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Bounded operators and von Neumann algebras

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ABSTRACT. We discuss the algebras of bounded operators $A \subset B(H)$, in the case where A is weakly closed, and has a trace $tr : A \rightarrow \mathbb{C}$. In this case we have $A = L^\infty(X)$, with X being a quantum space, and the free case, where $Z(A) = \mathbb{C}$, is of particular interest. We explain this material, following von Neumann and Connes, Jones, Voiculescu.

Preface

Quantum mechanics as we know it is the source of many puzzling questions. The simplest quantum mechanical system is the hydrogen atom, consisting of a negative charge, an electron, spinning around a positive charge, a proton. This reminds electrodynamics, and accepting the fact that the electron is a bit of a slippery particle, whose position and speed are described by probability, rather than by exact formulae, the hydrogen atom can indeed be solved, by starting with electrodynamics, and making a long series of “corrections”, for the most coming from various experiments, but sometimes coming as well from intuition, with the idea in mind that beautiful mathematics should correspond to true physics. The solution, as we presently know it, is something quite complicated.

Mathematically, the commonly accepted belief is that the good framework for the study of quantum mechanics is an infinite dimensional complex Hilbert space H , whose vectors can be thought of as being states of the system, and with the linear operators $T : H \rightarrow H$ corresponding to the observables. This is however to be taken with care, because in order to do “true physics”, things must be far sharper than that. Always remember indeed that the simplest object of quantum mechanics is the hydrogen atom, whose simplest state is the ground state, which is something quite complicated, and whose simplest observable is the first spectral line, which is something complicated too. Thus when talking about “states and observables”, we have a whole continuum of possible considerations and theories, ranging from true physics to very abstract mathematics.

For making things worse, even the existence and exact relevance of the Hilbert H is subject to debate. This is something more philosophical, with the 2-body problem evoked above, involving the hydrogen atom, being now replaced by a puzzling “1-body problem”, which has twisted the minds of many scientists, starting with Einstein and others. Can we get someday to a better quantum mechanics, by adding some kind of hidden variables to those available inside H ? No one really knows the answer here.

The present book is an introduction to the algebras $A \subset B(H)$ that the bounded linear operators $T : H \rightarrow H$ can form, once a Hilbert space H is given. There has been an enormous amount of work on such algebras, starting with von Neumann and others, and we will insist here on the aspects which are beautiful. With the idea, or rather hope in mind, that beautiful mathematics should correspond to true physics.

So, what is beauty, in the operator algebra framework? In our opinion, the source of all possible beauty is an old result of von Neumann, related to the Spectral Theorem for normal operators, which states that any commutative von Neumann algebra $A \subset B(H)$ must be of the form $A = L^\infty(X)$, with X being a measured space.

This is something subtle and interesting, which suggests doing several things with the arbitrary von Neumann algebras $A \subset B(H)$. Given such an algebra we can write the center as $Z(A) = L^\infty(X)$, we have then a decomposition of type $A = \int_X A_x dx$, and the problem is that of understanding the structure of the fibers, called “factors”. This is what von Neumann himself, and then Connes and others, did. Another idea, more speculative, following later work of Connes, and in parallel work of Voiculescu, is that of writing $A = L^\infty(X)$, with X being an abstract “quantum measured space”, and then trying to understand the geometry and probabilistic theory of X . Finally, yet another beautiful idea, due this time to Jones, is that of looking at the inclusions $A_0 \subset A_1$ of von Neumann algebras, instead at the von Neumann algebras themselves, the point being that the “symmetries” of such an inclusion lead to very interesting combinatorics.

All in all, many interesting things that can be done with a von Neumann algebra $A \subset B(H)$, and explaining the basics of the theory, plus having a look at the above 4 directions of research, is already what a graduate textbook, covering what can be taught during a 1-year graduate course, can cover. And this book is written exactly with this idea in mind. We will talk about all the above, keeping things as simple as possible, and with everything being accessible with a minimal knowledge of basic linear algebra, basic measure theory, basic functional analysis, and basic probability theory.

The book is organized in 4 parts, with Part I explaining the basics of operator theory, Part II explaining the basics of operator algebra theory, with a look into geometry and probability too, then Part III going into the structure of the von Neumann factors, and finally Part IV being an introduction to the subfactor theory of Jones.

There will be no physics in this book, which is meant to be a purely mathematical introduction to the operator algebras. Written of course with some physics in mind. We will give however references to more advanced books, some dealing with physics.

This book contains, besides the basics of the theory, a few recent contributions as well, in relation with the quantum group aspects. I am grateful to Julien Bichon, Benoît Collins, Steve Curran, Roland Speicher and my other coauthors, for working all this out. Many thanks go as well to my cats. Their views and opinions on mathematics, and knowledge of advanced functional analysis, have always been of great help.

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Part I

Bounded operators

*Does anybody here remember Vera Lynn
Remember how she said that
We would meet again
Some sunny day*

CHAPTER 1

Linear algebra

1a. Linear maps

We are interested in the bounded linear operators $T : H \rightarrow H$ from a Hilbert space to itself, and in the algebras $\langle T_i \rangle \subset B(H)$ that these operators can form. In this chapter we discuss the finite dimensional case, where $H = \mathbb{C}^N$, with usual scalar product:

$$\langle x, y \rangle = \sum_i x_i \bar{y}_i$$

In this case the boundedness condition is automatic, and our objects of study are the usual matrices $A \in M_N(\mathbb{C})$, due to the following fundamental result:

THEOREM 1.1. *The linear maps $T : \mathbb{C}^N \rightarrow \mathbb{C}^N$ are in correspondence with the square matrices $A \in M_N(\mathbb{C})$, with the linear map associated to such a matrix being*

$$Tx = Ax$$

and with the matrix associated to a linear map being:

$$A_{ij} = \langle Te_j, e_i \rangle$$

Moreover, such linear maps $T : \mathbb{C}^N \rightarrow \mathbb{C}^N$ are automatically bounded, in the sense that

$$\|T\| = \sup_{\|x\|=1} \|Tx\|$$

is finite.

PROOF. The first assertion is clear, because a linear map $T : \mathbb{C}^N \rightarrow \mathbb{C}^N$ must send a vector $x \in \mathbb{C}^N$ to a certain vector $Tx \in \mathbb{C}^N$, all whose components are linear combinations of the components of x . Thus, we can write, for certain complex numbers $A_{ij} \in \mathbb{C}$:

$$T \begin{pmatrix} x_1 \\ \vdots \\ x_N \end{pmatrix} = \begin{pmatrix} A_{11}x_1 + \dots + A_{1N}x_N \\ \vdots \\ A_{N1}x_1 + \dots + A_{NN}x_N \end{pmatrix}$$

Now the parameters $A_{ij} \in \mathbb{C}$ can be regarded as being the entries of a square matrix $A \in M_N(\mathbb{C})$, and with the usual convention for the rectangular matrix multiplication, the

above formula is precisely the one in the statement, namely:

$$Tx = Ax$$

Regarding the second assertion, if we denote by e_1, \dots, e_N the standard basis of \mathbb{C}^N , then we have the following formula, which gives $\langle Te_j, e_i \rangle = A_{ij}$:

$$Te_j = \begin{pmatrix} A_{1j} \\ \vdots \\ \vdots \\ A_{Nj} \end{pmatrix}$$

Regarding now the last assertion, let us estimate the norm of T , given by:

$$\|T\| = \sup_{\|x\|=1} \|Tx\|$$

Assuming $\|x\| = 1$ we have, by using the Cauchy-Schwarz inequality:

$$\begin{aligned} \|Tx\| &= \sqrt{\sum_{i=1}^N |A_{i1}x_1 + \dots + A_{iN}x_N|^2} \\ &\leq \sqrt{\sum_{i=1}^N |A_{i1}|^2 + \dots + |A_{iN}|^2} \\ &= \sqrt{\sum_{ij} |A_{ij}|^2} \end{aligned}$$

Thus the norm $\|T\|$ is bounded by the constant found above. \square

Our claim now is that, no matter what we want to do with T or A , of advanced type, we will run at some point into their adjoints T^* and A^* , constructed as follows:

PROPOSITION 1.2. *The adjoint operator $T^* : \mathbb{C}^N \rightarrow \mathbb{C}^N$, which is given by*

$$\langle Tx, y \rangle = \langle x, T^*y \rangle$$

corresponds to the adjoint matrix $A^ \in M_N(\mathbb{C})$, given by*

$$(A^*)_{ij} = \bar{A}_{ji}$$

via the correspondence between linear maps and matrices constructed above.

PROOF. Given a linear map $T : \mathbb{C}^N \rightarrow \mathbb{C}^N$, fix $y \in \mathbb{C}^N$, and consider the linear form $\varphi(x) = \langle Tx, y \rangle$. This form must be as follows, for a certain vector $T^*y \in \mathbb{C}^N$:

$$\varphi(x) = \langle x, T^*y \rangle$$

Thus, we have constructed a map $y \rightarrow T^*y$ as in the statement, which is obviously linear, and that we can call T^* . Now by taking the vectors $x, y \in \mathbb{C}^N$ to be elements of the standard basis of \mathbb{C}^N , our defining formula for T^* reads:

$$\langle Te_i, e_j \rangle = \langle e_i, T^*e_j \rangle$$

By reversing the scalar product on the right, this formula can be written as:

$$\langle T^*e_j, e_i \rangle = \overline{\langle Te_i, e_j \rangle}$$

But this means that the matrix associated to T^* is given by:

$$(A^*)_{ij} = \bar{A}_{ji}$$

Thus, we obtain indeed the adjoint matrix, as claimed. \square

Getting back now to our claim, when getting into advanced linear algebra, the adjoint linear maps and matrices are indeed ubiquitous. As an illustration here, we have:

THEOREM 1.3. *Consider the usual scalar product on \mathbb{C}^N , namely:*

$$\langle x, y \rangle = \sum_i x_i \bar{y}_i$$

(1) *A linear map $T : \mathbb{C}^N \rightarrow \mathbb{C}^N$, written as $T(x) = Ux$ with $U \in M_N(\mathbb{C})$, is an isometry precisely when U is unitary, in the sense that:*

$$U^* = U^{-1}$$

(2) *A linear map $T : \mathbb{C}^N \rightarrow \mathbb{C}^N$, written as $T(x) = Px$ with $P \in M_N(\mathbb{C})$, is a projection precisely when P is projection, in the sense that:*

$$P = P^2 = P^*$$

PROOF. We use Proposition 1.2, and more specifically the following formula coming from there, valid for any matrix $A \in M_N(\mathbb{C})$ and any two vectors $x, y \in \mathbb{C}^N$:

$$\langle Ax, y \rangle = \langle x, A^*y \rangle$$

We will need as well the complex polarization identity, whose proof is elementary, which allows us to recover scalar products out of distances, as follows:

$$\begin{aligned} 4 \langle x, y \rangle &= \|x + y\|^2 - \|x - y\|^2 \\ &+ i\|x + iy\|^2 - i\|x - iy\|^2 \end{aligned}$$

(1) Given a matrix $U \in M_N(\mathbb{C})$, we have indeed the following equivalences, with the first one coming from the polarization identity, and the other ones being clear:

$$\begin{aligned} \|Ux\| = \|x\| &\iff \langle Ux, Uy \rangle = \langle x, y \rangle \\ &\iff \langle x, U^*Uy \rangle = \langle x, y \rangle \\ &\iff U^*Uy = y \\ &\iff U^*U = 1 \\ &\iff U^* = U^{-1} \end{aligned}$$

(2) Given a matrix $P \in M_N(\mathbb{C})$, in order for $x \rightarrow Px$ to be a projection, we must have $P^2 = P$. Now observe that this projection is orthogonal precisely when:

$$\begin{aligned} \langle Px - x, Py \rangle = 0 &\iff \langle P^*Px - P^*x, y \rangle = 0 \\ &\iff P^*Px - P^*x = 0 \\ &\iff P^*P - P^* = 0 \\ &\iff P^*P = P^* \end{aligned}$$

Now observe that by conjugating the last formula, we obtain $P^*P = P^*$. Thus we must have $P = P^*$, and this gives the result. \square

Summarizing, the linear operators come in pairs T, T^* , and the associated matrices come as well in pairs A, A^* . We will keep this in mind, and come back to it later.

1b. Diagonalization

Let us discuss now the diagonalization question for linear maps and matrices. The basic theory here, formulated in a standard way, in terms of matrices, is as follows:

PROPOSITION 1.4. *A vector $v \in \mathbb{C}^N$ is called eigenvector of $A \in M_N(\mathbb{C})$, with corresponding eigenvalue λ , when A multiplies by λ in the direction of v :*

$$Av = \lambda v$$

In the case where \mathbb{C}^N has a basis v_1, \dots, v_N formed by eigenvectors of A , with corresponding eigenvalues $\lambda_1, \dots, \lambda_N$, in this new basis A becomes diagonal, as follows:

$$A \sim \begin{pmatrix} \lambda_1 & & \\ & \ddots & \\ & & \lambda_N \end{pmatrix}$$

Equivalently, if we denote by $D = \text{diag}(\lambda_1, \dots, \lambda_N)$ the above diagonal matrix, and by $P = [v_1 \dots v_N]$ the square matrix formed by the eigenvectors of A , we have:

$$A = PDP^{-1}$$

In this case we say that the matrix A is diagonalizable.

PROOF. This is something which is clear, the idea being as follows:

(1) The first assertion is clear, because the matrix which multiplies each basis element v_i by a number λ_i is precisely the diagonal matrix $D = \text{diag}(\lambda_1, \dots, \lambda_N)$.

(2) The second assertion follows from the first one, by changing the basis. We can prove this by a direct computation as well, because we have $Pe_i = v_i$, and so:

$$\begin{aligned} PDP^{-1}v_i &= PDe_i \\ &= P\lambda_i e_i \\ &= \lambda_i Pe_i \\ &= \lambda_i v_i \end{aligned}$$

Thus, the matrices A and PDP^{-1} coincide, as stated. \square

Let us recall as well that the basic example of a non diagonalizable matrix, over the complex numbers as above, is the following matrix:

$$A = \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix}$$

Indeed, the eigenvectors here are the vectors of type $\begin{pmatrix} x \\ 0 \end{pmatrix}$, all with eigenvalue 0. Thus, we have not enough eigenvectors for constructing a basis of \mathbb{C}^2 , out of them.

In general, in order to study the diagonalization problem, the idea is that the eigenvectors can be grouped into linear spaces, called eigenspaces, as follows:

THEOREM 1.5. *Let $A \in M_N(\mathbb{C})$, and for any eigenvalue $\lambda \in \mathbb{C}$ define the corresponding eigenspace as being the vector space formed by the corresponding eigenvectors:*

$$E_\lambda = \left\{ v \in \mathbb{C}^N \mid Av = \lambda v \right\}$$

These eigenspaces E_λ are then in a direct sum position, in the sense that given vectors $v_1 \in E_{\lambda_1}, \dots, v_k \in E_{\lambda_k}$ corresponding to different eigenvalues $\lambda_1, \dots, \lambda_k$, we have:

$$\sum_i c_i v_i = 0 \implies c_i = 0$$

In particular, we have $\sum_\lambda \dim(E_\lambda) \leq N$, with the sum being over all the eigenvalues, and our matrix is diagonalizable precisely when we have equality.

PROOF. We prove the first assertion by recurrence on $k \in \mathbb{N}$. Assume by contradiction that we have a formula as follows, with the scalars c_1, \dots, c_k being not all zero:

$$c_1 v_1 + \dots + c_k v_k = 0$$

By dividing by one of these scalars, we can assume that our formula is:

$$v_k = c_1 v_1 + \dots + c_{k-1} v_{k-1}$$

Now let us apply A to this vector. On the left we obtain:

$$\begin{aligned} Av_k &= \lambda_k v_k \\ &= \lambda_k c_1 v_1 + \dots + \lambda_k c_{k-1} v_{k-1} \end{aligned}$$

On the right we obtain something different, as follows:

$$\begin{aligned} A(c_1 v_1 + \dots + c_{k-1} v_{k-1}) &= c_1 A v_1 + \dots + c_{k-1} A v_{k-1} \\ &= c_1 \lambda_1 v_1 + \dots + c_{k-1} \lambda_{k-1} v_{k-1} \end{aligned}$$

We conclude from this that the following equality must hold:

$$\lambda_k c_1 v_1 + \dots + \lambda_k c_{k-1} v_{k-1} = c_1 \lambda_1 v_1 + \dots + c_{k-1} \lambda_{k-1} v_{k-1}$$

On the other hand, we know by recurrence that the vectors v_1, \dots, v_{k-1} must be linearly independent. Thus, the coefficients must be equal, at right and at left:

$$\lambda_k c_1 = c_1 \lambda_1$$

$$\vdots$$

$$\lambda_k c_{k-1} = c_{k-1} \lambda_{k-1}$$

Now since at least one c_i must be nonzero, from $\lambda_k c_i = c_i \lambda_i$ we obtain $\lambda_k = \lambda_i$, which is a contradiction. Thus our proof by recurrence of the first assertion is complete.

As for the second assertion, this follows from the first one. \square

In order to reach now to more advanced results, we can use the characteristic polynomial, which appears via the following fundamental result:

THEOREM 1.6. *Given a matrix $A \in M_N(\mathbb{C})$, consider its characteristic polynomial:*

$$P(x) = \det(A - x1_N)$$

The eigenvalues of A are then the roots of P . Also, we have the inequality

$$\dim(E_\lambda) \leq m_\lambda$$

where m_λ is the multiplicity of λ , as root of P .

PROOF. The first assertion follows from the following computation, using the fact that a linear map is bijective when the determinant of the associated matrix is nonzero:

$$\begin{aligned} \exists v, Av = \lambda v &\iff \exists v, (A - \lambda 1_N)v = 0 \\ &\iff \det(A - \lambda 1_N) = 0 \end{aligned}$$

Regarding now the second assertion, given an eigenvalue λ of our matrix A , consider the dimension $d_\lambda = \dim(E_\lambda)$ of the corresponding eigenspace. By changing the basis of \mathbb{C}^N , as for the eigenspace E_λ to be spanned by the first d_λ basis elements, our matrix becomes as follows, with B being a certain smaller matrix:

$$A \sim \begin{pmatrix} \lambda 1_{d_\lambda} & 0 \\ 0 & B \end{pmatrix}$$

We conclude that the characteristic polynomial of A is of the following form:

$$\begin{aligned} P_A &= P_{\lambda 1_{d_\lambda}} P_B \\ &= (\lambda - x)^{d_\lambda} P_B \end{aligned}$$

Thus the multiplicity m_λ of our eigenvalue λ , as a root of P , satisfies $m_\lambda \geq d_\lambda$, and this leads to the conclusion in the statement. \square

Now recall that we are over \mathbb{C} , which is something that we have not used yet, in our last two statements. And the point here is that we have the following key result:

THEOREM 1.7. *Any polynomial $P \in \mathbb{C}[X]$ decomposes as*

$$P = c(X - a_1) \dots (X - a_N)$$

with $c \in \mathbb{C}$ and with $a_1, \dots, a_N \in \mathbb{C}$.

PROOF. It is enough to prove that P has one root, and we prove this by contradiction. So, assume that P has no roots, and pick a number $z \in \mathbb{C}$ where $|P|$ attains its minimum:

$$|P(z)| = \min_{x \in \mathbb{C}} |P(x)| > 0$$

Since $Q(t) = P(z+t) - P(z)$ is a polynomial which vanishes at $t = 0$, this polynomial must be of the form $ct^k + \text{higher terms}$, with $c \neq 0$, and with $k \geq 1$ being an integer. We obtain from this that, with $t \in \mathbb{C}$ small, we have the following estimate:

$$P(z+t) \simeq P(z) + ct^k$$

Now let us write $t = rw$, with $r > 0$ small, and with $|w| = 1$. Our estimate becomes:

$$P(z+rw) \simeq P(z) + cr^k w^k$$

Now recall that we have assumed $P(z) \neq 0$. We can therefore choose $w \in \mathbb{T}$ such that cw^k points in the opposite direction to that of $P(z)$, and we obtain in this way:

$$\begin{aligned} |P(z+rw)| &\simeq |P(z) + cr^k w^k| \\ &= |P(z)|(1 - |c|r^k) \end{aligned}$$

Now by choosing $r > 0$ small enough, as for the error in the first estimate to be small, and overcome by the negative quantity $-|c|r^k$, we obtain from this:

$$|P(z+rw)| < |P(z)|$$

But this contradicts our definition of $z \in \mathbb{C}$, as a point where $|P|$ attains its minimum. Thus P has a root, and by recurrence it has N roots, as stated. \square

Now by putting everything together, we obtain the following result:

THEOREM 1.8. *Given a matrix $A \in M_N(\mathbb{C})$, consider its characteristic polynomial*

$$P(X) = \det(A - X1_N)$$

then factorize this polynomial, by computing the complex roots, with multiplicities,

$$P(X) = (-1)^N (X - \lambda_1)^{n_1} \dots (X - \lambda_k)^{n_k}$$

and finally compute the corresponding eigenspaces, for each eigenvalue found:

$$E_i = \left\{ v \in \mathbb{C}^N \mid Av = \lambda_i v \right\}$$

The dimensions of these eigenspaces satisfy then the following inequalities,

$$\dim(E_i) \leq n_i$$

and A is diagonalizable precisely when we have equality for any i .

PROOF. The statement is well formulated, thanks to Theorem 1.7. By summing the inequalities $\dim(E_\lambda) \leq m_\lambda$ from Theorem 1.6, we obtain an inequality as follows:

$$\sum_{\lambda} \dim(E_\lambda) \leq \sum_{\lambda} m_\lambda \leq N$$

On the other hand, we know from Theorem 1.5 that our matrix is diagonalizable when we have global equality. Thus, we are led to the conclusion in the statement. \square

This was for the main result of linear algebra. There are countless applications of this, and generally speaking, advanced linear algebra consists in building on Theorem 1.8.

Let us record as well a useful algorithmic version of the above result:

THEOREM 1.9. *The square matrices $A \in M_N(\mathbb{C})$ can be diagonalized as follows:*

- (1) *Compute the characteristic polynomial.*
- (2) *Factorize the characteristic polynomial.*
- (3) *Compute the eigenvectors, for each eigenvalue found.*
- (4) *If there are no N eigenvectors, A is not diagonalizable.*
- (5) *Otherwise, A is diagonalizable, $A = PDP^{-1}$.*

PROOF. This is an informal reformulation of Theorem 1.8 above, with (4) referring to the total number of linearly independent eigenvectors found in (3), and with $A = PDP^{-1}$ in (5) being the usual diagonalization formula, with P, D being as before. \square

As a remark here, in step (3) it is always better to start with the eigenvalues having big multiplicity. Indeed, a multiplicity 1 eigenvalue, for instance, can never lead to the end of the computation, via (4), simply because the eigenvectors always exist.

1c. Matrix tricks

At the level of basic examples of diagonalizable matrices, we first have the following result, which provides us with the “generic” examples:

THEOREM 1.10. *For a matrix $A \in M_N(\mathbb{C})$ the following conditions are equivalent,*

- (1) *The eigenvalues are different, $\lambda_i \neq \lambda_j$,*
- (2) *The characteristic polynomial P has simple roots,*
- (3) *The characteristic polynomial satisfies $(P, P') = 1$,*
- (4) *The resultant of P, P' is nonzero, $R(P, P') \neq 0$,*
- (5) *The discriminant of P is nonzero, $\Delta(P) \neq 0$,*

and in this case, the matrix is diagonalizable.

PROOF. The last assertion holds indeed, due to Theorem 1.8. As for the equivalences in the statement, these are all standard, the idea for their proofs, along with some more theory, needed for using in practice the present result, being as follows:

- (1) \iff (2) This follows from Theorem 1.8.
- (2) \iff (3) This is standard, the double roots of P being roots of P' .
- (3) \iff (4) The idea here is that associated to any two polynomials P, Q is their resultant $R(P, Q)$, which checks whether P, Q have a common root. Let us write:

$$P = c(X - a_1) \dots (X - a_k)$$

$$Q = d(X - b_1) \dots (X - b_l)$$

We can define then the resultant as being the following quantity:

$$R(P, Q) = c^l d^k \prod_{ij} (a_i - b_j)$$

The point now, that we will explain as well, is that this is a polynomial in the coefficients of P, Q , with integer coefficients. Indeed, this can be checked as follows:

– We can expand the formula of $R(P, Q)$, and in what regards a_1, \dots, a_k , which are the roots of P , we obtain in this way certain symmetric functions in these variables, which will be therefore polynomials in the coefficients of P , with integer coefficients.

– We can then look what happens with respect to the remaining variables b_1, \dots, b_l , which are the roots of Q . Once again what we have here are certain symmetric functions, and so polynomials in the coefficients of Q , with integer coefficients.

– Thus, we are led to the above conclusion, that $R(P, Q)$ is a polynomial in the coefficients of P, Q , with integer coefficients, and with the remark that the $c^l d^k$ factor is there for these latter coefficients to be indeed integers, instead of rationals.

Alternatively, let us write our two polynomials in usual form, as follows:

$$P = p_k X^k + \dots + p_1 X + p_0$$

$$Q = q_l X^l + \dots + q_1 X + q_0$$

The corresponding resultant appears then as the determinant of an associated matrix, having size $k + l$, and having 0 coefficients at the blank spaces, as follows:

$$R(P, Q) = \begin{vmatrix} p_k & & & q_l & & \\ \vdots & \ddots & & \vdots & \ddots & \\ p_0 & & p_k & q_0 & & q_k \\ & & \ddots & \vdots & & \ddots \\ & & & p_0 & & q_0 \end{vmatrix}$$

Indeed, this follows by doing some linear algebra computations, mixed with algebra, in the spirit of those from the proof of the Vandermonde determinant theorem.

(4) \iff (5) Once again this is something standard, the idea here being that the discriminant $\Delta(P)$ of a polynomial $P \in \mathbb{C}[X]$ is, modulo scalars, the resultant $R(P, P')$. To be more precise, let us write our polynomial as follows:

$$P(X) = cX^N + dX^{N-1} + \dots$$

Its discriminant is then defined as being the following quantity:

$$\Delta(P) = \frac{(-1)^{\binom{N}{2}}}{c} R(P, P')$$

This is a polynomial in the coefficients of P , with integer coefficients, with the division by c being indeed possible, under \mathbb{Z} , and with the sign being there for various reasons, including the compatibility with some well-known formulae, at small values of N . \square

All this might seem a bit complicated, so as an illustration, let us work out an example. Consider the case of a polynomial of degree 2, and a polynomial of degree 1:

$$P = ax^2 + bx + c$$

$$Q = dx + e$$

In order to compute the resultant, let us factorize our polynomials:

$$P = a(x - p)(x - q)$$

$$Q = d(x - r)$$

The resultant can be then computed as follows, by using the two-step method:

$$\begin{aligned}
 R(P, Q) &= ad^2(p-r)(q-r) \\
 &= ad^2(pq - (p+q)r + r^2) \\
 &= cd^2 + bd^2r + ad^2r^2 \\
 &= cd^2 - bde + ae^2
 \end{aligned}$$

Observe that $R(P, Q) = 0$ corresponds indeed to the fact that P, Q have a common root. Indeed, the root of Q is $r = -e/d$, and we have:

$$\begin{aligned}
 P(r) &= \frac{ae^2}{d^2} - \frac{be}{d} + c \\
 &= \frac{R(P, Q)}{d^2}
 \end{aligned}$$

We can recover as well the resultant as a determinant, as follows:

$$\begin{aligned}
 R(P, Q) &= \begin{vmatrix} a & d & 0 \\ b & e & d \\ c & 0 & e \end{vmatrix} \\
 &= ae^2 - bde + cd^2
 \end{aligned}$$

Finally, in what regards the discriminant, let us see what happens in degree 2. Here we must compute the resultant of the following two polynomials:

$$\begin{aligned}
 P &= aX^2 + bX + c \\
 P' &= 2aX + b
 \end{aligned}$$

The resultant is then given by the following formula:

$$\begin{aligned}
 R(P, P') &= ab^2 - b(2a)b + c(2a)^2 \\
 &= 4a^2c - ab^2 \\
 &= -a(b^2 - 4ac)
 \end{aligned}$$

Now by doing the discriminant normalizations, we obtain, as we should:

$$\Delta(P) = b^2 - 4ac$$

There are many other things that can be said about resultants and discriminants, and all this can be found in any good matrix analysis book.

As already mentioned before, one can prove that the matrices having distinct eigenvalues are “generic”, and so the above result basically captures the whole situation.

We have in fact the following collection of density results, which are quite advanced, and with all being very useful statements, in practice, coming as a complement to what has already been explained in the above:

THEOREM 1.11. *The following happen, inside $M_N(\mathbb{C})$:*

- (1) *The invertible matrices are dense.*
- (2) *The matrices having distinct eigenvalues are dense.*
- (3) *The diagonalizable matrices are dense.*

PROOF. These are quite advanced linear algebra results, which can be proved as follows, with the technology that we have so far:

(1) This is clear, intuitively speaking, because the invertible matrices are given by the condition $\det A \neq 0$. Thus, the set formed by these matrices appears as the complement of the surface $\det A = 0$, and so must be dense inside $M_N(\mathbb{C})$, as claimed.

(2) Here we can use a similar argument, this time by saying that the set formed by the matrices having distinct eigenvalues appears as the complement of the surface given by $\Delta(P_A) = 0$, and so must be dense inside $M_N(\mathbb{C})$, as claimed.

(3) This follows from (2), via the fact that the matrices having distinct eigenvalues are diagonalizable, that we know from Theorem 1.10 above. There are of course some other proofs as well, for instance by putting the matrix in Jordan form. \square

As an application of the above results, and of our methods in general, we have:

THEOREM 1.12. *The following happen:*

- (1) *We have $P_{AB} = P_{BA}$, for any two matrices $A, B \in M_N(\mathbb{C})$.*
- (2) *AB, BA have the same eigenvalues, with the same multiplicities.*
- (3) *If A has eigenvalues $\lambda_1, \dots, \lambda_N$, then $f(A)$ has eigenvalues $f(\lambda_1), \dots, f(\lambda_N)$.*

PROOF. These results can be deduced by using Theorem 1.11, as follows:

(1) It follows from definitions that the characteristic polynomial of a matrix is invariant under conjugation, in the sense that we have:

$$P_C = P_{ACA^{-1}}$$

Now observe that, when assuming that A is invertible, we have:

$$AB = A(BA)A^{-1}$$

Thus, we have the result when A is invertible. By using now Theorem 1.11 (1) above, we conclude that this formula holds for any matrix A , by continuity.

(2) This is a reformulation of (1) above, via the fact that P encodes the eigenvalues, with multiplicities, which is hard to prove with bare hands.

(3) This is something more informal, the idea being that this is clear for the diagonal matrices D , then for the diagonalizable matrices PDP^{-1} , and finally for all the matrices, by using Theorem 1.11 (3), provided that f has suitable regularity properties. \square

We will be back to all this later on, and especially on (3) above, which is something quite subtle, directly in the case of linear operators on arbitrary Hilbert spaces, finite dimensional or not, when systematically doing operator theory.

Let us go back to the main problem raised by the diagonalization procedure, namely the computation of the roots of characteristic polynomials. We have here:

THEOREM 1.13. *The complex eigenvalues of a matrix $A \in M_N(\mathbb{C})$, counted with multiplicities, have the following properties:*

- (1) *Their sum is the trace.*
- (2) *Their product is the determinant.*

PROOF. Consider indeed the characteristic polynomial P of the matrix:

$$\begin{aligned} P(X) &= \det(A - X1_N) \\ &= (-1)^N X^N + (-1)^{N-1} \text{Tr}(A) X^{N-1} + \dots + \det(A) \end{aligned}$$

We can factorize this polynomial, by using its N complex roots, and we obtain:

$$\begin{aligned} P(X) &= (-1)^N (X - \lambda_1) \dots (X - \lambda_N) \\ &= (-1)^N X^N + (-1)^{N-1} \left(\sum_i \lambda_i \right) X^{N-1} + \dots + \prod_i \lambda_i \end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

Regarding now the intermediate terms, we have here:

THEOREM 1.14. *Assume that $A \in M_N(\mathbb{C})$ has eigenvalues $\lambda_1, \dots, \lambda_N \in \mathbb{C}$, counted with multiplicities. The basic symmetric functions of these eigenvalues, namely*

$$c_k = \sum_{i_1 < \dots < i_k} \lambda_{i_1} \dots \lambda_{i_k}$$

are then given by the fact that the characteristic polynomial of the matrix is:

$$P(X) = (-1)^N \sum_{k=0}^N (-1)^k c_k X^k$$

Moreover, all symmetric functions of the eigenvalues, such as the sums of powers

$$d_s = \lambda_1^s + \dots + \lambda_N^s$$

appear as polynomials in these characteristic polynomial coefficients c_k .

PROOF. These results can be proved by doing some algebra, as follows:

(1) Consider indeed the characteristic polynomial P of the matrix, factorized by using its N complex roots, taken with multiplicities. By expanding, we obtain:

$$\begin{aligned} P(X) &= (-1)^N (X - \lambda_1) \dots (X - \lambda_N) \\ &= (-1)^N X^N + (-1)^{N-1} \left(\sum_i \lambda_i \right) X^{N-1} + \dots + \prod_i \lambda_i \\ &= (-1)^N X^N + (-1)^{N-1} c_1 X^{N-1} + \dots + (-1)^0 c_N \\ &= (-1)^N (X^N - c_1 X^{N-1} + \dots + (-1)^N c_N) \end{aligned}$$

With the convention $c_0 = 1$, we are led to the conclusion in the statement.

(2) This is something standard, coming by doing some abstract algebra. Working out the formulae for the sums of powers $d_s = \sum_i \lambda_i^s$, at small values of the exponent $s \in \mathbb{N}$, is an excellent exercise, which shows how to proceed in general, by recurrence. \square

1d. Spectral theorems

Let us go back now to the diagonalization question. Here is a key result:

THEOREM 1.15. *Any matrix $A \in M_N(\mathbb{C})$ which is self-adjoint, $A = A^*$, is diagonalizable, with the diagonalization being of the following type,*

$$A = UDU^*$$

with $U \in U_N$, and with $D \in M_N(\mathbb{R})$ diagonal. The converse holds too.

PROOF. As a first remark, the converse trivially holds, because if we take a matrix of the form $A = UDU^*$, with U unitary and D diagonal and real, then we have:

$$\begin{aligned} A^* &= (UDU^*)^* \\ &= UD^*U^* \\ &= UDU^* \\ &= A \end{aligned}$$

In the other sense now, assume that A is self-adjoint, $A = A^*$. Our first claim is that the eigenvalues are real. Indeed, assuming $Av = \lambda v$, we have:

$$\begin{aligned} \lambda \langle v, v \rangle &= \langle \lambda v, v \rangle \\ &= \langle Av, v \rangle \\ &= \langle v, Av \rangle \\ &= \langle v, \lambda v \rangle \\ &= \bar{\lambda} \langle v, v \rangle \end{aligned}$$

Thus we obtain $\lambda \in \mathbb{R}$, as claimed. Our next claim now is that the eigenspaces corresponding to different eigenvalues are pairwise orthogonal. Assume indeed that:

$$Av = \lambda v$$

$$Aw = \mu w$$

We have then the following computation, using $\lambda, \mu \in \mathbb{R}$:

$$\begin{aligned} \lambda \langle v, w \rangle &= \langle \lambda v, w \rangle \\ &= \langle Av, w \rangle \\ &= \langle v, Aw \rangle \\ &= \langle v, \mu w \rangle \\ &= \mu \langle v, w \rangle \end{aligned}$$

Thus $\lambda \neq \mu$ implies $v \perp w$, as claimed. In order now to finish the proof, it remains to prove that the eigenspaces of A span the whole space \mathbb{C}^N . For this purpose, we will use a recurrence method. Let us pick an eigenvector of our matrix:

$$Av = \lambda v$$

Assuming now that we have a vector w orthogonal to it, $v \perp w$, we have:

$$\begin{aligned} \langle Aw, v \rangle &= \langle w, Av \rangle \\ &= \langle w, \lambda v \rangle \\ &= \lambda \langle w, v \rangle \\ &= 0 \end{aligned}$$

Thus, if v is an eigenvector, then the vector space v^\perp is invariant under A . In order to do now the recurrence, it still remains to prove that the restriction of A to the vector space v^\perp is self-adjoint. But this comes from a general property of the self-adjoint matrices, that we will explain now. Our claim is that an arbitrary square matrix A is self-adjoint precisely when the following happens, for any vector v :

$$\langle Av, v \rangle \in \mathbb{R}$$

Indeed, the fact that the above scalar product is real is equivalent to:

$$\langle (A - A^*)v, v \rangle = 0$$

But this is equivalent to $A = A^*$, by using the complex polarization identity. Now back to our questions, it is clear from our self-adjointness criterion above that the restriction of A to any invariant subspace, and in particular to the subspace v^\perp , is self-adjoint. Thus, we can proceed by recurrence, and we obtain the result. \square

As basic examples of self-adjoint matrices, we have the orthogonal projections. The diagonalization result regarding them is as follows:

PROPOSITION 1.16. *The matrices $P \in M_N(\mathbb{C})$ which are projections,*

$$P^2 = P = P^*$$

are precisely those which diagonalize as follows,

$$P = UDU^*$$

with $U \in U_N$, and with $D \in M_N(0, 1)$ being diagonal.

PROOF. The equation for the projections being $P^2 = P = P^*$, the eigenvalues λ are real, and we have as well the following condition, coming from $P^2 = P$:

$$\begin{aligned} \lambda \langle v, v \rangle &= \langle \lambda v, v \rangle \\ &= \langle Pv, v \rangle \\ &= \langle P^2v, v \rangle \\ &= \langle Pv, Pv \rangle \\ &= \langle \lambda v, \lambda v \rangle \\ &= \lambda^2 \langle v, v \rangle \end{aligned}$$

Thus we obtain $\lambda \in \{0, 1\}$, as claimed, and as a final conclusion here, the diagonalization of the self-adjoint matrices is as follows, with $e_i \in \{0, 1\}$:

$$P \sim \begin{pmatrix} e_1 & & \\ & \ddots & \\ & & e_N \end{pmatrix}$$

To be more precise, the number of 1 values is the dimension of the image of P , and the number of 0 values is the dimension of space of vectors sent to 0 by P . \square

An important class of self-adjoint matrices, which includes for instance all the projections, are the positive matrices. The theory here is as follows:

THEOREM 1.17. *For a matrix $A \in M_N(\mathbb{C})$ the following conditions are equivalent, and if they are satisfied, we say that A is positive:*

- (1) $A = B^2$, with $B = B^*$.
- (2) $A = CC^*$, for some $C \in M_N(\mathbb{C})$.
- (3) $\langle Ax, x \rangle \geq 0$, for any vector $x \in \mathbb{C}^N$.
- (4) $A = A^*$, and the eigenvalues are positive, $\lambda_i \geq 0$.
- (5) $A = UDU^*$, with $U \in U_N$ and with $D \in M_N(\mathbb{R}_+)$ diagonal.

PROOF. The idea is that the equivalences in the statement basically follow from some elementary computations, with only Theorem 1.15 needed, at some point:

- (1) \implies (2) This is clear, because we can take $C = B$.

(2) \implies (3) This follows from the following computation:

$$\begin{aligned} \langle Ax, x \rangle &= \langle CC^*x, x \rangle \\ &= \langle C^*x, C^*x \rangle \\ &\geq 0 \end{aligned}$$

(3) \implies (4) By using the fact that $\langle Ax, x \rangle$ is real, we have:

$$\begin{aligned} \langle Ax, x \rangle &= \langle x, A^*x \rangle \\ &= \langle A^*x, x \rangle \end{aligned}$$

Thus we have $A = A^*$, and the remaining assertion, regarding the eigenvalues, follows from the following computation, assuming $Ax = \lambda x$:

$$\begin{aligned} \langle Ax, x \rangle &= \langle \lambda x, x \rangle \\ &= \lambda \langle x, x \rangle \\ &\geq 0 \end{aligned}$$

(4) \implies (5) This follows by using Theorem 1.15 above.

(5) \implies (1) Assuming $A = UDU^*$ is as in the statement, with $U \in U_N$, and with $D \in M_N(\mathbb{R}_+)$ being diagonal, we can set:

$$B = U\sqrt{D}U^*$$

Then B is self-adjoint, and its square is given by:

$$\begin{aligned} B^2 &= U\sqrt{D}U^* \cdot U\sqrt{D}U^* \\ &= UDU^* \\ &= A \end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

Let us record as well the following technical version of the above result:

THEOREM 1.18. *For a matrix $A \in M_N(\mathbb{C})$ the following conditions are equivalent, and if they are satisfied, we say that A is strictly positive:*

- (1) $A = B^2$, with $B = B^*$, invertible.
- (2) $A = CC^*$, for some $C \in M_N(\mathbb{C})$ invertible.
- (3) $\langle Ax, x \rangle > 0$, for any nonzero vector $x \in \mathbb{C}^N$.
- (4) $A = A^*$, and the eigenvalues are strictly positive, $\lambda_i > 0$.
- (5) $A = UDU^*$, with $U \in U_N$ and with $D \in M_N(\mathbb{R}_+^*)$ diagonal.

PROOF. This follows either from Theorem 1.17, by adding the various extra assumptions in the statement, or from the proof of Theorem 1.17, by modifying where needed. \square

We will be back to positive matrices later on, at the end of the present chapter.

Let us discuss now the case of the unitary matrices. We have here:

THEOREM 1.19. *Any matrix $U \in M_N(\mathbb{C})$ which is unitary, $U^* = U^{-1}$, is diagonalizable, with the eigenvalues on \mathbb{T} . More precisely we have*

$$U = VDV^*$$

with $V \in U_N$, and with $D \in M_N(\mathbb{T})$ diagonal. The converse holds too.

PROOF. As a first remark, the converse trivially holds, because given a matrix of type $U = VDV^*$, with $V \in U_N$, and with $D \in M_N(\mathbb{T})$ being diagonal, we have:

$$\begin{aligned} U^* &= (VDV^*)^* \\ &= VD^*V^* \\ &= VD^{-1}V^{-1} \\ &= (V^*)^{-1}D^{-1}V^{-1} \\ &= (VDV^*)^{-1} \\ &= U^{-1} \end{aligned}$$

Let us prove now the first assertion, stating that the eigenvalues of a unitary matrix $U \in U_N$ belong to \mathbb{T} . Indeed, assuming $Uv = \lambda v$, we have:

$$\begin{aligned} \langle v, v \rangle &= \langle U^*Uv, v \rangle \\ &= \langle Uv, Uv \rangle \\ &= \langle \lambda v, \lambda v \rangle \\ &= |\lambda|^2 \langle v, v \rangle \end{aligned}$$

Thus we obtain $\lambda \in \mathbb{T}$, as claimed. Our next claim now is that the eigenspaces corresponding to different eigenvalues are pairwise orthogonal. Assume indeed that:

$$Uv = \lambda v$$

$$Uw = \mu w$$

We have then the following computation, using $U^* = U^{-1}$ and $\lambda, \mu \in \mathbb{T}$:

$$\begin{aligned} \lambda \langle v, w \rangle &= \langle \lambda v, w \rangle \\ &= \langle Uv, w \rangle \\ &= \langle v, U^*w \rangle \\ &= \langle v, U^{-1}w \rangle \\ &= \langle v, \mu^{-1}w \rangle \\ &= \mu \langle v, w \rangle \end{aligned}$$

Thus $\lambda \neq \mu$ implies $v \perp w$, as claimed. In order now to finish the proof, it remains to prove that the eigenspaces of U span the whole space \mathbb{C}^N . For this purpose, we will use a recurrence method. Let us pick an eigenvector of our matrix:

$$Uv = \lambda v$$

Assuming that we have a vector w orthogonal to it, $v \perp w$, we have:

$$\begin{aligned} \langle Uw, v \rangle &= \langle w, U^*v \rangle \\ &= \langle w, U^{-1}v \rangle \\ &= \langle w, \lambda^{-1}v \rangle \\ &= \lambda \langle w, v \rangle \\ &= 0 \end{aligned}$$

Thus, if v is an eigenvector, then the vector space v^\perp is invariant under U . Now since U is an isometry, so is its restriction to this space v^\perp . Thus this restriction is a unitary, and so we can proceed by recurrence, and we obtain the result. \square

The self-adjoint matrices and the unitary matrices are particular cases of the general notion of a “normal matrix”, and we have here:

THEOREM 1.20. *Any matrix $A \in M_N(\mathbb{C})$ which is normal, $AA^* = A^*A$, is diagonalizable, with the diagonalization being of the following type,*

$$A = UDU^*$$

with $U \in U_N$, and with $D \in M_N(\mathbb{C})$ diagonal. The converse holds too.

PROOF. As a first remark, the converse trivially holds, because if we take a matrix of the form $A = UDU^*$, with U unitary and D diagonal, then we have:

$$\begin{aligned} AA^* &= UDU^* \cdot UD^*U^* \\ &= UDD^*U^* \\ &= UD^*DU^* \\ &= UD^*U^* \cdot UDU^* \\ &= A^*A \end{aligned}$$

In the other sense now, this is something more technical. Our first claim is that a matrix A is normal precisely when the following happens, for any vector v :

$$\|Av\| = \|A^*v\|$$

Indeed, the above equality can be written as follows:

$$\langle AA^*v, v \rangle = \langle A^*Av, v \rangle$$

But this is equivalent to $AA^* = A^*A$, by using the polarization identity. Our claim now is that A, A^* have the same eigenvectors, with conjugate eigenvalues:

$$Av = \lambda v \implies A^*v = \bar{\lambda}v$$

Indeed, this follows from the following computation, and from the trivial fact that if A is normal, then so is any matrix of type $A - \lambda 1_N$:

$$\begin{aligned} \|(A^* - \bar{\lambda}1_N)v\| &= \|(A - \lambda 1_N)^*v\| \\ &= \|(A - \lambda 1_N)v\| \\ &= 0 \end{aligned}$$

Let us prove now, by using this, that the eigenspaces of A are pairwise orthogonal. Assume that we have two eigenvectors, corresponding to different eigenvalues, $\lambda \neq \mu$:

$$Av = \lambda v \quad , \quad Aw = \mu w$$

We have the following computation, which shows that $\lambda \neq \mu$ implies $v \perp w$:

$$\begin{aligned} \lambda \langle v, w \rangle &= \langle \lambda v, w \rangle \\ &= \langle Av, w \rangle \\ &= \langle v, A^*w \rangle \\ &= \langle v, \bar{\mu}w \rangle \\ &= \mu \langle v, w \rangle \end{aligned}$$

In order to finish, it remains to prove that the eigenspaces of A span the whole \mathbb{C}^N . This is something that we have already seen for the self-adjoint matrices, and for unitaries, and we will use here these results, in order to deal with the general normal case. As a first observation, given an arbitrary matrix A , the matrix AA^* is self-adjoint:

$$(AA^*)^* = AA^*$$

Thus, we can diagonalize this matrix AA^* , as follows, with the passage matrix being a unitary, $V \in U_N$, and with the diagonal form being real, $E \in M_N(\mathbb{R})$:

$$AA^* = VEV^*$$

Now observe that, for matrices of type $A = UDU^*$, which are those that we supposed to deal with, we have the following formulae:

$$V = U \quad , \quad E = D\bar{D}$$

In particular, the matrices A and AA^* have the same eigenspaces. So, this will be our idea, proving that the eigenspaces of AA^* are eigenspaces of A . In order to do so, let us pick two eigenvectors v, w of the matrix AA^* , corresponding to different eigenvalues, $\lambda \neq \mu$. The eigenvalue equations are then as follows:

$$AA^*v = \lambda v \quad , \quad AA^*w = \mu w$$

We have the following computation, using the normality condition $AA^* = A^*A$, and the fact that the eigenvalues of AA^* , and in particular μ , are real:

$$\begin{aligned}
\lambda \langle Av, w \rangle &= \langle \lambda Av, w \rangle \\
&= \langle A\lambda v, w \rangle \\
&= \langle AAA^*v, w \rangle \\
&= \langle AA^*Av, w \rangle \\
&= \langle Av, AA^*w \rangle \\
&= \langle Av, \mu w \rangle \\
&= \mu \langle Av, w \rangle
\end{aligned}$$

We conclude that we have $\langle Av, w \rangle = 0$. But this reformulates as follows:

$$\lambda \neq \mu \implies A(E_\lambda) \perp E_\mu$$

Now since the eigenspaces of AA^* are pairwise orthogonal, and span the whole \mathbb{C}^N , we deduce from this that these eigenspaces are invariant under A :

$$A(E_\lambda) \subset E_\lambda$$

But with this result in hand, we can finish. Indeed, we can decompose the problem, and the matrix A itself, following these eigenspaces of AA^* , which in practice amounts in saying that we can assume that we only have 1 eigenspace. By rescaling, this is the same as assuming that we have $AA^* = 1$, and so we are now into the unitary case, that we know how to solve, as explained in Theorem 1.19 above. \square

As a first application, we have the following result:

THEOREM 1.21. *Given a matrix $A \in M_N(\mathbb{C})$, we can construct a matrix $|A|$ as follows, by using the fact that A^*A is diagonalizable, with positive eigenvalues:*

$$|A| = \sqrt{A^*A}$$

*This matrix $|A|$ is then positive, and its square is $|A|^2 = A^*A$. In the case $N = 1$, we obtain in this way the usual absolute value of the complex numbers.*

PROOF. Consider indeed the matrix A^*A , which is normal. According to Theorem 1.20, we can diagonalize this matrix as follows, with $U \in U_N$, and with D diagonal:

$$A^*A = UDU^*$$

From $A^*A \geq 0$ we obtain $D \geq 0$. But this means that the entries of D are real, and positive. Thus we can extract the square root \sqrt{D} , and then set:

$$\sqrt{A^*A} = U\sqrt{D}U^*$$

Thus, we are basically done. Indeed, if we call this latter matrix $|A|$, then we are led to the conclusions in the statement. Finally, the last assertion is clear from definitions. \square

We can now formulate a first polar decomposition result, as follows:

THEOREM 1.22. *Any invertible matrix $A \in M_N(\mathbb{C})$ decomposes as*

$$A = U|A|$$

with $U \in U_N$, and with $|A| = \sqrt{A^*A}$ as above.

PROOF. This is routine, and follows by comparing the actions of $A, |A|$ on the vectors $v \in \mathbb{C}^N$, and deducing from this the existence of a unitary $U \in U_N$ as above. \square

Observe that at $N = 1$ we obtain in this way the usual polar decomposition of the nonzero complex numbers. There are of course many other examples.

More generally now, we have the following result:

THEOREM 1.23. *Any square matrix $A \in M_N(\mathbb{C})$ decomposes as*

$$A = U|A|$$

with U being a partial isometry, and with $|A| = \sqrt{A^*A}$ as above.

PROOF. Once again, this follows by comparing the actions of $A, |A|$ on the vectors $v \in \mathbb{C}^N$, and deducing from this the existence of a partial isometry U as above. Alternatively, we can get this from Theorem 1.22, applied on the complement of the 0-eigenvectors. \square

1e. Exercises

Linear algebra is a very wide topic, and there are countless interesting matrices, and exercises about them. Here is one selected such exercise, which is a must-do:

EXERCISE 1.24. *Prove that the flat matrix, which is the all-one $N \times N$ matrix, diagonalizes over the complex numbers as follows,*

$$\begin{pmatrix} 1 & \cdots & 1 \\ \vdots & & \vdots \\ 1 & \cdots & 1 \end{pmatrix} = \frac{1}{N} F_N \begin{pmatrix} N & & & \\ & 0 & & \\ & & \cdots & \\ & & & 0 \end{pmatrix} F_N^*$$

with $F_N = (w^{ij})_{ij}$ with $w = e^{2\pi i/N}$ being the Fourier matrix.

This is something very instructive. Normally you have to look for eigenvectors for the flat matrix, and you are led in this way to the equation $x_0 + \dots + x_{N-1} = 0$. The problem however is that this equation, while looking very gentle, has no “canonical” solutions over the real numbers. Thus you are led to the complex numbers, and more specifically to the roots of unity, and their magic, leading to the above result. Enjoy.

CHAPTER 2

Bounded operators

2a. Hilbert spaces

We discuss in this chapter, and in the next two ones, an extension of the linear algebra results obtained in the previous chapter, in infinite or arbitrary dimensions.

There are many subtleties here, and to begin with, abstract linear algebra is not very interesting over arbitrary complex vector spaces, and this due to a number of technical reasons, having to do with existence problems for bases, and other complicated issues.

We must therefore use vector spaces with some extra structure. And the ideal such extra structure is a scalar product:

DEFINITION 2.1. *A scalar product on a complex vector space H is an operation*

$$H \times H \rightarrow \mathbb{C}$$

denoted $(x, y) \rightarrow \langle x, y \rangle$, satisfying the following conditions:

- (1) $\langle x, y \rangle$ is linear in x , and antilinear in y .
- (2) $\overline{\langle x, y \rangle} = \langle y, x \rangle$, for any x, y .
- (3) $\langle x, x \rangle \geq 0$, for any $x \neq 0$.

As a basic example here, we have the finite dimensional vector space $H = \mathbb{C}^N$, with its usual scalar product, which is as follows:

$$\langle x, y \rangle = \sum_i x_i \bar{y}_i$$

There are many other examples, and notably various spaces of L^2 functions, which naturally appear in problems coming from physics. We will discuss them later on.

In order to study the scalar products, let us formulate the following definition:

DEFINITION 2.2. *The norm of a vector $x \in H$ is the following quantity:*

$$\|x\| = \sqrt{\langle x, x \rangle}$$

We also call this number length of x , or distance from x to the origin.

The terminology comes from what happens in \mathbb{C}^N , where the length of the vector, as defined above, coincides with the usual length, given by:

$$\|x\| = \sqrt{\sum_i |x_i|^2}$$

In analogy with what happens in finite dimensions, we have two important results regarding the norms. First we have the Cauchy-Schwarz inequality, as follows:

THEOREM 2.3. *We have the Cauchy-Schwarz inequality*

$$|\langle x, y \rangle| \leq \|x\| \cdot \|y\|$$

and the equality case holds precisely when x, y are proportional.

PROOF. This is something very standard. Consider indeed the following quantity, depending on a real variable $t \in \mathbb{R}$, and on a variable on the unit circle, $w \in \mathbb{T}$:

$$f(t) = \|twx + y\|^2$$

By developing f , we see that this is a degree 2 polynomial in t :

$$\begin{aligned} f(t) &= \langle twx + y, twx + y \rangle \\ &= t^2 \langle x, x \rangle + tw \langle x, y \rangle + t\bar{w} \langle y, x \rangle + \langle y, y \rangle \\ &= t^2 \|x\|^2 + 2t \operatorname{Re}(w \langle x, y \rangle) + \|y\|^2 \end{aligned}$$

Since f is obviously positive, its discriminant must be negative:

$$4 \operatorname{Re}(w \langle x, y \rangle)^2 - 4 \|x\|^2 \cdot \|y\|^2 \leq 0$$

But this is equivalent to the following condition:

$$|\operatorname{Re}(w \langle x, y \rangle)| \leq \|x\| \cdot \|y\|$$

Now the point is that we can arrange for the number $w \in \mathbb{T}$ to be such that the quantity $w \langle x, y \rangle$ is real. Thus, we obtain the following inequality:

$$|\langle x, y \rangle| \leq \|x\| \cdot \|y\|$$

Finally, the study of the equality case is straightforward, by using the fact that the discriminant of f vanishes precisely when we have a root. But this leads to the conclusion in the statement, namely that the vectors x, y must be proportional. \square

As a second main result now, we have the Minkowski inequality:

THEOREM 2.4. *We have the Minkowski inequality*

$$\|x + y\| \leq \|x\| + \|y\|$$

and the equality case holds precisely when x, y are proportional.

PROOF. This follows indeed from the Cauchy-Schwarz inequality, as follows:

$$\begin{aligned}
& \|x + y\| \leq \|x\| + \|y\| \\
\iff & \|x + y\|^2 \leq (\|x\| + \|y\|)^2 \\
\iff & \|x\|^2 + \|y\|^2 + 2\operatorname{Re} \langle x, y \rangle \leq \|x\|^2 + \|y\|^2 + 2\|x\| \cdot \|y\| \\
\iff & \operatorname{Re} \langle x, y \rangle \leq \|x\| \cdot \|y\|
\end{aligned}$$

As for the equality case, this is clear from Cauchy-Schwarz as well. \square

As a consequence of this, we have the following result:

THEOREM 2.5. *The following function is a distance on H ,*

$$d(x, y) = \|x - y\|$$

in the usual sense, that of the abstract metric spaces.

PROOF. This follows indeed from the Minkowski inequality, which corresponds to the triangle inequality, the other two axioms for a distance being trivially satisfied. \square

The above result is quite important, because it shows that we can do geometry and analysis in our present setting, with distances and angles, a bit as in the finite dimensional case. In order to do such abstract geometry, we will often need the following key result, which shows that everything can be recovered in terms of distances:

PROPOSITION 2.6. *The scalar products can be recovered from distances, via the formula*

$$\begin{aligned}
4 \langle x, y \rangle &= \|x + y\|^2 - \|x - y\|^2 \\
&+ i\|x + iy\|^2 - i\|x - iy\|^2
\end{aligned}$$

called complex polarization identity.

PROOF. This is something that we have already met in finite dimensions. In arbitrary dimensions the proof is similar, as follows:

$$\begin{aligned}
& \|x + y\|^2 - \|x - y\|^2 + i\|x + iy\|^2 - i\|x - iy\|^2 \\
= & \|x\|^2 + \|y\|^2 - \|x\|^2 - \|y\|^2 + i\|x\|^2 + i\|y\|^2 - i\|x\|^2 - i\|y\|^2 \\
& + 2\operatorname{Re}(\langle x, y \rangle) + 2\operatorname{Re}(\langle x, y \rangle) + 2i\operatorname{Im}(\langle x, y \rangle) + 2i\operatorname{Im}(\langle x, y \rangle) \\
= & 4 \langle x, y \rangle
\end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

Let us discuss now some more advanced aspects. In order to do analysis on our spaces, we need the Cauchy sequences that we construct to converge. This is something which is automatic in finite dimensions, but in arbitrary dimensions, this can fail.

Thus, we must add an extra axiom, stating that H is complete with respect to the norm. It is convenient here to formulate a detailed new definition, as follows, which will be the starting point for our various considerations to follow:

DEFINITION 2.7. A Hilbert space is a complex vector space H given with a scalar product $\langle x, y \rangle$, satisfying the following conditions:

- (1) $\langle x, y \rangle$ is linear in x , and antilinear in y .
- (2) $\overline{\langle x, y \rangle} = \langle y, x \rangle$, for any x, y .
- (3) $\langle x, x \rangle \geq 0$, for any $x \neq 0$.
- (4) H is complete with respect to the norm $\|x\| = \sqrt{\langle x, x \rangle}$.

In other words, we have taken here Definition 2.1 above, and added the condition that H must be complete with respect to the norm $\|x\| = \sqrt{\langle x, x \rangle}$, that we know indeed to be a norm, according to the Minkowski inequality proved above.

As a basic example, we have the space $H = \mathbb{C}^N$, with its usual scalar product:

$$\langle x, y \rangle = \sum_i x_i \bar{y}_i$$

More generally now, we have the following construction of Hilbert spaces:

PROPOSITION 2.8. The sequences of complex numbers (x_i) which are square-summable,

$$\sum_i |x_i|^2 < \infty$$

form a Hilbert space $l^2(\mathbb{N})$, with the following scalar product:

$$\langle x, y \rangle = \sum_i x_i \bar{y}_i$$

In fact, given any index set I , we can construct a Hilbert space $l^2(I)$, in this way.

PROOF. The fact that we have indeed a complex vector space with a scalar product is elementary, and the fact that this space is indeed complete is very standard too. \square

On the other hand, we can talk as well about spaces of functions, as follows:

PROPOSITION 2.9. Given an interval $X \subset \mathbb{R}$, the quantity

$$\langle f, g \rangle = \int_X f(x) \overline{g(x)} dx$$

is a scalar product, making $H = L^2(X)$ a Hilbert space.

PROOF. Once again this is routine, coming this time from basic measure theory, with $H = L^2(X)$ being the space of square-integrable functions $f : X \rightarrow \mathbb{C}$, with the convention that two such functions are identified when they coincide almost everywhere. \square

We can unify the above two constructions, as follows:

THEOREM 2.10. *Given a measured space X , the quantity*

$$\langle f, g \rangle = \int_X f(x)\overline{g(x)}dx$$

is a scalar product, making $H = L^2(X)$ a Hilbert space.

PROOF. Here the first assertion is clear, and the fact that the Cauchy sequences converge is clear as well, by taking the pointwise limit, and using a standard argument. \square

Observe that with $X = \{1, \dots, N\}$ we obtain the space $H = \mathbb{C}^N$. Also, with $X = \mathbb{N}$, with the counting measure, we obtain the space $H = l^2(\mathbb{N})$. In fact, with an arbitrary set I , once again with the counting measure, we obtain the space $H = l^2(I)$.

Thus, as a conclusion of all this, the Hilbert space construction in Theorem 2.10 unifies all the Hilbert space constructions that we have so far.

Quite remarkably, the converse of this holds, in the sense that any Hilbert space must be of the form $L^2(X)$. This follows indeed from the following key result, which tells us that, in addition to this, we can always assume that $X = I$ is a discrete space:

THEOREM 2.11. *Let H be a Hilbert space.*

- (1) *Any algebraic basis of this space $\{f_i\}_{i \in I}$ can be turned into an orthonormal basis $\{e_i\}_{i \in I}$, by using the Gram-Schmidt procedure.*
- (2) *Thus, H has an orthonormal basis, and so we have $H \simeq l^2(I)$, with I being the indexing set for this orthonormal basis.*

PROOF. All this is standard, by recurrence in finite dimensions, using Gram-Schmidt, as stated, and by recurrence as well in infinite, countable dimensions. As for the case of infinite, uncountable dimensions, here the result holds as well, with the proof using technical transfinite recurrence arguments, borrowed from logic. \square

The above result, and its relation with Theorem 2.10, is something quite subtle, so let us further get into this. First, we have the following definition, based on the above:

DEFINITION 2.12. *A Hilbert space H is called separable when the following equivalent conditions are satisfied:*

- (1) *H has a countable algebraic basis $\{f_i\}_{i \in \mathbb{N}}$.*
- (2) *H has a countable orthonormal basis $\{e_i\}_{i \in \mathbb{N}}$.*
- (3) *We have $H \simeq l^2(\mathbb{N})$, isomorphism of Hilbert spaces.*

In what follows we will be mainly interested in the separable Hilbert spaces, where most of the questions coming from quantum physics take place. In view of the above, the following philosophical question appears: why not simply talking about $l^2(\mathbb{N})$?

In answer to this, we cannot really do so, because many of the separable spaces that we are interested in appear as spaces of functions, and such spaces do not necessarily have a very simple or explicit orthonormal basis, as shown by the following result:

PROPOSITION 2.13. *The Hilbert space $H = L^2[0, 1]$ is separable, having as orthonormal basis the orthonormalized version of the algebraic basis $f_n = x^n$ with $n \in \mathbb{N}$.*

PROOF. This follows from the Weierstrass theorem, which provides us with the basis $f_n = x^n$, which can be orthogonalized by using the Gram-Schmidt procedure, as explained in Theorem 2.11. Working out the details here is actually an excellent exercise. \square

As a conclusion to all this, we are interested in 1 space, namely the unique separable Hilbert space H , but due to various technical reasons, it is often better to forget that we have $H = l^2(\mathbb{N})$, and say instead that we have $H = L^2(X)$, with X being a separable measured space, or simply say that H is an abstract separable Hilbert space.

2b. Linear operators

Let us get now into the study of linear operators $T : H \rightarrow H$, which will eventually lead us into the correct infinite dimensional version of linear algebra. We first have:

THEOREM 2.14. *Let H be an arbitrary Hilbert space, with orthonormal basis $\{e_i\}_{i \in I}$. The algebra of all linear operators from H to itself,*

$$\mathcal{L}(H) = \left\{ T : H \rightarrow H \text{ linear} \right\}$$

embeds then into the algebra of the $I \times I$ complex matrices,

$$M_I(\mathbb{C}) = \left\{ (M_{ij})_{i,j \in I} \mid M_{ij} \in \mathbb{C} \right\}$$

with such a matrix M coming from the following operator,

$$Tx = Mx$$

and with an operator T corresponding to the following matrix:

$$M_{ij} = \langle Te_j, e_i \rangle$$

In the case $H = \mathbb{C}^N$ we obtain in this way the usual isomorphism $\mathcal{L}(H) \simeq M_N(\mathbb{C})$. In the separable case we obtain in this way a proper embedding $\mathcal{L}(H) \subset M_\infty(\mathbb{C})$.

PROOF. We have three assertions to be proved, the idea being as follows:

(1) Regarding the first assertion, given a linear operator $T : H \rightarrow H$, let us associate to it a matrix $M \in M_I(\mathbb{C})$ as in the statement, by the following formula:

$$M_{ij} = \langle Te_j, e_i \rangle$$

It is clear that this correspondence $T \rightarrow M$ is linear, and also that its kernel is $\{0\}$. Thus, we have an embedding of linear spaces, as follows:

$$\mathcal{L}(H) \subset M_I(\mathbb{C})$$

Our claim now is that this embedding is a morphism of algebras. But this is clear too, because if we denote by $T \rightarrow M_T$ our correspondence, we have:

$$\begin{aligned} (M_{ST})_{ij} &= \langle STe_j, e_i \rangle \\ &= \left\langle S \sum_k \langle Te_j, e_k \rangle e_k, e_i \right\rangle \\ &= \sum_k \langle Se_k, e_i \rangle \langle Te_j, e_k \rangle \\ &= \sum_k (M_S)_{ik} (M_T)_{kj} \\ &= (M_S M_T)_{ij} \end{aligned}$$

Finally, we must prove that the original operator $T : H \rightarrow H$ can be recovered from its matrix $M \in M_I(\mathbb{C})$ via the formula in the statement, namely $Tx = Mx$. But this latter formula holds for the vectors of the basis, $x = e_j$, because we have:

$$\begin{aligned} (Te_j)_i &= \langle Te_j, e_i \rangle \\ &= M_{ij} \\ &= (Me_j)_i \end{aligned}$$

Now by linearity we obtain from this that the formula $Tx = Mx$ holds everywhere, on any vector $x \in H$, and this finishes the proof of the first assertion.

(2) In finite dimensions we obtain an isomorphism, because any matrix $M \in M_N(\mathbb{C})$ determines an operator $T : \mathbb{C}^N \rightarrow \mathbb{C}^N$, according to the formula $\langle Te_j, e_i \rangle = M_{ij}$.

(3) In infinite dimensions, however, we do not have an isomorphism. For instance on $H = l^2(\mathbb{N})$ the following matrix does not define an operator:

$$M = \begin{pmatrix} 1 & 1 & \dots \\ 1 & 1 & \dots \\ \vdots & \vdots & \end{pmatrix}$$

Indeed, $T(e_1)$ should be the all-1 vector, but this vector is not square-summable. \square

In connection with our previous comments, the above result is something quite theoretical, because for basic Hilbert spaces like $L^2[0, 1]$, which do not have a simple orthonormal basis, the embedding $\mathcal{L}(H) \subset M_\infty(\mathbb{C})$ that we obtain is not something very useful.

In short, while the operators $T : H \rightarrow H$ are basically some infinite matrices, it is better to think of these operators as being objects on their own.

In what follows we will be interested in the operators $T : H \rightarrow H$ which are bounded. Regarding such operators, we have the following result:

THEOREM 2.15. *Given a Hilbert space H , the linear operators $T : H \rightarrow H$ which are bounded, in the sense that we have*

$$\|T\| = \sup_{\|x\| \leq 1} \|Tx\| < \infty$$

form a complex algebra with unit $B(H)$, having the property

$$\|ST\| \leq \|S\| \cdot \|T\|$$

and which is complete with respect to the norm.

PROOF. The fact that we have indeed an algebra, satisfying the product condition in the statement, follows from the following estimates, which are all elementary:

$$\|S + T\| \leq \|S\| + \|T\|$$

$$\|\lambda T\| = |\lambda| \cdot \|T\|$$

$$\|ST\| \leq \|S\| \cdot \|T\|$$

Regarding now the last assertion, if $\{T_n\} \subset B(H)$ is Cauchy then $\{T_n x\}$ is Cauchy for any $x \in H$, so we can define the limit $T = \lim_{n \rightarrow \infty} T_n$ by setting:

$$Tx = \lim_{n \rightarrow \infty} T_n x$$

Let us first check that the application $x \rightarrow Tx$ is linear. We have:

$$\begin{aligned} T(x + y) &= \lim_{n \rightarrow \infty} T_n(x + y) \\ &= \lim_{n \rightarrow \infty} T_n(x) + T_n(y) \\ &= \lim_{n \rightarrow \infty} T_n(x) + \lim_{n \rightarrow \infty} T_n(y) \\ &= T(x) + T(y) \end{aligned}$$

Similarly, we have as well the following computation:

$$\begin{aligned} T(\lambda x) &= \lim_{n \rightarrow \infty} T_n(\lambda x) \\ &= \lambda \lim_{n \rightarrow \infty} T_n(x) \\ &= \lambda T(x) \end{aligned}$$

Thus we have $T \in \mathcal{L}(H)$. It remains now to prove that we have $T \in B(H)$, and that we have $T_n \rightarrow T$ in norm. For this purpose, observe that we have:

$$\begin{aligned} & \|T_n - T_m\| \leq \varepsilon, \quad \forall n, m \geq N \\ \implies & \|T_n x - T_m x\| \leq \varepsilon, \quad \forall \|x\| = 1, \quad \forall n, m \geq N \\ \implies & \|T_n x - T x\| \leq \varepsilon, \quad \forall \|x\| = 1, \quad \forall n \geq N \\ \implies & \|T_N x - T x\| \leq \varepsilon, \quad \forall \|x\| = 1 \\ \implies & \|T_N - T\| \leq \varepsilon \end{aligned}$$

As a first consequence, we obtain $T \in B(H)$, because we have:

$$\begin{aligned} \|T\| &= \|T_N + (T - T_N)\| \\ &\leq \|T_N\| + \|T - T_N\| \\ &\leq \|T_N\| + \varepsilon \\ &< \infty \end{aligned}$$

As a second consequence, we obtain $T_N \rightarrow T$ in norm, and we are done. \square

As a first observation, in connection with our previous comments, in relation with the construction from Theorem 2.14 above, we have:

PROPOSITION 2.16. *We have embeddings of algebras as follows,*

$$B(H) \subset \mathcal{L}(H) \subset M_I(\mathbb{C})$$

which are both proper, in the infinite dimensional case.

PROOF. According to Theorem 2.14, the algebra $B(H)$ consists of the $I \times I$ complex matrices which define indeed linear maps $T : H \rightarrow H$, and which satisfy as well a second boundedness condition, coming from the boundedness of the norm of T :

$$\|T\| < \infty$$

In finite dimensions we have equalities everywhere, but in general this is not true, the standard example of a matrix which does not produce a linear operator being:

$$M = \begin{pmatrix} 1 & 1 & \dots \\ 1 & 1 & \dots \\ \vdots & \vdots & \end{pmatrix}$$

As for the examples of linear operators which are well-defined, on the whole Hilbert space, but which are not bounded, these can be constructed as well, by using methods from logic. We will not need these counterexamples in what follows. \square

As already mentioned after Theorem 2.14, all this is something quite theoretical, because for basic function spaces like $L^2[0, 1]$, which do not have a simple orthonormal basis, the embedding $B(H) \subset M_I(\mathbb{C})$ that we obtain is not very useful.

As a conclusion, while the bounded operators $T : H \rightarrow H$ are basically some infinite matrices, it is better to think of these operators as being objects on their own.

As in the finite dimensional case, we can talk about adjoint operators, in this setting, the definition and main properties of the construction $T \rightarrow T^*$ being as follows:

THEOREM 2.17. *Given a bounded operator $T \in B(H)$, the following formula defines a bounded operator $T^* \in B(H)$, called adjoint of T :*

$$\langle Tx, y \rangle = \langle x, T^*y \rangle$$

The correspondence $T \rightarrow T^$ is antilinear, antimultiplicative, and is an involution, and an isometry. In finite dimensions, we recover the usual adjoint operator.*

PROOF. There are several statements here, the proof being as follows:

(1) The existence of the adjoint operator T^* , given by the formula in the statement, comes from the fact that, given $x \in H$, the formula $\varphi(x) = \langle Tx, y \rangle$ defines a linear map $H \rightarrow \mathbb{C}$. Thus, we must have a formula as follows, for a certain vector $T^*y \in H$:

$$\varphi(x) = \langle x, T^*y \rangle$$

Moreover, this vector $T^*y \in H$ is unique with this property, and we conclude from this that the formula $y \rightarrow T^*y$ defines a certain map $T^* : H \rightarrow H$, which is unique with the property in the statement, namely $\langle Tx, y \rangle = \langle x, T^*y \rangle$ for any x, y .

(2) Let us prove that we have $T^* \in B(H)$. By using once again the uniqueness of T^* , we conclude that we have the following formulae, which show that T^* is linear:

$$T^*(x + y) = T^*x + T^*y$$

$$T^*(\lambda x) = \lambda T^*x$$

Observe also that T^* is bounded as well, because we have:

$$\begin{aligned} \|T\| &= \sup_{\|x\|=1} \sup_{\|y\|=1} \langle Tx, y \rangle \\ &= \sup_{\|y\|=1} \sup_{\|x\|=1} \langle x, T^*y \rangle \\ &= \|T^*\| \end{aligned}$$

(3) The fact that the correspondence $T \rightarrow T^*$ is antilinear, antimultiplicative, and is an involution comes from the following formulae, coming from uniqueness:

$$(S + T)^* = S^* + T^*$$

$$(\lambda T)^* = \bar{\lambda} T^*$$

$$(ST)^* = T^* S^*$$

$$(T^*)^* = T$$

As for the isometry property with respect to the operator norm, $\|T\| = \|T^*\|$, this is something that we already know, from the proof of (2) above.

(4) Regarding finite dimensions, let us first examine the general case where our Hilbert space comes with a basis, $H = l^2(I)$. We know from Theorem 2.14 that the operators $T \in B(H)$ correspond to matrices $M \in M_I(\mathbb{C})$, the connecting formula being:

$$M_{ij} = \langle Te_j, e_i \rangle$$

We can compute the matrix $M^* \in M_I(\mathbb{C})$ associated to the operator $T^* \in B(H)$, by using the formula $\langle Tx, y \rangle = \langle x, T^*y \rangle$, in the following way:

$$\begin{aligned} (M^*)_{ij} &= \langle T^*e_i, e_j \rangle \\ &= \overline{\langle e_j, T^*e_i \rangle} \\ &= \overline{\langle Te_j, e_i \rangle} \\ &= \overline{M_{ji}} \end{aligned}$$

Thus, we have reached to the usual formula for the adjoints of matrices, and in the particular case $H = \mathbb{C}^N$, we conclude that T^* comes indeed from the usual M^* . \square

As in finite dimensions, the operators T, T^* can be thought of as being “twin brothers”, and there is a lot of interesting mathematics connecting them.

We will be back to this later, on several occasions. For the moment, let us just record the following formula, which is something technical, and very useful:

THEOREM 2.18. *We have the following formula,*

$$\|TT^*\| = \|T\|^2$$

valid for any operator $T \in B(H)$.

PROOF. We recall from Theorem 2.17 that the correspondence $T \rightarrow T^*$ is an isometry with respect to the operator norm, in the sense that we have:

$$\|T\| = \|T^*\|$$

In order to prove now the formula in the statement, observe first that we have:

$$\|TT^*\| \leq \|T\| \cdot \|T^*\| = \|T\|^2$$

On the other hand, we have as well the following estimate:

$$\begin{aligned} \|T\|^2 &= \sup_{\|x\|=1} |\langle Tx, Tx \rangle| \\ &= \sup_{\|x\|=1} |\langle x, T^*Tx \rangle| \\ &\leq \|T^*T\| \end{aligned}$$

By replacing $T \rightarrow T^*$ we obtain from this that we have:

$$\|T\|^2 \leq \|TT^*\|$$

Thus, we have obtained the needed inequality, and we are done. \square

As a technical comment here, one may wonder whether there are other interesting norm formulae relating T, T^* , besides those that we already know, namely $\|T\| = \|T^*\|$ and $\|TT^*\| = \|T\|^2$. The answer here is non-trivial, basically stating that all possible norm formulae relating T, T^* , including the basic isometry formula $\|T\| = \|T^*\|$, can be deduced from the “master formula” $\|TT^*\| = \|T\|^2$. We will be back to this.

2c. Unitaries, projections

Let us discuss now some explicit examples of operators, in analogy with what happens in finite dimensions. The most basic examples of linear transformations are the rotations, symmetries and projections. Then, we have certain remarkable classes of linear transformations, such as the positive, self-adjoint and normal ones. In what follows we will develop the basic theory of such transformations, in the present Hilbert space setting.

Let us begin with the rotations. The situation here is quite tricky in arbitrary dimensions, and we have several notions instead of one. We first have the following result:

THEOREM 2.19. *For a linear operator $U \in B(H)$ the following conditions are equivalent, and if they are satisfied, we say that U is an isometry:*

- (1) U is a metric space isometry, $d(Ux, Uy) = d(x, y)$.
- (2) U is a normed space isometry, $\|Ux\| = \|x\|$.
- (3) U preserves the scalar product, $\langle Ux, Uy \rangle = \langle x, y \rangle$.
- (4) U satisfies the isometry condition $U^*U = 1$.

In finite dimensions, we recover in this way the usual unitary transformations.

PROOF. The proofs are similar to those in finite dimensions, as follows:

- (1) \iff (2) This follows indeed from the formula of the distances, namely:

$$d(x, y) = \|x - y\|$$

(2) \iff (3) This is once again standard. Indeed, we can pass from scalar products to distances, by using the following formula:

$$\|x\| = \sqrt{\langle x, x \rangle}$$

Conversely, we can compute the scalar products in terms of distances, by using the complex polarization identity, which is as follows:

$$\begin{aligned}
& \|x+y\|^2 - \|x-y\|^2 + i\|x+iy\|^2 - i\|x-iy\|^2 \\
= & \|x\|^2 + \|y\|^2 - \|x\|^2 - \|y\|^2 + i\|x\|^2 + i\|y\|^2 - i\|x\|^2 - i\|y\|^2 \\
& + 2\operatorname{Re}\langle x, y \rangle + 2\operatorname{Re}\langle x, y \rangle + 2i\operatorname{Im}\langle x, y \rangle + 2i\operatorname{Im}\langle x, y \rangle \\
= & 4\langle x, y \rangle
\end{aligned}$$

(3) \iff (4) We have indeed the following equivalences, by using the standard formula $\langle Tx, y \rangle = \langle x, T^*y \rangle$, which defines the adjoint operator:

$$\begin{aligned}
\langle Ux, Uy \rangle = \langle x, y \rangle & \iff \langle x, U^*Uy \rangle = \langle x, y \rangle \\
& \iff U^*Uy = y \\
& \iff U^*U = 1
\end{aligned}$$

Thus, we are led to the conclusions in the statement. \square

The point now is that the condition $U^*U = 1$ does not imply in general $UU^* = 1$, the simplest counterexample here being the shift operator on $l^2(\mathbb{N})$:

PROPOSITION 2.20. *The shift operator on the space $l^2(\mathbb{N})$, given by*

$$S(e_i) = e_{i+1}$$

*is an isometry, $S^*S = 1$. However, we have $SS^* \neq 1$.*

PROOF. The adjoint of the shift is given by the following formula:

$$S^*(e_i) = \begin{cases} e_{i-1} & \text{if } i > 0 \\ 0 & \text{if } i = 0 \end{cases}$$

When composing S, S^* , in one sense we obtain the following formula:

$$S^*S(e_i) = e_i$$

In other other sense now, we obtain the following formula:

$$SS^*(e_i) = \begin{cases} e_i & \text{if } i > 0 \\ 0 & \text{if } i = 0 \end{cases}$$

Summarizing, the compositions are given by the following formulae:

$$S^*S = 1$$

$$SS^* = \operatorname{Proj}(e_0^\perp)$$

Thus, we are led to the conclusions in the statement. \square

As a conclusion, the notion of isometry is not the correct infinite dimensional analogue of the notion of unitary, and the unitary operators must be introduced as follows:

THEOREM 2.21. *For a linear operator $U \in B(H)$ the following conditions are equivalent, and if they are satisfied, we say that U is a unitary:*

- (1) U is an isometry, which is invertible.
- (2) U, U^{-1} are both isometries.
- (3) U, U^* are both isometries.
- (4) $UU^* = U^*U = 1$.
- (5) $U^* = U^{-1}$.

Moreover, the unitary operators form a group $U(H) \subset B(H)$.

PROOF. There are several statements here, the idea being as follows:

(1) The various equivalences in the statement are all clear from definitions, and from Theorem 2.19 in what regards the various possible notions of isometries which can be used, by using the formula $(ST)^* = T^*S^*$ for the adjoints of the products of operators.

(2) The fact that the products and inverses of unitaries are unitaries is also clear, and we conclude that the unitary operators form a group $U(H) \subset B(H)$, as stated. \square

As a conclusion to all this, in infinite dimensions we have three notions in relation with the usual rotations, namely the isometries, the adjoints of isometries, and the unitaries. The example to be kept in mind is the shift operator on S on the separable Hilbert space $l^2(\mathbb{N})$, which is an isometry, $S^*S = 1$, but not a unitary, $SS^* \neq 1$.

Let us discuss now the projections. Modulo the fact that all the subspaces $K \subset H$ where these projections project must be assumed to be closed, in the present setting, here the result is perfectly similar to the one in finite dimensions, as follows:

THEOREM 2.22. *For a linear operator $P \in B(H)$ the following conditions are equivalent, and if they are satisfied, we say that P is a projection:*

- (1) P is the orthogonal projection on a closed subspace $K \subset H$.
- (2) P satisfies the projection equations $P^2 = P = P^*$.

PROOF. As in finite dimensions, P is an abstract projection, not necessarily orthogonal, when it is an idempotent, algebraically speaking, in the sense that we have:

$$P^2 = P$$

The point now is that this projection is orthogonal when:

$$\begin{aligned} \langle Px - x, Py \rangle = 0 &\iff \langle P^*Px - P^*x, y \rangle = 0 \\ &\iff P^*Px - P^*x = 0 \\ &\iff P^*P - P^* = 0 \\ &\iff P^*P = P^* \end{aligned}$$

Thus we must have $P^* = P^*P$. Now observe that by conjugating, we obtain:

$$\begin{aligned} P &= (P^*P)^* \\ &= P^*(P^*)^* \\ &= P^*P \end{aligned}$$

Now by comparing with the original relation, $P^* = P^*P$, we conclude that $P = P^*$. Thus, we have shown that any orthogonal projection must satisfy, as claimed:

$$P^2 = P = P^*$$

Conversely, if this condition is satisfied, $P^2 = P$ shows that P is a projection, and $P = P^*$ shows via the above computation that P is indeed orthogonal. \square

There is a relation between the projections and the general isometries, such as the shift S that we met before, and we have the following result:

PROPOSITION 2.23. *Given an isometry $U \in B(H)$, the following operator*

$$P = UU^*$$

is a projection, namely the orthogonal projection on $\text{Im}(U)$.

PROOF. Assume indeed that we have an isometry, $U^*U = 1$. The fact that $P = UU^*$ is indeed a projection can be checked abstractly, as follows:

$$(UU^*)^* = UU^*$$

$$UU^*UU^* = UU^*$$

As for the last assertion, this is something that we already met, for the shift, and the situation in general is similar, with the result itself being clear. \square

More generally now, along the same lines, and clarifying the whole situation with the unitaries and isometries, we have the following result:

THEOREM 2.24. *An operator $U \in B(H)$ is a partial isometry, in the usual geometric sense, when the following two operators are projections:*

$$P = UU^*$$

$$Q = U^*U$$

Moreover, the isometries, adjoints of isometries and unitaries are respectively characterized by the conditions $Q = 1$, $P = 1$, $P = Q = 1$.

PROOF. The first assertion is a straightforward extension of Proposition 2.23, and the second assertion follows from various results regarding isometries established above. \square

Summarizing, we have infinite dimensional analogues of the rotations, projections and partial isometries, and the theory is quite similar to what we knew before, with a few twists, however, coming from counterexamples such as the shift operator on $l^2(\mathbb{N})$.

It is possible to talk as well about symmetries, in the following way:

DEFINITION 2.25. *An operator $S \in B(H)$ is called a symmetry when $S^2 = 1$, and a unitary symmetry when one of the following equivalent conditions is satisfied:*

- (1) S is a unitary, $S^* = S^{-1}$, and a symmetry as well, $S^2 = 1$.
- (2) S satisfies the equations $S = S^* = S^{-1}$.

As in the finite dimensional case, there are many interesting examples of symmetries. We will be back to this, and to some more general theory as well, later on.

With these basic examples of operators discussed, let us study now some larger classes of operators, which are of particular importance. The idea indeed is that in finite dimensions we have the positive operators, the self-adjoint operators and the normal operators, and our next task will be that of talking about such operators, in infinite dimensions.

Let us start with the self-adjoint operators. The basic theory here is as follows:

THEOREM 2.26. *For an operator $T \in B(H)$, the following conditions are equivalent, and if they are satisfied, we call T self-adjoint:*

- (1) $T = T^*$.
- (2) $\langle Tx, x \rangle \in \mathbb{R}$.

In finite dimensions, we recover in this way the usual self-adjointness notion.

PROOF. There are several assertions here, the idea being as follows:

- (1) \implies (2) This is clear, because we have:

$$\begin{aligned} \overline{\langle Tx, x \rangle} &= \langle x, Tx \rangle \\ &= \langle T^*x, x \rangle \\ &= \langle Tx, x \rangle \end{aligned}$$

(2) \implies (1) This is clear as well, because the beginning of the above computation shows that, when assuming $\langle Tx, x \rangle \in \mathbb{R}$, the following happens:

$$\langle Tx, x \rangle = \langle T^*x, x \rangle$$

Thus, in terms of the operator $S = T - T^*$, we have:

$$\langle Sx, x \rangle = 0$$

Now by using the polarization identity we obtain from this that we have $\|Sx\| = 0$, and so $S = 0$. But this proves that we have $T = T^*$, as desired.

(3) Finally, in what regards the finite dimensions, or more generally the case where our Hilbert space comes with a basis, $H = l^2(I)$, here the condition $T = T^*$ corresponds to the usual self-adjointness condition $M = M^*$ at the level of the associated matrices. \square

At the level of the basic examples, the situation is as follows:

PROPOSITION 2.27. *The following operators are self-adjoint:*

- (1) *The projections, $P^2 = P = P^*$. In fact, an abstract, algebraic projection is an orthogonal projection precisely when it is self-adjoint.*
- (2) *The unitary symmetries, $S = S^* = S^{-1}$. In fact, a unitary is a unitary symmetry precisely when it is self-adjoint.*

PROOF. The fact that the projections and the unitary symmetries are indeed self-adjoint is clear, the self-adjointness condition being part of the equations to be satisfied. As for the other assertions, these are things that we know, which are trivial as well. \square

Next in line, we have the notion of positive operator. We have here:

THEOREM 2.28. *The positive operators, which are the operators $T \in B(H)$ satisfying $\langle Tx, x \rangle \geq 0$, have the following properties:*

- (1) *They are self-adjoint, $T = T^*$.*
- (2) *As examples, we have the projections, $P^2 = P = P^*$.*
- (3) *More generally, $T = SS^*$ is positive, for any $S \in B(H)$.*
- (4) *In finite dimensions, we recover the usual positive operators.*

PROOF. All these assertions are elementary, the idea being as follows:

(1) This follows from the equivalence in Theorem 2.26 above, because the positivity condition $\langle Tx, x \rangle \geq 0$ assumes that we have $\langle Tx, x \rangle \in \mathbb{R}$.

(2) This is clear from $P^2 = P = P^*$, because we have:

$$\begin{aligned} \langle Px, x \rangle &= \langle P^2x, x \rangle \\ &= \langle Px, P^*x \rangle \\ &= \langle Px, Px \rangle \\ &= \|Px\|^2 \end{aligned}$$

(3) This follows from a similar computation, namely:

$$\begin{aligned} \langle SS^*x, x \rangle &= \langle S^*x, S^*x \rangle \\ &= \|S^*x\|^2 \end{aligned}$$

(4) This is well-known, the idea being that the condition $\langle Tx, x \rangle \geq 0$ corresponds to the usual positivity condition $A \geq 0$, at the level of the associated matrix. \square

Once again in analogy with what we know from finite dimensions, it is possible to talk as well about strictly positive operators, and we have here the following result:

THEOREM 2.29. *The strictly positive operators, which are the operators $T \in B(H)$ satisfying $\langle Tx, x \rangle > 0$, for any $x \neq 0$, have the following properties:*

- (1) *They are self-adjoint, $T = T^*$.*
- (2) *As examples, $T = SS^*$ is positive, for any $S \in B(H)$ invertible.*
- (3) *In finite dimensions, we recover the usual strictly positive operators.*

PROOF. As before, all these assertions are elementary, the idea being as follows:

(1) This follows from the equivalence in Theorem 2.26 above, because the positivity condition $\langle Tx, x \rangle > 0$ assumes that we have $\langle Tx, x \rangle \in \mathbb{R}$.

(2) This follows from the injectivity of S^* , because for any $x \neq 0$ we have:

$$\begin{aligned} \langle SS^*x, x \rangle &= \langle S^*x, S^*x \rangle \\ &= \|S^*x\|^2 \\ &> 0 \end{aligned}$$

(3) This is well-known, the idea being that the condition $\langle Tx, x \rangle > 0$ corresponds to the usual strict positivity condition $A > 0$, at the level of the associated matrix. \square

As a comment here, in relation with the above, there are in fact some subtleties in the infinite dimensional case. Indeed, while any strictly positive matrix $A > 0$ is well-known to be invertible, with this being standard, the analogue of this fact does not hold in infinite dimensions. Consider indeed the following operator on $l^2(\mathbb{N})$:

$$T = \begin{pmatrix} 1 & & & \\ & \frac{1}{2} & & \\ & & \frac{1}{3} & \\ & & & \ddots \end{pmatrix}$$

This operator T is well-defined and bounded, and we have $\langle Tx, x \rangle > 0$ for any $x \neq 0$. However T is not invertible, and so we have our counterexample.

We will be back to positive and strictly positive operators later on, in chapter 4 below, with a systematic study of them, including the construction of square roots.

As a last remarkable class of operators, we have the normal ones. We have here:

THEOREM 2.30. *For an operator $T \in B(H)$, the following conditions are equivalent, and if they are satisfied, we call T normal:*

- (1) $TT^* = T^*T$.
- (2) $\|Tx\| = \|T^*x\|$.

In finite dimensions, we recover in this way the usual normality notion.

PROOF. There are several assertions here, the idea being as follows:

(1) \implies (2) This is clear, because we have:

$$\begin{aligned} \|Tx\|^2 &= \langle Tx, Tx \rangle \\ &= \langle T^*Tx, x \rangle \\ &= \langle TT^*x, x \rangle \\ &= \langle T^*x, T^*x \rangle \\ &= \|T^*x\|^2 \end{aligned}$$

(2) \implies (1) This is clear as well, because the above computation shows that, when assuming $\|Tx\| = \|T^*x\|$, the following happens:

$$\langle TT^*x, x \rangle = \langle T^*Tx, x \rangle$$

Thus, in terms of the operator $S = TT^* - T^*T$, we have:

$$\langle Sx, x \rangle = 0$$

Now by using the polarization identity we obtain from this that we have:

$$\|Sx\| = 0$$

Thus $S = 0$. But this proves that we have $TT^* = T^*T$, as desired.

(3) Finally, in what regards the finite dimensions, or more generally the case where our Hilbert space comes with a basis, $H = l^2(I)$, here the condition $TT^* = T^*T$ corresponds to the usual normality condition $MM^* = M^*M$ at the level of the associated matrices. \square

As a conclusion to all this, we have all sorts of classes of operators. By ignoring the partial isometries, unitaries, symmetries and the strictly positive operators, which are more specialized objects, we have the following hierarchy of bounded operators:

- (1) Projections, $P^2 = P = P^*$.
- (2) Positive operators, $T \geq 0$.
- (3) Self-adjoint operators, $T = T^*$.
- (4) Normal operators, $TT^* = T^*T$.

In addition, the class of normal operators covers all the unitaries, and in particular, all the unitary symmetries. The non-unitary isometries, however, are not normal.

2d. Diagonal operators

Let us work out now what happens in the case that we are mostly interested in, namely $H = L^2(X)$, with X being a measured space. We first have:

THEOREM 2.31. *Given a measured space X , consider the Hilbert space $H = L^2(X)$. Associated to any function $f \in L^\infty(X)$ is then the multiplication operator*

$$T_f : H \rightarrow H$$

$$T_f(g) = fg$$

which is well-defined, linear and bounded, having norm as follows:

$$\|T_f\| = \|f\|_\infty$$

Moreover, the correspondence $f \rightarrow T_f$ is linear, multiplicative and involutive.

PROOF. There are several assertions here, the idea being as follows:

(1) We must first prove that the formula in the statement, $T_f(g) = fg$, defines indeed an operator $H \rightarrow H$, which amounts in saying that we have:

$$f \in L^\infty(X), g \in L^2(X) \implies fg \in L^2(X)$$

But this follows from the following explicit estimate:

$$\begin{aligned} \|fg\|_2 &= \sqrt{\int_X |f(x)|^2 |g(x)|^2 d\mu(x)} \\ &\leq \sup_{x \in X} |f(x)|^2 \sqrt{\int_X |g(x)|^2 d\mu(x)} \\ &= \|f\|_\infty \|g\|_2 \\ &< \infty \end{aligned}$$

(2) Next in line, we must prove that T is linear and bounded. We have:

$$T_f(g + h) = T_f(g) + T_f(h)$$

$$T_f(\lambda g) = \lambda T_f(g)$$

As for the boundedness condition, this follows from the estimate from the proof of (1), which gives, in terms of the operator norm of $B(H)$:

$$\|T_f\| \leq \|f\|_\infty$$

(3) Let us prove now that we have equality, $\|T_f\| = \|f\|_\infty$, in the above estimate. For this purpose, we use the well-known fact that the L^∞ functions can be approximated by

L^2 functions. Indeed, with such an approximation $g_n \rightarrow f$ we obtain:

$$\begin{aligned} \|fg_n\|_2 &= \sqrt{\int_X |f(x)|^2 |g_n(x)|^2 d\mu(x)} \\ &\simeq \sup_{x \in X} |f(x)|^2 \sqrt{\int_X |g_n(x)|^2 d\mu(x)} \\ &= \|f\|_\infty \|g_n\|_2 \end{aligned}$$

Thus, with $n \rightarrow \infty$ we obtain the following inequality:

$$\|T_f\| \geq \|f\|_\infty$$

But this is precisely the inequality which is reverse to the one obtained in the above proof of (2), and this leads to the conclusion in the statement.

(4) Regarding now the fact that the correspondence $f \rightarrow T_f$ is indeed linear and multiplicative, the corresponding formulae are as follows, both clear:

$$\begin{aligned} T_{f+h}(g) &= T_f(g) + T_h(g) \\ T_{\lambda f}(g) &= \lambda T_f(g) \end{aligned}$$

(5) Finally, let us prove that the correspondence $f \rightarrow T_f$ is involutive, in the sense that it transforms the standard involution $f \rightarrow \bar{f}$ of the algebra $L^\infty(X)$ into the standard involution $T \rightarrow T^*$ of the algebra $B(H)$. We must prove that we have:

$$T_f^* = T_{\bar{f}}$$

But this follows from the following computation:

$$\begin{aligned} \langle T_f g, h \rangle &= \langle fg, h \rangle \\ &= \int_X f(x)g(x)\bar{h}(x)d\mu(x) \\ &= \int_X g(x)f(x)\bar{h}(x)d\mu(x) \\ &= \langle g, \bar{f}h \rangle \\ &= \langle g, T_{\bar{f}}h \rangle \end{aligned}$$

Indeed, since the adjoint is unique, we obtain from this that we have:

$$T_f^* = T_{\bar{f}}$$

Thus the correspondence $f \rightarrow T_f$ is indeed involutive, as claimed. \square

In what regards now the basic classes of operators, the above construction provides us with many new examples, which are very explicit, and are complementary to the usual finite dimensional examples that we usually have in mind, as follows:

THEOREM 2.32. *The multiplication operators $T_f(g) = fg$ on the Hilbert space $H = L^2(X)$ associated to the functions $f \in L^\infty(X)$ are as follows:*

- (1) T_f is unitary when $f : X \rightarrow \mathbb{T}$.
- (2) T_f is a symmetry when $f : X \rightarrow \{-1, 1\}$.
- (3) T_f is a projection when $f = \chi_Y$ with $Y \in X$.
- (4) There are no non-unitary isometries.
- (5) There are no non-unitary symmetries.
- (6) T_f is positive when $f : X \rightarrow \mathbb{R}_+$.
- (7) T_f is self-adjoint when $f : X \rightarrow \mathbb{R}$.
- (8) T_f is always normal, for any $f : X \rightarrow \mathbb{C}$.

PROOF. All these assertions are clear from definitions, and from the various properties of the correspondence $f \rightarrow T_f$, established above, as follows:

(1) The unitarity condition $U^* = U^{-1}$ for the operator T_f reads $\bar{f} = f^{-1}$, which means that we must have $f : X \rightarrow \mathbb{T}$, as claimed.

(2) The symmetry condition $S^2 = 1$ for the operator T_f reads $f^2 = 1$, which means that we must have $f : X \rightarrow \{-1, 1\}$, as claimed.

(3) The projection condition $P^2 = P = P^*$ for the operator T_f reads $f^2 = f = \bar{f}$, which means that we must have $f : X \rightarrow \{0, 1\}$, or equivalently, $f = \chi_Y$ with $Y \subset X$.

(4) A non-unitary isometry must satisfy by definition $U^*U = 1, UU^* \neq 1$, and for the operator T_f this means that we must have $|f|^2 = 1, |f|^2 \neq 1$, which is impossible.

(5) This follows from (1) and (2), because the solutions found in (2) for the symmetry problem are included in the solutions found in (1) for the unitarity problem.

(6) The fact that T_f is positive amounts in saying that we must have $\langle fg, g \rangle \geq 0$ for any $g \in L^2(X)$, and this is equivalent to the fact that we must have $f \geq 0$, as desired.

(7) The self-adjointness condition $T = T^*$ for the operator T_f reads $f = \bar{f}$, which means that we must have $f : X \rightarrow \mathbb{R}$, as claimed.

(8) The normality condition $TT^* = T^*T$ for the operator T_f reads $f\bar{f} = \bar{f}f$, which is automatic for any function $f : X \rightarrow \mathbb{C}$, as claimed. \square

The above result might look quite puzzling, at a first glance, messing up our intuition with various classes of operators, coming from usual linear algebra. However, a bit of further thinking tells us that there is no contradiction, and that Theorem 2.32 in fact is very similar to what we know about the diagonal matrices.

To be more precise, the diagonal matrices are unitaries precisely when their entries are in \mathbb{T} , there are no non-unitary isometries, all such matrices are normal, and so on.

In order to understand all this, and fully recover from Theorem 2.32, let us work out what happens with the correspondence $f \rightarrow T_f$, in finite dimensions. The situation here is in fact extremely simple, and illuminating, as follows:

THEOREM 2.33. *Assuming $X = \{1, \dots, N\}$ with the counting measure, the embedding*

$$L^\infty(X) \subset B(L^2(X))$$

constructed via multiplication operators, $T_f(g) = fg$, corresponds to the embedding

$$\mathbb{C}^N \subset M_N(\mathbb{C})$$

given by the diagonal matrices, constructed as follows:

$$f \rightarrow \text{diag}(f_1, \dots, f_N)$$

Thus, Theorem 2.31 generalizes what we know about the diagonal matrices.

PROOF. The idea is that all this is trivial, with not a single new computation needed, modulo some algebraic thinking, of quite soft type.

Let us go back indeed to Theorem 2.31 above and its proof, with the abstract measured space X appearing there being now the following finite space, with its counting measure:

$$X = \{1, \dots, N\}$$

Regarding the functions $f \in L^\infty(X)$, these are now functions as follows:

$$f : \{1, \dots, N\} \rightarrow \mathbb{C}$$

We can identify such a function with the corresponding vector $(f(i))_i \in \mathbb{C}^N$, and so we conclude that our input algebra $L^\infty(X)$ is the algebra \mathbb{C}^N :

$$L^\infty(X) = \mathbb{C}^N$$

Regarding now the Hilbert space $H = L^2(X)$, this is equal as well to \mathbb{C}^N , and for the same reasons, namely that $g \in L^2(X)$ can be identified with the vector $(g(i))_i \in \mathbb{C}^N$:

$$L^2(X) = \mathbb{C}^N$$

Observe that, due to our assumption that X comes with its counting measure, the scalar product that we obtain on \mathbb{C}^N is the usual one, without weights. Now, let us identify the operators on $L^2(X) = \mathbb{C}^N$ with the square matrices, in the usual way:

$$B(L^2(X)) = M_N(\mathbb{C})$$

This was our final identification, in order to get started. Now by getting back to Theorem 2.31, the embedding $L^\infty(X) \subset B(L^2(X))$ constructed there reads:

$$\mathbb{C}^N \subset M_N(\mathbb{C})$$

But this can only be the embedding given by the diagonal matrices, so are basically done. In order to finish, however, let us understand what the operator associated to an arbitrary vector $f \in \mathbb{C}^N$ is. We can regard this vector as a function, $f(i) = f_i$, and so

the action $T_f(g) = fg$ on the vectors of $L^2(X) = \mathbb{C}^N$ is by componentwise multiplication by the numbers f_1, \dots, f_N . But this is exactly the action of the diagonal matrix $\text{diag}(f_1, \dots, f_N)$, and so we are led to the conclusion in the statement. \square

There are other things that can be said about the embedding $L^\infty(X) \subset B(L^2(X))$, a key observation here, which is elementary to prove, being the fact that the image of $L^\infty(X)$ in this embedding is weakly closed. Thus, $L^\infty(X)$ is naturally a von Neumann algebra, on $L^2(X)$. We will be back to this, on numerous occasions, in what follows.

2e. Exercises

As before with linear algebra, operator theory is a wide area of mathematics, and there are many interesting operators, and exercises about them. We first have:

EXERCISE 2.34. *Find an explicit orthonormal basis for the Hilbert space*

$$H = L^2[0, 1]$$

by starting with the algebraic basis $f_n = x^n$ with $n \in \mathbb{N}$, and applying Gram-Schmidt.

This is actually quite non-trivial, and in case you're stuck with complicated computations, better look it up, and then write an account of what you found.

As a second exercise now, instructive and annoying as well, we have:

EXERCISE 2.35. *Clarify whether the linear operators*

$$T : H \rightarrow H$$

are automatically bounded, or not.

This was briefly discussed in the above, with that comment that the answer is in general is “no”, due to opaque reasons, coming from logic. Time to have this done.

CHAPTER 3

Spectral theorems

3a. Basic theory

We discuss in this chapter the diagonalization problem for the operators $T \in B(H)$, in analogy with the diagonalization problem for the usual matrices $A \in M_N(\mathbb{C})$. As in the usual matrix case, our main results will concern the case of the normal operators.

As a first observation, we can talk about eigenvalues and eigenvectors of arbitrary operators $T \in B(H)$, in the obvious way, as follows:

DEFINITION 3.1. *Given an operator $T \in B(H)$, assuming that we have*

$$Tx = \lambda x$$

we say that $x \in H$ is an eigenvector of T , with eigenvalue $\lambda \in \mathbb{C}$.

We know many basic things about the eigenvalues and eigenvectors, in the finite dimensional case. However, most of these will not extend to the infinite dimensional case, or at least not extend in a straightforward way, due to a number of reasons:

- (1) Most of the basic linear algebra is based on the fact that $Tx = \lambda x$ is equivalent to $(T - \lambda)x = 0$, so that λ is an eigenvalue precisely when $T - \lambda$ is not invertible. In the infinite dimensional setting $T - \lambda$ might be injective and not surjective, or vice versa, or invertible with $(T - \lambda)^{-1}$ not bounded, and so on.
- (2) The other basic fact is that $T - \lambda$ is not invertible precisely when $\det(T - \lambda) = 0$. In infinite dimensions, however, it is impossible to construct a determinant function $\det : B(H) \rightarrow \mathbb{C}$, and this for a myriad reasons, for instance because any definition of \det requires some kind of algorithm, which must stop.

Summarizing, we are in big trouble with our extension program, and this right from the beginning. In order to have some theory started, however, let us forget about (2), which obviously leads nowhere, and focus on the difficulties in (1).

In order to cut short the discussion there, regarding the various properties of $T - \lambda$, we can just say that $T - \lambda$ is either invertible with bounded inverse, the “good case”, or not. We are led in this way to the following definition:

DEFINITION 3.2. *The spectrum of an operator $T \in B(H)$ is the set*

$$\sigma(T) = \left\{ \lambda \in \mathbb{C} \mid T - \lambda \notin B(H)^{-1} \right\}$$

where $B(H)^{-1} \subset B(H)$ is the set of invertible operators.

As a basic example, in the finite dimensional case, $H = \mathbb{C}^N$, the spectrum of a usual matrix $A \in M_N(\mathbb{C})$ is the collection of its eigenvalues, taken without multiplicities. We will see many other explicit examples of spectra, in what follows.

In general, the spectrum has the following properties:

PROPOSITION 3.3. *The spectrum of $T \in B(H)$ contains the eigenvalue set*

$$\varepsilon(T) = \left\{ \lambda \in \mathbb{C} \mid \ker(T - \lambda) \neq \{0\} \right\}$$

and $\varepsilon(T) \subset \sigma(T)$ is an equality in finite dimensions, but not in infinite dimensions.

PROOF. We have several assertions here, the idea being as follows:

(1) First of all, the eigenvalue set is indeed the one in the statement, because $Tx = \lambda x$ tells us precisely that $T - \lambda$ must be not injective.

(2) The fact that we have $\varepsilon(T) \subset \sigma(T)$ is clear as well, because if $T - \lambda$ is not injective, it is not bijective, save of being bijective with bounded inverse.

(3) In finite dimensions we have $\varepsilon(T) = \sigma(T)$, because $T - \lambda$ is injective if and only if it is bijective, with the boundedness of the inverse being automatic.

(4) In infinite dimensions we can assume $H = l^2(\mathbb{N})$, and here we have the operator $T(e_i) = e_{2i}$, which is injective but not surjective. Thus $0 \in \sigma(T) - \varepsilon(T)$. \square

We will see more examples and counterexamples, and general theory, in a moment. Philosophically speaking, the best way of thinking at all this is as follows:

- (1) The numbers $\lambda \notin \sigma(T)$ are good, because we can invert $T - \lambda$.
- (2) The numbers $\lambda \in \sigma(T) - \varepsilon(T)$ are bad.
- (3) The eigenvalues $\lambda \in \varepsilon(T)$ are evil.

Note that this is somewhat contrary to what happens in linear algebra, where the eigenvalues are highly valued, and cherished, and regarded as being the source of all good things on Earth. Welcome to operator theory, where some things are upside down.

Let us develop now some general theory for the spectrum, or perhaps for its complement, with the promise to come back to eigenvalues later.

As a first result, we would like to prove that the spectra are non-empty. This is something non-trivial even for the usual matrices, with the standard proof using two non-trivial ingredients, namely the fact that the eigenvalues are roots of the characteristic polynomial, and the fact that any complex polynomial has a root.

In our general setting, in order to deal with such questions, we will need:

PROPOSITION 3.4. *The following happen:*

- (1) $\|T\| < 1 \implies (1 - T)^{-1} = 1 + T + T^2 + \dots$
- (2) *The set $B(H)^{-1}$ is open.*
- (3) *The map $T \rightarrow T^{-1}$ is differentiable.*

PROOF. All these assertions are elementary, as follows:

(1) This follows as in the scalar case, the computation being as follows, provided that everything converges under the norm, which amounts in saying that $\|T\| < 1$:

$$\begin{aligned} & (1 - T)(1 + T + T^2 + \dots) \\ &= 1 - T + T - T^2 + T^2 - T^3 + \dots \\ &= 1 \end{aligned}$$

(2) Assuming $T \in B(H)^{-1}$, let us pick $S \in B(H)$ such that:

$$\|T - S\| < \frac{1}{\|T^{-1}\|}$$

We have then the following estimate:

$$\begin{aligned} \|1 - T^{-1}S\| &= \|T^{-1}(T - S)\| \\ &\leq \|T^{-1}\| \cdot \|T - S\| \\ &< 1 \end{aligned}$$

Thus by (1) we obtain $T^{-1}S \in B(H)^{-1}$, and so $S \in B(H)^{-1}$, as desired.

(3) This follows as in the scalar case, where the derivative of $f(t) = t^{-1}$ is:

$$f'(t) = -t^{-2}$$

To be more precise, in the present normed space setting the derivative is no longer a number, but rather a linear transformation. But this linear transformation can be found by developing $f(T) = T^{-1}$ at order 1, as follows:

$$\begin{aligned} (T + S)^{-1} &= ((1 + ST^{-1})T)^{-1} \\ &= T^{-1}(1 + ST^{-1})^{-1} \\ &= T^{-1}(1 - ST^{-1} + (ST^{-1})^2 - \dots) \\ &\simeq T^{-1}(1 - ST^{-1}) \\ &= T^{-1} - T^{-1}ST^{-1} \end{aligned}$$

We conclude that the derivative that we are looking for is:

$$f'(T)S = -T^{-1}ST^{-1}$$

Thus, we are led to the conclusion in the statement. \square

We can now formulate our first theorem about spectra, as follows:

THEOREM 3.5. *The spectrum of a bounded operator $T \in B(H)$ is:*

- (1) *Compact.*
- (2) *Contained in the disc $D_0(\|T\|)$.*
- (3) *Non-empty.*

PROOF. This can be proved by using the above results, as follows:

(1) In view of (2) below, it is enough to prove that $\sigma(T)$ is closed. But this follows from the following computation, with $|\varepsilon|$ being small:

$$\begin{aligned} \lambda \notin \sigma(T) &\implies T - \lambda \in B(H)^{-1} \\ &\implies T - \lambda - \varepsilon \in B(H)^{-1} \\ &\implies \lambda + \varepsilon \notin \sigma(T) \end{aligned}$$

(2) This follows from the following computation:

$$\begin{aligned} \lambda > \|T\| &\implies \left\| \frac{T}{\lambda} \right\| < 1 \\ &\implies 1 - \frac{T}{\lambda} \in B(H)^{-1} \\ &\implies \lambda - T \in B(H)^{-1} \\ &\implies \lambda \notin \sigma(T) \end{aligned}$$

(3) Assume by contradiction $\sigma(T) = \emptyset$. Given a linear form $f \in B(H)^*$, consider the following map, which is well-defined, due to our assumption $\sigma(T) = \emptyset$:

$$\begin{aligned} \varphi : \mathbb{C} &\rightarrow \mathbb{C} \\ \lambda &\rightarrow f((T - \lambda)^{-1}) \end{aligned}$$

By using the fact that $T \rightarrow T^{-1}$ is differentiable, that we know from Proposition 3.4, we conclude that this map is differentiable, and so holomorphic. Also, we have:

$$\begin{aligned} \lambda \rightarrow \infty &\implies T - \lambda \rightarrow \infty \\ &\implies (T - \lambda)^{-1} \rightarrow 0 \\ &\implies f((T - \lambda)^{-1}) \rightarrow 0 \end{aligned}$$

Thus by the Liouville theorem we obtain:

$$\varphi = 0$$

With this in hand, by Hahn-Banach we obtain then:

$$(T - \lambda)^{-1} = 0$$

But this is a contradiction, as desired. \square

Here is now a second basic result regarding the spectra, inspired from what happens in finite dimensions, for the usual complex matrices, and which shows that things do not necessarily extend without troubles to the infinite dimensional setting:

THEOREM 3.6. *We have the following formula, valid for any operators S, T :*

$$\sigma(ST) \cup \{0\} = \sigma(TS) \cup \{0\}$$

In finite dimensions we have $\sigma(ST) = \sigma(TS)$, but this fails in infinite dimensions.

PROOF. There are several statements here, the idea being as follows:

(1) This is something that we know in finite dimensions, coming from the fact that the characteristic polynomials of the associated matrices A, B coincide:

$$P_{AB} = P_{BA}$$

Thus we obtain $\sigma(ST) = \sigma(TS)$ in this case, as claimed. Observe that this improves twice the general formula in the statement, first because we have no issues at 0, and second because what we obtain is actually an equality of sets, with multiplicities.

(2) What happens in finite dimensions is very special, and as claimed in the last assertion, when H is infinite dimensional the sets $\sigma(ST), \sigma(TS)$ do not coincide in general on 0. Indeed, in order to construct a counterexample, we can assume in practice that we are dealing with the separable Hilbert space $H = l^2(\mathbb{N})$. Let us take S to be the shift:

$$S(e_i) = e_{i+1}$$

As for T , we can take the adjoint of S , which is the following operator:

$$S^*(e_i) = \begin{cases} e_{i-1} & \text{if } i > 0 \\ 0 & \text{if } i = 0 \end{cases}$$

Let us compose now these two operators. In one sense, we have:

$$S^*S = 1 \implies 0 \notin \sigma(SS^*)$$

In the other sense, however, the situation is different, as follows:

$$SS^* = Proj(e_0^\perp) \implies 0 \in \sigma(SS^*)$$

Thus, the spectra do not match on 0, and we have our counterexample, as desired.

(3) It remains to prove the first, and main assertion, stating that $\sigma(ST), \sigma(TS)$ coincide outside 0. We first prove that we have the following implication:

$$1 \notin \sigma(ST) \implies 1 \notin \sigma(TS)$$

Assume indeed that $1 - ST$ is invertible, with inverse denoted R :

$$R = (1 - ST)^{-1}$$

We have then the following formulae, relating our variables R, S, T :

$$RST = STR = R - 1$$

By using $RST = R - 1$, we have the following computation:

$$\begin{aligned} (1 + TRS)(1 - TS) &= 1 + TRS - TS - TRSTS \\ &= 1 + TRS - TS - TRS + TS \\ &= 1 \end{aligned}$$

A similar computation, using $STR = R - 1$, shows that we have:

$$(1 - TS)(1 + TRS) = 1$$

Thus $1 - TS$ is invertible, with inverse $1 + TRS$, which proves our claim. Now by multiplying by scalars, we deduce from this that for any $\lambda \in \mathbb{C} - \{0\}$ we have:

$$\lambda \notin \sigma(ST) \implies \lambda \notin \sigma(TS)$$

But this leads to the conclusion in the statement. □

3b. Spectral radius

Let us develop now some systematic theory for the computation of the spectra, based on what we know about the eigenvalues of the usual complex matrices.

As a first result, which is well-known for the usual matrices, and has no special issues in the infinite dimensional case, we have:

THEOREM 3.7. *We have the “polynomial functional calculus” formula*

$$\sigma(P(T)) = P(\sigma(T))$$

valid for any polynomial $P \in \mathbb{C}[X]$, and any operator $T \in B(H)$.

PROOF. We pick a scalar $\lambda \in \mathbb{C}$, and we decompose the polynomial $P - \lambda$:

$$P(X) - \lambda = c(X - r_1) \dots (X - r_n)$$

We have then the following equivalences:

$$\begin{aligned} \lambda \notin \sigma(P(T)) &\iff P(T) - \lambda \in B(H)^{-1} \\ &\iff c(T - r_1) \dots (T - r_n) \in B(H)^{-1} \\ &\iff T - r_1, \dots, T - r_n \in B(H)^{-1} \\ &\iff r_1, \dots, r_n \notin \sigma(T) \\ &\iff \lambda \notin P(\sigma(T)) \end{aligned}$$

Thus, we are led to the formula in the statement. □

The above result is something very useful, and extending it will be our next task. As a first ingredient here, assuming that $A \in M_N(\mathbb{C})$ is invertible, we have:

$$\sigma(A^{-1}) = \sigma(A)^{-1}$$

It is possible to extend this formula to the arbitrary operators, and we will do this in a moment. Before starting, however, we have to think in advance on how to unify this potential result, that we have in mind, with Theorem 3.7 itself.

What we have to do here is to find a class of functions generalizing both the polynomials $P \in \mathbb{C}[X]$ and the inverse function $x \rightarrow x^{-1}$, and the answer to this question is provided by the rational functions, which are as follows:

DEFINITION 3.8. *A rational function $f \in \mathbb{C}(X)$ is a quotient of polynomials:*

$$f = \frac{P}{Q}$$

Assuming that P, Q are prime to each other, we can regard f as a usual function,

$$f : \mathbb{C} - X \rightarrow \mathbb{C}$$

with X being the set of zeros of Q , also called poles of f .

Now that we have our class of functions, the next step consists in applying them to operators. Here we cannot expect $f(T)$ to make sense for any f and any T , for instance because T^{-1} is defined only when T is invertible. We are led in this way to:

DEFINITION 3.9. *Given an operator $T \in B(H)$, and a rational function $f = P/Q$ having poles outside $\sigma(T)$, we can construct the following operator,*

$$f(T) = P(T)Q(T)^{-1}$$

that we can denote as a usual fraction, as follows,

$$f(T) = \frac{P(T)}{Q(T)}$$

due to the fact that $P(T), Q(T)$ commute, so that the order is irrelevant.

Summarizing, we have so far a good class of functions, generalizing both polynomials and the inverse map $x \rightarrow x^{-1}$, and we know as well how these functions can be applied to the operators. We can now extend Theorem 3.7, as follows:

THEOREM 3.10. *We have the “rational functional calculus” formula*

$$\sigma(f(T)) = f(\sigma(T))$$

valid for any rational function $f \in \mathbb{C}(X)$ having poles outside $\sigma(T)$.

PROOF. We pick a scalar $\lambda \in \mathbb{C}$, we write $f = P/Q$, and we set:

$$F = P - \lambda Q$$

By using now Theorem 3.7, for this polynomial, we obtain:

$$\begin{aligned} \lambda \in \sigma(f(T)) &\iff F(T) \notin B(H)^{-1} \\ &\iff 0 \in \sigma(F(T)) \\ &\iff 0 \in F(\sigma(T)) \\ &\iff \exists \mu \in \sigma(T), F(\mu) = 0 \\ &\iff \lambda \in f(\sigma(T)) \end{aligned}$$

Thus, we are led to the formula in the statement. \square

As an application of the above methods, we can investigate certain special classes of operators, such as the unitary ones, and the self-adjoint ones. We first have:

THEOREM 3.11. *The spectrum of a unitary operator*

$$U^* = U^{-1}$$

is on the unit circle, $\sigma(U) \subset \mathbb{T}$.

PROOF. Assuming $U^* = U^{-1}$, we have the following norm computation:

$$\|U\| = \sqrt{\|UU^*\|} = \sqrt{1} = 1$$

Now if we denote by D the unit disk, we obtain from this:

$$\sigma(U) \subset D$$

On the other hand, once again by using $U^* = U^{-1}$, we have as well:

$$\|U^{-1}\| = \|U^*\| = \|U\| = 1$$

Thus, as before with D being the unit disk in the complex plane, we have:

$$\sigma(U^{-1}) \subset D$$

Now by using the rational function $f(z) = z^{-1}$, we obtain from this:

$$\sigma(U) \subset D^{-1}$$

Now by putting everything together, we obtain:

$$\sigma(U) \subset D \cap D^{-1} = \mathbb{T}$$

Thus, we are led to the conclusion in the statement. \square

Before getting into the study of self-adjoint elements, let us record the following elementary result, regarding the behaviour of the spectrum when taking adjoints:

PROPOSITION 3.12. *We have the following formula*

$$\sigma(T^*) = \overline{\sigma(T)}$$

valid for any operator $T \in B(H)$.

PROOF. By using the fact that an operator S is invertible precisely when its adjoint S^* is, the spectrum of the adjoint operator T^* can be computed as follows:

$$\begin{aligned} \sigma(T^*) &= \left\{ \lambda \in \mathbb{C} \mid T^* - \lambda \notin B(H)^{-1} \right\} \\ &= \left\{ \lambda \in \mathbb{C} \mid T - \bar{\lambda} \notin B(H)^{-1} \right\} \\ &= \overline{\sigma(T)} \end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

Let us discuss now the case of the self-adjoint elements. The result here, which is quite similar to the one for the unitary operators, is as follows:

THEOREM 3.13. *The spectrum of a self-adjoint operator*

$$T = T^*$$

consists of real numbers, $\sigma(T) \subset \mathbb{R}$.

PROOF. As a first observation, this cannot follow from Proposition 3.12 a bit as in the proof of Theorem 3.11, because Proposition 3.12 only gives, in the case $T = T^*$:

$$\sigma(T) = \overline{\sigma(T)}$$

However, we can deduce the result from Theorem 3.11, by using the following remarkable rational function, depending on a parameter $r \in \mathbb{R}$:

$$f(z) = \frac{z + ir}{z - ir}$$

Indeed, for $r \gg 0$ the operator $f(T)$ is well-defined, and we have:

$$\begin{aligned} \left(\frac{T + ir}{T - ir} \right)^* &= \frac{(T + ir)^*}{(T - ir)^*} \\ &= \frac{T - ir}{T + ir} \\ &= \left(\frac{T + ir}{T - ir} \right)^{-1} \end{aligned}$$

Thus $f(T)$ is unitary, and by using Theorem 3.11 we obtain:

$$\begin{aligned}\sigma(T) &\subset f^{-1}(f(\sigma(T))) \\ &= f^{-1}(\sigma(f(T))) \\ &\subset f^{-1}(\mathbb{T}) \\ &= \mathbb{R}\end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

As a theoretical remark, it is possible to deduce as well Theorem 3.11 from Theorem 3.13, by performing the above computation in the other sense. Indeed, by assuming that Theorem 3.13 holds indeed, and starting with a unitary $U \in B(H)$, we obtain:

$$\begin{aligned}\sigma(U) &\subset f(f^{-1}(\sigma(U))) \\ &= f(\sigma(f^{-1}(U))) \\ &\subset f(\mathbb{R}) \\ &= \mathbb{T}\end{aligned}$$

However, proving Theorem 3.13 directly, with bare hands, is something not obvious, and the above order of presentation, with Theorem 3.11 coming first, is the good one.

As a conclusion, we have so far a beginning of spectral theory, with results allowing us to investigate the unitaries and the self-adjoints, and with the remark that these two classes of operators are related by a certain wizarding rational function, namely:

$$f(z) = \frac{z + ir}{z - ir}$$

Let us keep building on this, with more complex analysis involved. One key thing that we know about matrices, and which follows for instance by using the fact that the diagonalizable matrices are dense, is the following formula:

$$\sigma(e^A) = e^{\sigma(A)}$$

This is something heavily used across mathematics, as for instance in the theory of ordinary differential equations. We would like of course to have such formulae for the general operators $T \in B(H)$ too, and this is indeed possible, as follows:

THEOREM 3.14. *We have the “holomorphic functional calculus” formula*

$$\sigma(f(T)) = f(\sigma(T))$$

valid for any holomorphic function $f \in \text{Hol}(\sigma(T))$.

PROOF. This is something that we will not really need, for the purposes of the present book, which is quite soft on analytic aspects, so here is the general idea:

(1) Consider the rational calculus morphism from Definition 3.9, which is as follows, with the exponent standing for “having poles outside $\sigma(T)$ ”:

$$\begin{aligned}\mathbb{C}(X)^T &\rightarrow B(H) \\ f &\rightarrow f(T)\end{aligned}$$

As already mentioned before, the rational functions are holomorphic outside their poles, and this raises the question of extending this morphism, as follows:

$$\begin{aligned}Hol(\sigma(T)) &\rightarrow B(H) \\ f &\rightarrow f(T)\end{aligned}$$

(2) In order to construct this extension, let us first consider the case of the entire functions, such as the exponential function $f(z) = e^z$. Here we have a Taylor series available, and we can simply apply this Taylor series to our operator, as follows:

$$e^T = \sum_{k=0}^{\infty} \frac{T^k}{k!}$$

Observe that such infinite series converge indeed, due to:

$$\|T^k\| \leq \|T\|^k$$

It is possible to further build on this, a bit as in the polynomial function case, and in the end to obtain the formula $\sigma(f(T)) = f(\sigma(T))$ as well, for the exponential function $f(z) = e^z$, and for the other entire holomorphic functions.

(3) In general, however, the holomorphic functions are not entire, and the above method won't actually cover the rational functions $f \in \mathbb{C}(X)^T$ that we want to generalize. Thus, we must use something else. And the answer here comes from the Cauchy formula:

$$f(t) = \frac{1}{2\pi i} \int_{\gamma} \frac{f(z)}{z-t} dz$$

Indeed, given a rational function $f \in \mathbb{C}(X)^T$, the corresponding operator $f(T) \in B(H)$, constructed in Definition 3.9, can be recaptured in an analytic way, as follows:

$$f(T) = \frac{1}{2\pi i} \int_{\gamma} \frac{f(z)}{z-T} dz$$

Now given an arbitrary function $f \in Hol(\sigma(T))$, we can define $f(T) \in B(H)$ by the exactly same formula, and we obtain in this way the desired correspondence:

$$\begin{aligned}Hol(\sigma(T)) &\rightarrow B(H) \\ f &\rightarrow f(T)\end{aligned}$$

(4) In practice now, all this needs a bit of care, notably with the verification of the fact that the operator $f(T) \in B(H)$ does not depend on γ , and with the technical remark that a winding number must be added to the above Cauchy formulae, for things to be

correct. But this can be done via a standard study, keeping in mind the fact that in the case $H = \mathbb{C}$, where our operators are usual numbers, $B(H) = \mathbb{C}$, what we want to do is simply proving that the usual Cauchy formula holds indeed.

(5) Now with this correspondence $f \rightarrow f(T)$ constructed, and so with the formula in the statement, namely $\sigma(f(T)) = f(\sigma(T))$, making now sense, it remains to prove that this formula holds indeed. But this follows as well via a careful use of the Cauchy formula, or by using approximation by polynomials, or rational functions. \square

As already said, the above result is important for advanced operator theory and applications, and we will not get further into this subject, the present book being more on the algebraic side of the theory. We will be back later, however, to the various functional calculus formulae established above, with some applications and improvements, in the special case of the normal operators, which is of particular interest for us.

In order to formulate our next result, we will need the following notion:

DEFINITION 3.15. *Given an operator $T \in B(H)$, its spectral radius*

$$\rho(T) \in (0, \|T\|)$$

is the radius of the smallest disk centered at 0 containing $\sigma(T)$.

Here we have included for simplicity, and for further reference in what follows, a number of results that we already know, from Theorem 3.5 above, namely the fact that the spectrum is indeed non-empty, and is contained in the disk $D_0(\|T\|)$.

With the above notion in hand, we have the following key result, improving among others our key result so far, namely $\sigma(T) \neq \emptyset$, from Theorem 3.5:

THEOREM 3.16. *The spectral radius of an operator $T \in B(H)$ is given by*

$$\rho(T) = \lim_{n \rightarrow \infty} \|T^n\|^{1/n}$$

and in this formula, we can replace the limit by an inf.

PROOF. We have to prove two inequalities, and the idea is as follows:

(1) In one sense, we use the following polynomial calculus formula:

$$\sigma(T^n) = \sigma(T)^n$$

But this gives the following estimate, valid for any n , as desired:

$$\begin{aligned}
\rho(T) &= \sup_{\lambda \in \sigma(T)} |\lambda| \\
&= \sup_{\rho \in \sigma(T)^n} |\rho|^{1/n} \\
&= \sup_{\rho \in \sigma(T^n)} |\rho|^{1/n} \\
&= \rho(T^n)^{1/n} \\
&\leq \|T^n\|^{1/n}
\end{aligned}$$

(2) For the reverse inequality, we fix a number as follows:

$$\rho > \rho(T)$$

We have then the following computation, based on the Cauchy formula:

$$\begin{aligned}
\int_{|z|=\rho} \frac{z^n}{z-T} dz &= \int_{|z|=\rho} \sum_{k=0}^{\infty} z^{n-k-1} T^k dz \\
&= \sum_{k=0}^{\infty} \left(\int_{|z|=\rho} z^{n-k-1} dz \right) T^k \\
&= \sum_{k=0}^{\infty} \delta_{n,k+1} T^k \\
&= T^{n-1}
\end{aligned}$$

By applying the norm and taking n -th roots we obtain from this formula, modulo some elementary manipulations, the following estimate:

$$\rho \geq \lim_{n \rightarrow \infty} \|T^n\|^{1/n}$$

Now recall that ρ was by definition an arbitrary number satisfying $\rho > \rho(T)$. Thus, we have obtained the following estimate, valid for any $T \in B(H)$:

$$\rho(T) \geq \lim_{n \rightarrow \infty} \|T^n\|^{1/n}$$

Thus, we are led to the conclusion in the statement. □

In the case of the normal elements, we have the following finer result:

THEOREM 3.17. *The spectral radius of a normal element,*

$$TT^* = T^*T$$

is equal to its norm.

PROOF. We can proceed in two steps, as follows:

Step 1. In the case $T = T^*$ we have $\|T^n\| = \|T\|^n$ for any exponent of the form $n = 2^k$, by using the formula $\|TT^*\| = \|T\|^2$, and by taking n -th roots we get:

$$\rho(T) \geq \|T\|$$

Thus, we are done with the self-adjoint case, with the result $\rho(T) = \|T\|$.

Step 2. In the general normal case $TT^* = T^*T$ we have $T^n(T^n)^* = (TT^*)^n$, and by using this, along with the result from Step 1, applied to TT^* , we obtain:

$$\begin{aligned} \rho(T) &\geq \lim_{n \rightarrow \infty} \|T^n\|^{1/n} \\ &= \sqrt{\lim_{n \rightarrow \infty} \|T^n(T^n)^*\|^{1/n}} \\ &= \sqrt{\lim_{n \rightarrow \infty} \|(TT^*)^n\|^{1/n}} \\ &= \sqrt{\rho(TT^*)} \\ &= \sqrt{\|T\|^2} \\ &= \|T\| \end{aligned}$$

Thus, we are led to the conclusion in the statement. □

As a first comment, the spectral radius formula $\rho(T) = \|T\|$ does not hold in general, the simplest counterexample being the following non-normal matrix:

$$A = \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix}$$

As another comment, we can combine the formula $\rho(T) = \|T\|$ for normal operators with the formula $\|TT^*\| = \|T\|^2$, and we are led to the following statement:

PROPOSITION 3.18. *The norm of $B(H)$ is given by*

$$\|T\| = \sqrt{\sup \left\{ \lambda \in \mathbb{C} \mid TT^* - \lambda \notin B(H)^{-1} \right\}}$$

and so is a purely algebraic quantity.

PROOF. We have the following computation, using the formula $\|TT^*\| = \|T\|^2$, then the spectral radius formula for TT^* , and finally the definition of the spectral radius:

$$\begin{aligned} \|T\| &= \sqrt{\|TT^*\|} \\ &= \sqrt{\rho(TT^*)} \\ &= \sqrt{\sup \left\{ \lambda \in \mathbb{C} \mid \lambda \in \sigma(TT^*) \right\}} \\ &= \sqrt{\sup \left\{ \lambda \in \mathbb{C} \mid TT^* - \lambda \notin B(H)^{-1} \right\}} \end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

The above result is quite interesting, philosophically speaking. We will be back to this, with further results and comments on $B(H)$, and other algebras of the same type.

3c. Normal operators

By using Theorem 3.17 we can say a number of non-trivial things concerning the normal operators, commonly known as “spectral theorem for normal operators”.

As a first result here, we can improve the polynomial functional calculus formula:

THEOREM 3.19. *Given $T \in B(H)$ normal, we have a unique morphism of algebras*

$$\mathbb{C}[X] \rightarrow B(H)$$

$$P \rightarrow P(T)$$

given by $X \rightarrow T$, which has the following properties:

- (1) $\sigma(P(T)) = P(\sigma(T))$.
- (2) $\|P(T)\| = \|P_{|\sigma(T)}\|$.
- (3) $Tx = \lambda x \implies P(T)x = P(\lambda)x$.
- (4) $[S, T] = 0 \implies [S, P(T)] = 0$.

PROOF. This is an improvement of Theorem 3.7 in the normal case, as follows:

(1) This is something that we know from Theorem 3.7, valid for any $T \in B(H)$.

(2) The element $P(T)$ being normal, we can apply to it the spectral radius formula for normal elements, and by using (1) we obtain, as desired:

$$\begin{aligned} \|P(T)\| &= \rho(P(T)) \\ &= \sup_{\lambda \in \sigma(P(T))} |\lambda| \\ &= \sup_{\lambda \in P(\sigma(T))} |\lambda| \\ &= \|P_{|\sigma(T)}\| \end{aligned}$$

(3) This is something clear, by linearity and multiplicativity.

(4) Once again this is clear, by linearity and multiplicativity. \square

We can improve as well the rational calculus formula, as follows:

THEOREM 3.20. *Given $T \in B(H)$ normal, we have a unique morphism of algebras as follows, with the exponent standing for “having poles outside $\sigma(T)$ ”,*

$$\mathbb{C}(X)^T \rightarrow B(H)$$

$$f \rightarrow f(T)$$

given by $X \rightarrow T$, which has the following properties:

- (1) $\sigma(f(T)) = f(\sigma(T))$.
- (2) $\|f(T)\| = \|f|_{\sigma(T)}\|$.
- (3) $Tx = \lambda x \implies f(T)x = f(\lambda)x$.
- (4) $[S, T] = 0 \implies [S, f(T)] = 0$.

PROOF. This is an improvement of Theorem 3.10 in the normal case, as follows:

(1) This is something that we know from Theorem 3.10, valid for any $T \in B(H)$.

(2) This follows from (1), via the same computation as for polynomials.

(3) This is something clear, by linearity and multiplicativity.

(4) Once again this is clear, by linearity and multiplicativity. \square

It is possible to improve as well the holomorphic calculus formula, as follows:

THEOREM 3.21. *Given $T \in B(H)$ normal, we have a unique morphism of algebras*

$$\text{Hol}(\sigma(T)) \rightarrow B(H)$$

$$f \rightarrow f(T)$$

given by $X \rightarrow T$, which has the following properties:

- (1) $\sigma(f(T)) = f(\sigma(T))$.
- (2) $\|f(T)\| = \|f\|$.
- (3) $Tx = \lambda x \implies f(T)x = f(\lambda)x$.
- (4) $[S, T] = 0 \implies [S, f(T)] = 0$.

PROOF. This is an improvement of Theorem 3.14 in the normal case, as follows:

(1) This is something that we know from Theorem 3.14, valid for any $T \in B(H)$.

(2) This follows from (1), via the same computation as for polynomials.

(3) This is something clear, by linearity and multiplicativity.

(4) Once again this is clear, by linearity and multiplicativity. \square

Importantly now, in the case of normal elements we have some new functional calculus results, using more general functions than those used before. First, we have:

THEOREM 3.22. *Given $T \in B(H)$ normal, we have a unique continuous morphism of algebras*

$$\begin{aligned} C(\sigma(T)) &\rightarrow B(H) \\ f &\rightarrow f(T) \end{aligned}$$

given by $z \rightarrow T$, which has the following properties:

- (1) $\|f(T)\| = \|f\|$.
- (2) $\sigma(f(T)) = f(\sigma(T))$.
- (3) $Tx = \lambda x \implies f(T)x = f(\lambda)x$.
- (4) $[S, T] = 0 \implies [S, f(T)] = 0$.

PROOF. The idea here is to “complete” the morphisms in Theorems 3.19 or 3.20 or 3.21. Consider indeed the morphism constructed in Theorem 3.19, namely:

$$\begin{aligned} \mathbb{C}[X] &\rightarrow B(H) \\ P &\rightarrow P(T) \end{aligned}$$

We know from Theorem 3.19 that this morphism is continuous, and in fact isometric, when regarding the polynomials $P \in \mathbb{C}[X]$ as functions on $\sigma(T)$:

$$\|P(T)\| = \|P_{\sigma(T)}\|$$

Thus, by Stone-Weierstrass, we have a unique continuous extension of this morphism, as in the statement, namely:

$$\begin{aligned} C(\sigma(T)) &\rightarrow B(H) \\ f &\rightarrow f(T) \end{aligned}$$

Regarding now the assertions (1-4), the proof here goes as follows:

(1) This follows from our construction of $f \rightarrow f(T)$, via Stone-Weierstrass.

(2) This is something that we know for polynomials and rational functions, that we have now to extend to the continuous functions. We proceed by double inclusion:

“ \subset ” Given a continuous function $f \in C(\sigma(T))$, we must prove that we have:

$$\lambda \notin f(\sigma(T)) \implies \lambda \notin \sigma(f(T))$$

For this purpose, consider the following function, which is well-defined:

$$\frac{1}{f - \lambda} \in C(\sigma(T))$$

We can therefore apply this function to T , and we obtain:

$$\left(\frac{1}{f - \lambda} \right) T = \frac{1}{f(T) - \lambda}$$

In particular $f(T) - \lambda$ is invertible, so $\lambda \notin \sigma(f(T))$, as desired.

“ \supset ” Given a continuous function $f \in C(\sigma(T))$, we must prove that we have:

$$\lambda \in f(\sigma(T)) \implies \lambda \in \sigma(f(T))$$

But this is the same as proving that we have:

$$\mu \in \sigma(T) \implies f(\mu) \in \sigma(f(T))$$

For this purpose, we approximate our function by polynomials, $P_n \rightarrow f$, and we examine the following convergence, which follows from $P_n \rightarrow f$:

$$P_n(T) - P_n(\mu) \rightarrow f(T) - f(\mu)$$

We know from polynomial functional calculus that we have:

$$P_n(\mu) \in P_n(\sigma(T)) = \sigma(P_n(T))$$

Thus, the operators $P_n(T) - P_n(\mu)$ are not invertible. On the other hand, we know that the set formed by the invertible operators is open, so its complement is closed. Thus the limit $f(T) - f(\mu)$ is not invertible either, and so $f(\mu) \in \sigma(f(T))$, as desired.

(3) This is something clear, by linearity and multiplicativity.

(4) Once again this is clear, by linearity and multiplicativity. \square

As an important comment, Theorem 3.22 is not exactly in final form, because it misses an important point, namely that our correspondence maps:

$$\bar{z} \rightarrow T^*$$

However, this is something non-trivial, and we will be back to this later. Observe however that Theorem 3.22 is fully powerful for the self-adjoint operators, $T = T^*$, where the spectrum is real, and so where $z = \bar{z}$ on the spectrum. We will be back to this.

As a second result now, along the same lines, we can further extend Theorem 3.22 into a measurable functional calculus theorem, as follows:

THEOREM 3.23. *Given $T \in B(H)$ normal, we have a unique continuous morphism of algebras as follows, with L^∞ standing for abstract measurable functions*

$$L^\infty(\sigma(T)) \rightarrow B(H)$$

$$f \rightarrow f(T)$$

given by $z \rightarrow T$, which has the following properties:

- (1) $\|f(T)\| = \|f\|$.
- (2) $\sigma(f(T)) = f(\sigma(T))$.
- (3) $Tx = \lambda x \implies f(T)x = f(\lambda)x$.
- (4) $[S, T] = 0 \implies [S, f(T)] = 0$.

PROOF. As before, the idea is that of “completing” what we already have. For this purpose, we will use a polarization trick. Given $x \in H$, consider the following functional:

$$C(\sigma(T)) \rightarrow \mathbb{C} \quad , \quad f \rightarrow \langle f(T)x, x \rangle$$

By the Riesz theorem, this functional must be the integration with respect to a certain measure μ on the space $\sigma(T)$. Thus, we have a formula as follows:

$$\langle f(T)x, x \rangle = \int_{\sigma(T)} f(z) d\mu(z)$$

Now with this formula in hand, we can extend the continuous functional calculus into an abstract measurable one, by exactly the same formula, as follows:

$$\langle f(T)x, x \rangle = \int_{\sigma(T)} f(z) d\mu(z)$$

We have then a morphism of algebras, as desired, and the properties (1-4) in the statement hold too, with the extensions being as before, by using this time in the proofs approximations by continuous functions, and the polarization identity. \square

The same comments as before apply. Theorem 3.23 is not exactly in final form, because it misses an important point, namely that our correspondence maps:

$$\bar{z} \rightarrow T^*$$

However, this is something non-trivial, and we will be back to this later. Observe however that Theorem 3.23 is fully powerful for the self-adjoint operators, $T = T^*$, where the spectrum is real, and so where $z = \bar{z}$ on the spectrum. We will be back to this.

3d. Diagonalization

We can now diagonalize the normal operators. We will do this in 3 steps, first for the self-adjoint operators, then for the families of commuting self-adjoint operators, and finally for the general normal operators, by using a $T = \operatorname{Re}(T) + i\operatorname{Im}(T)$ trick.

The diagonalization in infinite dimensions is more tricky than in finite dimensions, and instead of writing a formula of type $T = UDU^*$, with $U, D \in B(H)$ being unitary and diagonal, we will express our operator as $T = U^*MU$, with $U : H \rightarrow K$ being a certain unitary, and $M \in B(K)$ being a certain diagonal operator. This is how the spectral theorem is best formulated, in view of applications. In practice, the explicit construction of U, M , which will be actually rather part of the proof, is also needed.

For the self-adjoint operators, the statement and proof are as follows:

THEOREM 3.24. *Any self-adjoint operator $T \in B(H)$ can be diagonalized,*

$$T = U^* M_f U$$

with $U : H \rightarrow L^2(X)$ being a unitary operator from H to a certain L^2 space associated to T , with $f : X \rightarrow \mathbb{R}$ being a certain function, once again associated to T , and with

$$M_f(g) = fg$$

being the usual multiplication operator by f , on the Hilbert space $L^2(X)$.

PROOF. The construction of U, f can be done in several steps, as follows:

(1) We first prove the result in the special case where our operator T has a cyclic vector $x \in H$, with this meaning that the following holds:

$$\overline{\text{span} \left(T^k x \mid n \in \mathbb{N} \right)} = H$$

For this purpose, let us go back to the proof of Theorem 3.23. We will use the following formula from there, with μ being the measure on $X = \sigma(T)$ associated to x :

$$\langle g(T)x, x \rangle = \int_{\sigma(T)} g(z) d\mu(z)$$

Our claim is that we can define a unitary $U : H \rightarrow L^2(X)$, first on the dense part spanned by the vectors $T^k x$, by the following formula, and then by continuity:

$$U[g(T)x] = g$$

Indeed, the following computation shows that U is well-defined, and isometric:

$$\begin{aligned} \|g(T)x\|^2 &= \langle g(T)x, g(T)x \rangle \\ &= \langle g(T)^* g(T)x, x \rangle \\ &= \langle |g|^2(T)x, x \rangle \\ &= \int_{\sigma(T)} |g(z)|^2 d\mu(z) \\ &= \|g\|_2^2 \end{aligned}$$

We can then extend U by continuity into a unitary $U : H \rightarrow L^2(X)$, as claimed. Now observe that we have the following formula:

$$\begin{aligned} UTU^*g &= U[Tg(T)x] \\ &= U[(zg)(T)x] \\ &= zg \end{aligned}$$

Thus our result is proved in the present case, with U as above, and with $f(z) = z$.

(2) We discuss now the general case. Our first claim is that H has a decomposition as follows, with each H_i being invariant under T , and admitting a cyclic vector x_i :

$$H = \bigoplus_i H_i$$

Indeed, this is something elementary, the construction being by recurrence in finite dimensions, in the obvious way, and by using the Zorn lemma in general.

Now with this decomposition result in hand, we can make a direct sum of the diagonalizations obtained in (1) above, for each of the restrictions $T|_{H_i}$, and we obtain the formula in the statement. \square

We have the following technical generalization of the above result:

THEOREM 3.25. *Any family of commuting self-adjoint operators $T_i \in B(H)$ can be jointly diagonalized,*

$$T_i = U^* M_{f_i} U$$

with $U : H \rightarrow L^2(X)$ being a unitary operator from H to a certain L^2 space associated to $\{T_i\}$, with $f_i : X \rightarrow \mathbb{R}$ being certain functions, once again associated to T_i , and with

$$M_{f_i}(g) = f_i g$$

being the usual multiplication operator by f_i , on the Hilbert space $L^2(X)$.

PROOF. This is similar to the proof of Theorem 3.24, by suitably modifying the measurable calculus formula, and the measure μ itself, as to have this formula working for all the operators T_i . With this modification done, everything extends. \square

We can now discuss the case of arbitrary normal operators, as follows:

THEOREM 3.26. *Any normal operator $T \in B(H)$ can be diagonalized,*

$$T = U^* M_f U$$

with $U : H \rightarrow L^2(X)$ being a unitary operator from H to a certain L^2 space associated to T , with $f : X \rightarrow \mathbb{C}$ being a certain function, once again associated to T , and with

$$M_f(g) = fg$$

being the usual multiplication operator by f , on the Hilbert space $L^2(X)$.

PROOF. This follows by applying Theorem 3.25 above to the real and imaginary parts of T , which are constructed as follows, and which are self-adjoint, and commuting:

$$T = \frac{T + T^*}{2} - i \cdot \frac{i(T - T^*)}{2}$$

Alternatively, we can use methods similar to those that we used in chapter 1 above, in order to deal with the usual normal matrices, involving the special relation between T and the operator TT^* , which is self-adjoint. We will be back to this. \square

This was for our series of diagonalization theorems. There is of course one more result here, regarding the families of commuting normal operators, as follows:

THEOREM 3.27. *Any family of commuting normal operators $T_i \in B(H)$ can be jointly diagonalized,*

$$T_i = U^* M_{f_i} U$$

with $U : H \rightarrow L^2(X)$ being a unitary operator from H to a certain L^2 space associated to $\{T_i\}$, with $f_i : X \rightarrow \mathbb{C}$ being certain functions, once again associated to T_i , and with

$$M_{f_i}(g) = f_i g$$

being the usual multiplication operator by f_i , on the Hilbert space $L^2(X)$.

PROOF. This is similar to the proof of Theorem 3.25 and Theorem 3.26, by combining the arguments there. To be more precise, this follows as Theorem 3.25, by using the decomposition trick from the proof of Theorem 3.26. \square

With the above diagonalization results in hand, we can now “fix” the continuous and measurable functional calculus theorems, with a key complement, as follows:

THEOREM 3.28. *Given a normal operator $T \in B(H)$, the following hold, for both the functional calculus and the measurable calculus morphisms:*

- (1) *These morphisms are $*$ -morphisms.*
- (2) *The function \bar{z} gets mapped to T^* .*
- (3) *The functions $\operatorname{Re}(z)$, $\operatorname{Im}(z)$ get mapped to $\operatorname{Re}(T)$, $\operatorname{Im}(T)$.*
- (4) *The function $|z|^2$ gets mapped to $TT^* = T^*T$.*
- (5) *If f is real, then $f(T)$ is self-adjoint.*

PROOF. These assertions are more or less equivalent, with (1) being the main one, which obviously implies everything else. But this assertion (1) follows from the diagonalization result for normal operators, from Theorem 3.26 above. \square

This was for the spectral theory of arbitrary and normal operators. As a conclusion here, our main results are as follows:

- (1) Regarding the arbitrary operators, the main results here, or rather the most advanced results, are the holomorphic calculus formula from Theorem 3.14, and the spectral radius estimate from Theorem 3.16.
- (2) For the self-adjoint operators, the main results are the spectral radius formula from Theorem 3.17, the measurable calculus formula from Theorem 3.23, and the diagonalization result from Theorem 3.24.
- (3) For general normal operators, the main results are the spectral radius formula from Theorem 3.17, the measurable calculus formula from Theorem 3.23, complemented by Theorem 3.28, and the diagonalization result in Theorem 3.26.

There are of course many other things that can be said about the spectral theory of the bounded operators $T \in B(H)$, all the above being just an introduction to the subject. As a complement, we recommend any good operator theory book.

3e. Exercises

The main theoretical notion introduced in this chapter was that of the spectrum of an operator, and as a first exercise here, we have:

EXERCISE 3.29. *Prove that for the usual matrices $A, B \in M_N(\mathbb{C})$ we have*

$$\sigma^+(AB) = \sigma^+(BA)$$

where σ^+ denotes the set of eigenvalues, taken with multiplicities.

As a remark, we have seen in the above that $\sigma(AB) = \sigma(BA)$ holds outside $\{0\}$, and the equality on $\{0\}$ holds as well, because AB is invertible if and only if BA is invertible. However, in what regards the eigenvalues taken with multiplicities, things are more tricky here, and the answer should be somewhere inside your linear algebra knowledge.

At a more theoretical level now, also in connection with the spectrum, we have:

EXERCISE 3.30. *Clarify, with examples and counterexamples, the relation between the eigenvalues of an operator $T \in B(H)$, and its spectrum $\sigma(T) \subset \mathbb{C}$.*

Here, as usual, the counterexamples could only come from the shift operator S , on the space $H = l^2(\mathbb{N})$. As a bonus exercise here, try computing the spectrum of S .

In relation now with functional calculus, we have:

EXERCISE 3.31. *Prove, with full details, that the functional calculus formula*

$$\sigma(f(T)) = f(\sigma(T))$$

made indeed sense for any holomorphic function $f \in \text{Hol}(\sigma(T))$.

This is something that we already discussed in the above, with the comment that the proof is quite technical, involving countless applications of the Cauchy formula, and that this will not be really needed, in what follows. However, this result is very important when doing more advanced operator theory, and time now to have it understood.

As a more concrete exercise now, in relation with spectral calculus, we have:

EXERCISE 3.32. *Draw the picture of the following rational function, and of its inverse,*

$$f(z) = \frac{z + ir}{z - ir}$$

with $r \in \mathbb{R}$, and prove that for $r \gg 0$ and $T = T^$, the element $f(T)$ is well-defined.*

This is something that we actually used in the above, when computing spectra of self-adjoints and unitaries, and the problem is that of working out all the details.

In relation now with the spectral radius theorem, we have:

EXERCISE 3.33. *Comment on the spectral radius theorem, stating that for a normal operator, $TT^* = T^*T$, the spectral radius is equal to the norm,*

$$\rho(T) = \|T\|$$

with examples and counterexamples, and simpler proofs of well, in various particular cases of interest, such as the finite dimensional one.

This is of course something a bit philosophical, but the spectral radius theorem being our key technical result so far, some further thinking of it is definitely a good thing.

On the same topic, and even more philosophically now, we have:

EXERCISE 3.34. *Develop a theory of $*$ -algebras A for which the quantity*

$$\|a\| = \sqrt{\sup \left\{ \lambda \in \mathbb{C} \mid aa^* - \lambda \notin A^{-1} \right\}}$$

defines a norm, for the elements $a \in A$.

As pointed out in the above, the spectral radius formula shows that for the full operator algebra $A = B(H)$ the norm is given by the above formula, and so there should be such a theory of “good” $*$ -algebras, with $A = B(H)$ as a main example. So, try to have this done, and do not depress or something if you get lost. Math is sometimes mysterious.

In relation now with the spectral theorem for the normal operators, which is quite complicated business, we have, as a first exercise:

EXERCISE 3.35. *Find and write down a proof for the spectral theorem for normal operators in the spirit of the proof for normal matrices from chapter 1, and vice versa.*

To be more precise, the problem is that the proof of the spectral theorem for the usual matrices, from chapter 1, was using a certain kind of trick, while the proof of the spectral theorem for the arbitrary operators, given in this chapter, was using some other kind of trick. Thus, for full understanding all this, working out more proofs, both for the usual matrices and for the arbitrary operators, is a useful thing.

Finally, also in connection with the spectral theorem, we have:

EXERCISE 3.36. *Find and write down an enhancement of the proof given above for the spectral theorem, as for $\bar{z} \rightarrow T^*$ to appear way before the end of the proof.*

This is something a bit philosophical, and check here first the various comments made above, and maybe work out this as well in parallel with the previous exercise.

CHAPTER 4

Compact operators

4a. Polar decomposition

We have seen so far the basic theory of bounded operators, in the arbitrary, normal and self-adjoint cases, and in a few other cases of interest. In this chapter we discuss a number of more specialized questions, for the most dealing with the compact operators, which are particularly close, conceptually speaking, to the usual complex matrices.

Before getting into this subject, which is quite technical, and as a first application of the methods that we have, we can now develop the theory of positive operators, and then establish polar decomposition results for the operators $T \in B(H)$. We will need:

PROPOSITION 4.1. *Given a bounded operator $T \in B(H)$, the following happen:*

- (1) $\ker T^* = (ImT)^\perp$.
- (2) $\overline{ImT^*} = (\ker T)^\perp$.

PROOF. Both these assertions are elementary, as follows:

- (1) Let us first prove “ \subset ”. Assuming $T^*x = 0$, we have indeed $x \perp ImT$, because:

$$\langle x, Ty \rangle = \langle T^*x, y \rangle = 0$$

As for “ \supset ”, assuming $\langle x, Ty \rangle = 0$ for any y , we have $T^*x = 0$, because:

$$\langle T^*x, y \rangle = \langle x, Ty \rangle = 0$$

- (2) This can be proved by some similar computations, or can be simply deduced from the formula (1), applied to the operator T^* , as follows:

$$(\ker T)^\perp = (ImT^*)^{\perp\perp} = \overline{ImT^*}$$

Here we have used the following formula, valid for any linear subspace $K \subset H$ of a Hilbert space, which is well-known, and standard, coming from Hahn-Banach:

$$K^{\perp\perp} = \bar{K}$$

Thus, we are led to the conclusions in the statement. \square

We can now develop the theory of positive operators $T \in B(H)$, in analogy with the theory of the positive matrices $A \in M_N(\mathbb{C})$. We have already seen some theory in chapter 2 above, that we can now complete, into a final statement on the subject, as follows:

THEOREM 4.2. For an operator $T \in B(H)$, the following are equivalent:

- (1) T is positive, in the sense that $\langle Tx, x \rangle \geq 0$.
- (2) T is normal, and $\sigma(T) \subset [0, \infty)$.
- (3) $T = S^2$, for some $S \in B(H)$ satisfying $S = S^*$.
- (4) $T = R^*R$, for some $R \in B(H)$.

PROOF. This is something quite tricky, as follows:

- (1) \implies (2) Assuming $\langle Tx, x \rangle \geq 0$, with $S = T - T^*$ we have:

$$\begin{aligned} \langle Sx, x \rangle &= \langle Tx, x \rangle - \langle T^*x, x \rangle \\ &= \langle Tx, x \rangle - \langle x, Tx \rangle \\ &= \langle Tx, x \rangle - \overline{\langle Tx, x \rangle} \\ &= 0 \end{aligned}$$

By using the polarization identity we obtain that we have $\|Sx\| = 0$, and so:

$$S = 0$$

Thus T is self-adjoint, and so normal, and also, by self-adjointness, we have:

$$\sigma(T) \subset \mathbb{R}$$

In order to prove now that we have indeed $\sigma(T) \subset [0, \infty)$, we must invert $T + \lambda$, for any $\lambda > 0$. For this purpose, observe that we have:

$$\begin{aligned} \langle (T + \lambda)x, x \rangle &= \langle Tx, x \rangle + \langle \lambda x, x \rangle \\ &\geq \langle \lambda x, x \rangle \\ &= \lambda \|x\|^2 \end{aligned}$$

But this shows that $T + \lambda$ is injective. In order to prove now the surjectivity, and the boundedness of the inverse, observe first that we have, by using Proposition 4.1:

$$\begin{aligned} \text{Im}(T + \lambda)^\perp &= \ker(T + \lambda)^* \\ &= \ker(T + \lambda) \\ &= \{0\} \end{aligned}$$

Thus $\text{Im}(T + \lambda)$ is dense. On the other hand, observe that we have:

$$\begin{aligned} \|(T + \lambda)x\|^2 &= \langle Tx + \lambda x, Tx + \lambda x \rangle \\ &= \|Tx\|^2 + 2\lambda \langle Tx, x \rangle + \lambda^2 \|x\|^2 \\ &\geq \lambda^2 \|x\|^2 \end{aligned}$$

Thus for any vector in the image $y \in \text{Im}(T + \lambda)$ we have:

$$\|y\| \geq \lambda \|(T + \lambda)^{-1}y\|$$

As a conclusion to what we have so far, $T + \lambda$ is bijective and invertible as a bounded operator from H onto its image, with the following norm bound:

$$\|(T + \lambda)^{-1}\| \leq \lambda^{-1}$$

But this shows that $Im(T + \lambda)$ is complete, hence closed, and since we already knew that $Im(T + \lambda)$ is dense, our operator $T + \lambda$ is surjective, and we are done.

(2) \implies (3) Since T is normal, and with spectrum contained in $[0, \infty)$, we can use the continuous functional calculus formula for the normal operators, from chapter 3 above, with the function $f(x) = \sqrt{x}$, as to construct a square root $S = \sqrt{T}$.

(3) \implies (4) This is trivial, because we can set $R = S$.

(4) \implies (1) This is clear, because we have the following computation:

$$\langle R^*Rx, x \rangle = \langle Rx, Rx \rangle = \|Rx\|^2$$

Thus, we have the equivalences in the statement. □

Let us record as well a “strict” version of the above result, as follows:

THEOREM 4.3. *For an operator $T \in B(H)$, the following are equivalent:*

- (1) T is positive and invertible.
- (2) T is normal, and $\sigma(T) \subset (0, \infty)$.
- (3) $T = S^2$, for some $S \in B(H)$ invertible, satisfying $S = S^*$.
- (4) $T = R^*R$, for some $R \in B(H)$ invertible.

If these conditions hold, then $\langle Tx, x \rangle > 0$ for any $x \neq 0$, but the converse is not true.

PROOF. Our claim is that the above conditions (1-4) are precisely the conditions (1-4) in Theorem 4.2, with the assumption “ T is invertible” added. Indeed:

(1) This is clear by definition.

(2) In the context of Theorem 4.2 (2), namely when T is normal, and $\sigma(T) \subset [0, \infty)$, the invertibility of T , which means $0 \notin \sigma(T)$, gives $\sigma(T) \subset (0, \infty)$, as desired.

(3) In the context of Theorem 4.2 (3), namely when $T = S^2$, with $S = S^*$, by using the basic properties of the functional calculus for normal operators, the invertibility of T is equivalent to the invertibility of its square root $S = \sqrt{T}$, as desired.

(4) In the context of Theorem 4.2 (4), namely when $T = RR^*$, the invertibility of T is equivalent to the invertibility of R . This can be either checked directly, or deduced via the equivalence (3) \iff (4) from Theorem 4.2, by using the above argument (3).

Thus, by using now Theorem 4.2, we are led to equivalences in the statement. Regarding now the last assertion, the inequality there can be deduced as follows, by using

the fact that the operator $S = \sqrt{T}$ is invertible, and in particular injective:

$$\begin{aligned} \langle Tx, x \rangle &= \langle S^2x, x \rangle \\ &= \langle Sx, S^*x \rangle \\ &= \langle Sx, Sx \rangle \\ &= \|Sx\|^2 \\ &> 0 \end{aligned}$$

Finally, regarding the converse, consider the following operator on $l^2(\mathbb{N})$:

$$T = \begin{pmatrix} 1 & & & \\ & \frac{1}{2} & & \\ & & \frac{1}{3} & \\ & & & \ddots \end{pmatrix}$$

This operator T is well-defined and bounded, and we have $\langle Tx, x \rangle > 0$ for any $x \neq 0$. However T is not invertible, and so the converse does not hold, as stated. \square

With the above positivity results in hand, let us discuss now some basic decomposition results for the bounded operators $T \in B(H)$, in analogy with what we know about the usual complex numbers $z \in \mathbb{C}$. From an additive viewpoint, we know that any complex number $z \in \mathbb{C}$ can be written as follows, with $a, b \in \mathbb{R}$:

$$z = a + ib$$

Also, we know that both the real and imaginary parts $a, b \in \mathbb{R}$, and more generally any real number $c \in \mathbb{R}$, can be written as follows, with $p, q \geq 0$:

$$c = p - q$$

Here is the operator theoretic generalization of these results:

PROPOSITION 4.4. *Given an operator $T \in B(H)$, the following hold:*

- (1) *We can write T as follows, with $Re(T), Im(T) \in B(H)$ being self-adjoint:*

$$T = Re(T) + iIm(T)$$

- (2) *When $T = T^*$, we can write T as follows, with $T_+, T_- \in B(H)$ being positive:*

$$T = T_+ - T_-$$

- (3) *Thus, we can write any T as a linear combination of 4 positive elements.*

PROOF. All this follows from basic spectral theory, as follows:

(1) This is something that we have already met in chapter 3 above, when proving the spectral theorem in its general form, the decomposition formula being as follows:

$$T = \frac{T + T^*}{2} - i \cdot \frac{i(T - T^*)}{2}$$

(2) This follows from the measurable functional calculus. Indeed, assuming $T = T^*$ we have $\sigma(T) \subset \mathbb{R}$, so we can use the following decomposition formula on \mathbb{R} :

$$1 = \chi_{[0,\infty)} + \chi_{(-\infty,0)}$$

To be more precise, let us multiply by z , and rewrite this formula as follows:

$$z = \chi_{[0,\infty)}z - \chi_{(-\infty,0)}(-z)$$

Now by applying these measurable functions to T , we obtain as formula as follows, with both the operators $T_+, T_- \in B(H)$ being positive, as desired:

$$T = T_+ - T_-$$

(3) This follows by combining the results in (1) and (2) above. □

We can see from the above the power of the measurable functional calculus. It is possible to do many more things along the same lines, as for instance cutting the bounded operators into spectral projections, once again by using suitable characteristic functions, and so on. We will discuss this later on, in the von Neumann algebra context.

Going ahead now with our decomposition results, let us record as well the following technical statement, which comes as a consequence of Proposition 4.4:

PROPOSITION 4.5. *Any bounded operator $T \in B(H)$ can be written as*

$$T = \lambda_1 U_1 + \lambda_2 U_2 + \lambda_3 U_3 + \lambda_4 U_4$$

with U_i unitaries. That is, T is a linear combination of 4 unitary operators.

PROOF. Indeed, by using Proposition 4.4 we can first write T as a linear combination of 2 self-adjoint operators, and then by functional calculus each of these 2 self-adjoint operators can be written as a linear linear combination of 2 unitary operators. □

Observe that the above result is not exactly along the lines of Proposition 4.4, sharply generalizing what we know about the complex numbers. Indeed, the following formula shows that any complex number $z \in \mathbb{C}$ appears as a real multiple of a unitary:

$$z = r e^{it}$$

However, the above result is technically quite useful. More on this later on.

All this gets us now into the multiplicative theory of the complex numbers, that we will attempt to generalize now. As a first construction, that we would like to generalize to the bounded operator setting, we have the construction of the modulus, as follows:

$$|z| = \sqrt{z\bar{z}}$$

The point now is that we can indeed generalize this construction, as follows:

PROPOSITION 4.6. *Given an operator $T \in B(H)$, we can construct a positive operator $|T| \in B(H)$ as follows, by using the fact that T^*T is positive:*

$$|T| = \sqrt{T^*T}$$

*The square of this operator is then $|T|^2 = T^*T$. In the case $H = \mathbb{C}$, we obtain in this way the usual absolute value of the complex numbers:*

$$|z| = \sqrt{z\bar{z}}$$

More generally, in the case where $H = \mathbb{C}^N$ is finite dimensional, we obtain in this way the usual moduli of the complex matrices $A \in M_N(\mathbb{C})$.

PROOF. We have several things to be proved, the idea being as follows:

(1) The first assertion follows from Theorem 4.2. Indeed, according to (4) there the operator T^*T is indeed positive, and then according to (2) there we can extract the square root of this latter positive operator, by applying to it the function \sqrt{z} .

(2) By functional calculus we have then $|T|^2 = T^*T$, as desired.

(3) In the case $H = \mathbb{C}$, we obtain indeed the absolute value of complex numbers.

(4) In the case where the space H is finite dimensional, $H = \mathbb{C}^N$, we obtain indeed the usual moduli of the complex matrices $A \in M_N(\mathbb{C})$. \square

As a comment here, it is possible to talk as well about the operator $\sqrt{TT^*}$, which is in general different from $\sqrt{T^*T}$. The reasons for using $\sqrt{T^*T}$ instead of $\sqrt{TT^*}$ are quite standard, coming from the polar decomposition formula, to be discussed in a second, that we would like to formulate as $T = U|T|$, rather than $T = |T|U$.

Note that when T is normal, $TT^* = T^*T$, there is no issue here, because we have:

$$\sqrt{TT^*} = \sqrt{T^*T}$$

Regarding now the polar decomposition formula, let us start with a weak version of this statement, regarding the invertible operators, as follows:

THEOREM 4.7. *We have the polar decomposition formula*

$$T = U\sqrt{T^*T}$$

with U being a unitary, for any $T \in B(H)$ invertible.

PROOF. According to our definition of $|T| = \sqrt{T^*T}$, we have:

$$\begin{aligned} \langle |T|x, |T|y \rangle &= \langle x, |T|^2y \rangle \\ &= \langle x, T^*Ty \rangle \\ &= \langle Tx, Ty \rangle \end{aligned}$$

Thus we can define a unitary operator $U \in B(H)$ as follows:

$$U(|T|x) = Tx$$

But this formula shows that we have $T = U|T|$, as desired. \square

Observe that we have uniqueness in the above result, in what regards the choice of the unitary $U \in B(H)$, due to the fact that we can write this unitary as follows:

$$U = T(\sqrt{T^*T})^{-1}$$

More generally now, we have the following result:

THEOREM 4.8. *We have the polar decomposition formula*

$$T = U\sqrt{T^*T}$$

with U being a partial isometry, for any $T \in B(H)$.

PROOF. As before, in the proof of Theorem 4.7, we have the following equality, valid for any two vectors $x, y \in H$:

$$\langle |T|x, |T|y \rangle = \langle Tx, Ty \rangle$$

We conclude from this equality that the following linear application is well-defined, and isometric:

$$U : \text{Im}|T| \rightarrow \text{Im}(T)$$

$$|T|x \rightarrow Tx$$

By continuity we can extend this map U into an isometry between Hilbert subspaces of H , as follows:

$$U : \overline{\text{Im}|T|} \rightarrow \overline{\text{Im}(T)}$$

$$|T|x \rightarrow Tx$$

Moreover, we can further extend U into a partial isometry $U : H \rightarrow H$, by setting $Ux = 0$, for any $x \in \overline{\text{Im}|T|}^\perp$, and with this convention, the result follows. \square

As before, we have uniqueness in the above result.

Summarizing, as a first application of our spectral theory methods, we have now a full generalization of the polar decomposition result for the usual matrices.

4b. Compact operators

We have seen so far the basic theory of the bounded operators, in the arbitrary, normal and self-adjoint cases, and in a few other cases of interest.

We will keep building on this, with a number of more specialized results, regarding the finite rank operators and compact operators, and other special classes of related operators, namely the trace class operators, and the Hilbert-Schmidt operators.

Let us start with a basic definition, as follows:

DEFINITION 4.9. *An operator $T \in B(H)$ is said to be of finite rank if its image*

$$\text{Im}(T) \subset H$$

is finite dimensional. The set of such operators is denoted $B_0(H)$.

There are many interesting examples of finite rank operators, the most basic ones being the finite rank projections, on the finite dimensional subspaces $K \subset H$.

Observe also that in the case where H is finite dimensional, any operator $T \in B(H)$ is automatically of finite rank. In general, we have the following result:

PROPOSITION 4.10. *The set of finite rank operators*

$$B_0(H) \subset B(H)$$

is a two-sided $$ -ideal.*

PROOF. We have several assertions to be proved, the idea being as follows:

(1) It is clear from definitions that $B_0(H)$ is indeed a vector space, with this due to the following formulae, valid for any $S, T \in B(H)$, which are both clear:

$$\dim(\text{Im}(S + T)) \leq \dim(\text{Im}(S)) + \dim(\text{Im}(T))$$

$$\dim(\text{Im}(\lambda T)) = \dim(\text{Im}(T))$$

(2) Let us prove now that $B_0(H)$ is stable under $*$. Given $T \in B_0(H)$, we can regard it as an invertible operator between finite dimensional Hilbert spaces, as follows:

$$T : (\ker T)^\perp \rightarrow \text{Im}(T)$$

Thus, we have the following dimension equality:

$$\dim((\ker T)^\perp) = \dim(\text{Im}(T))$$

Our claim now is that we have in fact equalities as follows:

$$\begin{aligned} \dim(\text{Im}(T^*)) &= \dim(\overline{\text{Im}(T^*)}) \\ &= \dim((\ker T)^\perp) \\ &= \dim(\text{Im}(T)) \end{aligned}$$

Indeed, the third equality is the one above, and the second equality is something that we know too, coming from Proposition 4.1. Now by combining these two equalities we deduce that $\text{Im}(T^*)$ is finite dimensional, and so the first equality holds as well. Thus, our equalities are proved, and this shows that we have $T^* \in B_0(H)$, as desired.

(3) Finally, regarding the ideal property, this follows from the following two formulae, valid for any $S, T \in B(H)$, which are once again clear from definitions:

$$\dim(\text{Im}(ST)) \leq \dim(\text{Im}(T))$$

$$\dim(\text{Im}(TS)) \leq \dim(\text{Im}(T))$$

Thus, we are led to the conclusion in the statement. \square

Let us discuss now the compact operators. These are introduced as follows:

DEFINITION 4.11. *An operator $T \in B(H)$ is said to be compact if the closed set*

$$\overline{T(B_1)} \subset H$$

is compact, where $B_1 \subset H$ is the unit ball. The set of such operators is denoted $B_\infty(H)$.

Equivalently, an operator $T \in B(H)$ is compact when for any sequence $\{x_n\} \subset B_1$, or more generally for any bounded sequence $\{x_n\} \subset H$, the sequence $\{T(x_n)\}$ has a convergence subsequence. We will see later some further criteria of compactness.

In finite dimensions any operator is compact. In general, as a first observation, any finite rank operator is compact. We have in fact the following result:

PROPOSITION 4.12. *Any finite rank operator is compact,*

$$B_0(H) \subset B_\infty(H)$$

and the finite rank operators are dense inside the compact operators.

PROOF. The first assertion is clear, because if $\text{Im}(T)$ is finite dimensional, then the following subset is closed and bounded, and so it is compact:

$$\overline{T(B_1)} \subset \text{Im}(T)$$

Regarding the second assertion, let us pick a compact operator $T \in B_\infty(H)$, and a number $\varepsilon > 0$. By compactness of T we can find a finite set $S \subset B_1$ such that:

$$T(B_1) \subset \bigcup_{x \in S} B_\varepsilon(Tx)$$

Consider now the orthogonal projection P onto the following finite dimensional space:

$$E = \text{span} \left(Tx \mid x \in S \right)$$

Since the set S is finite, this space E is finite dimensional, and so P is of finite rank, $P \in B_0(H)$. Now observe that for any norm one $y \in H$ and any $x \in S$ we have:

$$\begin{aligned} \|Ty - Tx\|^2 &= \|Ty - PTx\|^2 \\ &= \|Ty - PTy + PTy - PTx\|^2 \\ &= \|Ty - PTy\|^2 + \|PTx - PTy\|^2 \end{aligned}$$

Now by picking $x \in S$ such that the ball $B_\varepsilon(Tx)$ covers the point Ty , we conclude from this that we have the following estimate:

$$\|Ty - PTy\| \leq \|Ty - Tx\| \leq \varepsilon$$

Thus we have $\|T - PT\| \leq \varepsilon$, which gives the density result. \square

Quite remarkably, the set of compact operators is closed, and we have:

THEOREM 4.13. *The set of compact operators*

$$B_\infty(H) \subset B(H)$$

*is a closed two-sided *-ideal.*

PROOF. We have several assertions here, the idea being as follows:

(1) It is clear from definitions that $B_\infty(H)$ is indeed a vector space, with this due to the following formulae, valid for any $S, T \in B(H)$, which are both clear:

$$\begin{aligned} (S + T)(B_1) &\subset S(B_1) + T(B_1) \\ (\lambda T)(B_1) &= |\lambda| \cdot T(B_1) \end{aligned}$$

(2) In order to prove now that $B_\infty(H)$ is closed, assume that a sequence $T_n \in B_\infty(H)$ converges to $T \in B(H)$. Given $\varepsilon > 0$, let us pick $N \in \mathbb{N}$ such that:

$$\|T - T_N\| \leq \varepsilon$$

By compactness of T_N we can find a finite set $S \subset B_1$ such that:

$$T_N(B_1) \subset \bigcup_{x \in S} B_\varepsilon(T_N x)$$

We conclude that for any $y \in B_1$ there exists $x \in S$ such that:

$$\begin{aligned} \|Ty - Tx\| &\leq \|Ty - T_N y\| + \|T_N y - T_N x\| + \|T_N x - Tx\| \\ &\leq \varepsilon + \varepsilon + \varepsilon \\ &= 3\varepsilon \end{aligned}$$

Thus, we have an inclusion as follows, with $S \subset B_1$ being finite:

$$T(B_1) \subset \bigcup_{x \in S} B_{3\varepsilon}(Tx)$$

But this shows that our limiting operator T is compact, as desired.

(3) Regarding the fact that $B_\infty(H)$ is stable under involution, this follows from Proposition 4.10, Proposition 4.12 and (2). Indeed, by using Proposition 4.12, given $T \in B_\infty(H)$ we can write it as a limit of finite rank operators, as follows:

$$T = \lim_{n \rightarrow \infty} T_n$$

Now by applying the adjoint, we obtain that we have as well:

$$T^* = \lim_{n \rightarrow \infty} T_n^*$$

We know from Proposition 4.10 that the operators T_n^* are of finite rank, and so compact by Proposition 4.12, and by using (2) we obtain that T^* is compact too, as desired.

(4) Finally, regarding the ideal property, this follows from the following two formulae, valid for any $S, T \in B(H)$, which are once again clear from definitions:

$$(ST)(B_1) = S(T(B_1))$$

$$(TS)(B_1) \subset \|S\| \cdot T(B_1)$$

Thus, we are led to the conclusion in the statement. \square

Here is now a second key result regarding the compact operators:

THEOREM 4.14. *A bounded operator $T \in B(H)$ is compact precisely when*

$$Te_n \rightarrow 0$$

for any orthonormal system $\{e_n\} \subset H$.

PROOF. We have two implications to be proved, the idea being as follows:

“ \implies ” Assume that T is compact. By contradiction, assume $Te_n \not\rightarrow 0$. This means that there exists $\varepsilon > 0$ and a subsequence satisfying $\|Te_{n_k}\| > \varepsilon$, and by replacing $\{e_n\}$ with this subsequence, we can assume that the following holds, with $\varepsilon > 0$:

$$\|Te_n\| > \varepsilon$$

Since T was assumed to be compact, and the sequence $\{e_n\}$ is bounded, a certain subsequence $\{Te_{n_k}\}$ must converge. Thus, by replacing once again $\{e_n\}$ with a subsequence, we can assume that the following holds, with $x \neq 0$:

$$Te_n \rightarrow x$$

But this is a contradiction, because we obtain in this way:

$$\begin{aligned} \langle x, x \rangle &= \lim_{n \rightarrow \infty} \langle Te_n, x \rangle \\ &= \lim_{n \rightarrow \infty} \langle e_n, T^*x \rangle \\ &= 0 \end{aligned}$$

Thus our assumption $Te_n \not\rightarrow 0$ was wrong, and we obtain the result.

“ \Leftarrow ” Assume $Te_n \rightarrow 0$, for any orthonormal system $\{e_n\} \subset H$. In order to prove that T is compact, we use the various results established above, which show that this is the same as proving that T is in the closure of the space of finite rank operators:

$$T \in \overline{B_0(H)}$$

We do this by contradiction. So, assume that the above is wrong, and so that there exists $\varepsilon > 0$ such that the following holds:

$$S \in B_0(H) \implies \|T - S\| > \varepsilon$$

As a first observation, by using $S = 0$ we obtain $\|T\| > \varepsilon$. Thus, we can find a norm one vector $e_1 \in H$ such that the following holds:

$$\|Te_1\| > \varepsilon$$

Our claim, which will bring the desired contradiction, is that we can construct by recurrence vectors e_1, \dots, e_n such that the following holds, for any i :

$$\|Te_i\| > \varepsilon$$

Indeed, assume that we have constructed such vectors e_1, \dots, e_n . Let $E \subset H$ be the linear space spanned by these vectors, and let us set:

$$P = Proj(E)$$

Since the operator TP has finite rank, our assumption above shows that we have:

$$\|T - TP\| > \varepsilon$$

Thus, we can find a vector $x \in H$ such that:

$$\|(T - TP)x\| > \varepsilon$$

We have then $x \notin E$, and so we can consider the following nonzero vector:

$$y = (1 - P)x$$

With this nonzero vector y constructed, now let us set:

$$e_{n+1} = \frac{y}{\|y\|}$$

This vector e_{n+1} is then orthogonal to E , has norm one, and satisfies:

$$\|Te_{n+1}\| \geq \|y\|^{-1}\varepsilon \geq \varepsilon$$

Thus we are done with our construction by recurrence, and this contradicts our assumption that $Te_n \rightarrow 0$, for any orthonormal system $\{e_n\} \subset H$, as desired. \square

Summarizing, we have so far a number of results regarding the compact operators, in analogy with what we know about the usual complex matrices.

Let us discuss now the spectral theory of the compact operators. We first have:

PROPOSITION 4.15. *Assuming that $T \in B(H)$, with $\dim H = \infty$, is compact and self-adjoint, the following happen:*

- (1) *The eigenvalues of T form a sequence $\lambda_n \rightarrow 0$.*
- (2) *All eigenvalues $\lambda_n \neq 0$ have finite multiplicity.*

PROOF. We prove both the assertions at the same time. For this purpose, we fix a number $\varepsilon > 0$, we consider all the eigenvalues satisfying $|\lambda| \geq \varepsilon$, and for each such eigenvalue we consider the corresponding eigenspace $E_\lambda \subset H$. Let us set:

$$E = \text{span} \left(E_\lambda \mid |\lambda| \geq \varepsilon \right)$$

Our claim, which will prove both (1) and (2), is that this space E is finite dimensional. In order to prove now this claim, we can proceed as follows:

- (1) We know that we have $E \subset \text{Im}(T)$. Our claim is that we have:

$$\bar{E} \subset \text{Im}(T)$$

Indeed, assume that we have a sequence $g_n \in E$ which converges, $g_n \rightarrow g \in \bar{E}$. Let us write $g_n = Tf_n$, with $f_n \in H$. By definition of E , the following condition is satisfied:

$$h \in E \implies \|Th\| \geq \varepsilon\|h\|$$

Now since the sequence $\{g_n\}$ is Cauchy we obtain from this that the sequence $\{f_n\}$ is Cauchy as well, and with $f_n \rightarrow f$ we have $Tf_n \rightarrow Tf$, as desired.

- (2) Consider now the projection $P \in B(H)$ onto the above space \bar{E} . The composition PT is then as follows, surjective on its target:

$$PT : H \rightarrow \bar{E}$$

On the other hand since T is compact so must be PT , and it follows from this that the space \bar{E} is finite dimensional. Thus E itself must be finite dimensional too, and as explained in the beginning of the proof, this gives (1) and (2), as desired. \square

In order to construct now eigenvalues, we will need:

PROPOSITION 4.16. *If T is compact and self-adjoint, one of the numbers*

$$\|T\|, -\|T\|$$

must be an eigenvalue of T .

PROOF. We know from the spectral theory of the self-adjoint operators that the spectral radius $\|T\|$ of our operator T is attained, and so one of the numbers $\|T\|, -\|T\|$ must be in the spectrum. In order to prove now that one of these numbers must actually appear as an eigenvalue, we must use the compactness of T , as follows:

(1) First, we can assume $\|T\| = 1$. By functional calculus this implies $\|T^3\| = 1$ too, and so we can find a sequence of norm one vectors $x_n \in H$ such that:

$$|\langle T^3 x_n, x_n \rangle| \rightarrow 1$$

By using our assumption $T = T^*$, we can rewrite this formula as follows:

$$|\langle T^2 x_n, T x_n \rangle| \rightarrow 1$$

Now since T is compact, and $\{x_n\}$ is bounded, we can assume, up to changing the sequence $\{x_n\}$ to one of its subsequences, that the sequence $T x_n$ converges:

$$T x_n \rightarrow y$$

Thus, the convergence formula found above reformulates as follows, with $y \neq 0$:

$$|\langle T y, y \rangle| = 1$$

(2) Our claim now, which will finish the proof, is that this latter formula implies $T y = \pm y$. Indeed, by using Cauchy-Schwarz and $\|T\| = 1$, we have:

$$|\langle T y, y \rangle| \leq \|T y\| \cdot \|y\| \leq 1$$

We know that this must be an equality, so $T y, y$ must be proportional. But since T is self-adjoint the proportionality factor must be ± 1 , and so we obtain, as claimed:

$$T y = \pm y$$

Thus, we have constructed an eigenvector for $\lambda = \pm 1$, as desired. \square

We can further build on the above results in the following way:

PROPOSITION 4.17. *If T is compact and self-adjoint, there is an orthogonal basis of H made of eigenvectors of T .*

PROOF. We use Proposition 4.15. According to the results there, we can arrange the nonzero eigenvalues of T , taken with multiplicities, into a sequence $\lambda_n \rightarrow 0$. Let $y_n \in H$ be the corresponding eigenvectors, and consider the following space:

$$E = \overline{\text{span}(y_n)}$$

The result follows then from the following observations:

(1) Since we have $T = T^*$, both E and its orthogonal E^\perp are invariant under T .

(2) On the space E , our operator T is by definition diagonal.

(3) On the space E^\perp , our claim is that we have $T = 0$. Indeed, assuming that the restriction $S = T|_{E^\perp}$ is nonzero, we can apply Proposition 4.16 to this restriction, and we obtain an eigenvalue for S , and so for T , contradicting the maximality of E . \square

With the above results in hand, we can now formulate a first spectral theory result for compact operators, which closes the discussion in the self-adjoint case:

THEOREM 4.18. *Assuming that $T \in B(H)$, with $\dim H = \infty$, is compact and self-adjoint, the following happen:*

- (1) *The spectrum $\sigma(T) \subset \mathbb{R}$ consists of a sequence $\lambda_n \rightarrow 0$.*
- (2) *All spectral values $\lambda \in \sigma(T) - \{0\}$ are eigenvalues.*
- (3) *All eigenvalues $\lambda \in \sigma(T) - \{0\}$ have finite multiplicity.*
- (4) *There is an orthogonal basis of H made of eigenvectors of T .*

PROOF. This follows from the various results established above:

(1) In view of Proposition 4.15 (1), this will follow from (2) below.

(2) Assume that $\lambda \neq 0$ belongs to the spectrum $\sigma(T)$, but is not an eigenvalue. By using Proposition 4.17, let us pick an orthonormal basis $\{e_n\}$ of H consisting of eigenvectors of T , and then consider the following operator:

$$Sx = \sum_n \frac{\langle x, e_n \rangle}{\lambda_n - \lambda} e_n$$

Then S is an inverse for $T - \lambda$, and so we have $\lambda \notin \sigma(T)$, as desired.

(3) This is something that we know, from Proposition 4.15 (2).

(4) This is something that we know too, from Proposition 4.17. □

Finally, we have the following result, regarding the general case:

THEOREM 4.19. *The compact operators $T \in B(H)$, with $\dim H = \infty$, are the operators of the following form, with $\{e_n\}$, $\{f_n\}$ being orthonormal families, and with $\lambda_n \searrow 0$:*

$$T(x) = \sum_n \lambda_n \langle x, e_n \rangle f_n$$

The numbers λ_n , called singular values of T , are the eigenvalues of $|T|$. In fact, the polar decomposition of T is given by $T = U|T|$, with

$$|T|(x) = \sum_n \lambda_n \langle x, e_n \rangle e_n$$

and with U being given by $Ue_n = f_n$, and $U = 0$ on the complement of $\text{span}(e_i)$.

PROOF. This basically follows from Theorem 4.8 and Theorem 4.18, as follows:

(1) Given two orthonormal families $\{e_n\}$, $\{f_n\}$, and a sequence of real numbers $\lambda_n \searrow 0$, consider the linear operator given by the formula in the statement, namely:

$$T(x) = \sum_n \lambda_n \langle x, e_n \rangle f_n$$

Our first claim is that T is bounded. Indeed, when assuming $|\lambda_n| \leq \varepsilon$ for any n , which is something that we can do if we want to prove that T is bounded, we have:

$$\begin{aligned} \|T(x)\|^2 &= \left| \sum_n \lambda_n \langle x, e_n \rangle f_n \right|^2 \\ &= \sum_n |\lambda_n|^2 |\langle x, e_n \rangle|^2 \\ &\leq \varepsilon^2 \sum_n |\langle x, e_n \rangle|^2 \\ &\leq \varepsilon^2 \|x\|^2 \end{aligned}$$

(2) The next observation is that this operator is indeed compact, because it appears as the norm limit, $T_N \rightarrow T$, of the following sequence of finite rank operators:

$$T_N = \sum_{n \leq N} \lambda_n \langle x, e_n \rangle f_n$$

(3) Regarding now the polar decomposition assertion, for the above operator, this follows once again from definitions. Indeed, the adjoint is given by:

$$T^*(x) = \sum_n \lambda_n \langle x, f_n \rangle e_n$$

Thus, when composing T^* with T , we obtain the following operator:

$$T^*T(x) = \sum_n \lambda_n^2 \langle x, e_n \rangle e_n$$

Now by extracting the square root, we obtain the formula in the statement, namely:

$$|T|(x) = \sum_n \lambda_n \langle x, e_n \rangle e_n$$

(4) Conversely now, assume that $T \in B(H)$ is compact. Then T^*T , which is self-adjoint, must be compact as well, and so by Theorem 4.18 we have a formula as follows, with $\{e_n\}$ being a certain orthonormal family, and with $\lambda_n \searrow 0$:

$$T^*T(x) = \sum_n \lambda_n^2 \langle x, e_n \rangle e_n$$

By extracting the square root we obtain the formula of $|T|$ in the statement, and then by setting $U(e_n) = f_n$ we obtain a second orthonormal family, $\{f_n\}$, such that:

$$\begin{aligned} T(x) &= U|T| \\ &= \sum_n \lambda_n \langle x, e_n \rangle f_n \end{aligned}$$

Thus, our compact operator $T \in B(H)$ appears indeed as in the statement. \square

As a technical remark here, it is possible to slightly improve a part of the above statement. Consider indeed an operator of the following form, with $\{e_n\}$, $\{f_n\}$ being orthonormal families as before, and with $\lambda_n \rightarrow 0$ being now complex numbers:

$$T(x) = \sum_n \lambda_n \langle x, e_n \rangle f_n$$

Then the same proof as before shows that T is compact, and that the polar decomposition of T is given by $T = U|T|$, with the modulus $|T|$ being as follows:

$$|T|(x) = \sum_n |\lambda_n| \langle x, e_n \rangle e_n$$

As for the partial isometry U , this is given by $Ue_n = w_n f_n$, and $U = 0$ on the complement of $\text{span}(e_i)$, where $w_n \in \mathbb{T}$ are such that $\lambda_n = |\lambda_n| w_n$.

Thus, we have indeed an extension of the above results.

4c. Trace class operators

We have not talked so far about the trace of operators $T \in B(H)$, in analogy with the trace of the usual matrices $M \in M_N(\mathbb{C})$. This is because the trace can be finite or infinite, or even not well-defined, and we will discuss this now. Let us start with:

PROPOSITION 4.20. *Given a positive operator $T \in B(H)$, the quantity*

$$\text{Tr}(T) = \sum_n \langle T e_n, e_n \rangle \in [0, \infty]$$

is independent on the choice of an orthonormal basis $\{e_n\}$.

PROOF. If $\{f_n\}$ is another orthonormal basis, we have:

$$\begin{aligned} \sum_n \langle T f_n, f_n \rangle &= \sum_n \langle \sqrt{T} f_n, \sqrt{T} f_n \rangle \\ &= \sum_n \|\sqrt{T} f_n\|^2 \\ &= \sum_{mn} |\langle \sqrt{T} f_n, e_m \rangle|^2 \\ &= \sum_{mn} |\langle T^{1/4} f_n, T^{1/4} e_m \rangle|^2 \end{aligned}$$

Since this quantity is symmetric in e, f , this gives the result. \square

We can now introduce the trace class operators, as follows:

DEFINITION 4.21. An operator $T \in B(H)$ is said to be of trace class if:

$$\text{Tr}|T| < \infty$$

The set of such operators is denoted $B_1(H)$.

In finite dimensions, any operator is of course of trace class. In arbitrary dimension, finite or not, we first have the following result, regarding such operators:

PROPOSITION 4.22. Any finite rank operator is of trace class, and any trace class operator is compact, so that we have embeddings as follows:

$$B_0(H) \subset B_1(H) \subset B_\infty(H)$$

Moreover, for any compact operator $T \in B_\infty(H)$ we have the formula

$$\text{Tr}|T| = \sum_n \lambda_n$$

where $\lambda_n \geq 0$ are the singular values, and so $T \in B_1(H)$ precisely when $\sum_n \lambda_n < \infty$.

PROOF. We have several assertions here, the idea being as follows:

(1) If T is of finite rank, it is clearly of trace class.

(2) In order to prove now the second assertion, assume first that $T > 0$ is of trace class. For any orthonormal basis $\{e_n\}$ we have:

$$\begin{aligned} \sum_n \|\sqrt{T}e_n\|^2 &= \sum_n \langle Te_n, e_n \rangle \\ &\leq \text{Tr}(T) \\ &< \infty \end{aligned}$$

Thus we have $\sqrt{T}e_n \rightarrow 0$, and so \sqrt{T} is compact. Now since the compact operators form an ideal, it follows that $T = \sqrt{T} \cdot \sqrt{T}$ is compact as well.

(3) In order to prove now the second assertion in general, assume that $T \in B(H)$ is of trace class. Then $|T|$ is also of trace class, and so compact by (2), and since we have $T = U|T|$ by polar decomposition, it follows that T is compact too.

(4) Finally, in order to prove the last assertion, assume that T is compact. The singular value decomposition of $|T|$, from Theorem 4.19 above, is then as follows:

$$|T|(x) = \sum_n \lambda_n \langle x, e_n \rangle e_n$$

But this gives the formula for $\text{Tr}|T|$ in the statement, and proves the last assertion. \square

Here is a useful reformulation of the above result, or rather of the above result coupled with Theorem 4.19, without reference to compact operators:

THEOREM 4.23. *The trace class operators are precisely the operators of the form*

$$|T|(x) = \sum_n \lambda_n \langle x, e_n \rangle f_n$$

with $\{e_n\}, \{f_n\}$ being orthonormal systems, and with $\lambda \searrow 0$ being a sequence satisfying:

$$\sum_n \lambda_n < \infty$$

Moreover, for such an operator we have the following estimate:

$$|Tr(T)| \leq Tr|T| = \sum_n \lambda_n$$

PROOF. This follows indeed from Proposition 4.22, or rather for step (4) in the proof of Proposition 4.22, coupled with Theorem 4.19 above. \square

Next, we have the following result, which comes as a continuation of Proposition 4.22, and is our central result here, regarding the trace class operators:

THEOREM 4.24. *The space of trace class operators, which appears as an intermediate space between the finite rank operators and the compact operators,*

$$B_0(H) \subset B_1(H) \subset B_\infty(H)$$

is a two-sided $*$ -ideal of $B_\infty(H)$. The following is a Banach space norm on $B_1(H)$,

$$\|T\|_1 = Tr|T|$$

satisfying $\|T\| \leq \|T\|_1$, and for $T \in B_1(H)$ and $S \in B(H)$ we have:

$$\|ST\|_1 \leq \|S\| \cdot \|T\|_1$$

Also, the subspace $B_0(H)$ is dense inside $B_1(H)$, with respect to this norm.

PROOF. There are several assertions here, the idea being as follows:

(1) In order to prove that $B_1(H)$ is a linear space, and that $\|T\|_1 = Tr|T|$ is a norm on it, the only non-trivial point is that of proving the following inequality:

$$Tr|S + T| \leq Tr|S| + Tr|T|$$

For this purpose, consider the polar decompositions of these operators:

$$S = U|S|$$

$$T = V|T|$$

$$S + T = W|S + T|$$

Given an orthonormal basis $\{e_n\}$, we have the following formula:

$$\begin{aligned} \text{Tr}|S + T| &= \sum_n \langle |S + T|e_n, e_n \rangle \\ &= \sum_n \langle W^*(S + T)e_n, e_n \rangle \\ &= \sum_n \langle W^*U|S|e_n, e_n \rangle + \sum_n \langle W^*V|T|e_n, e_n \rangle \end{aligned}$$

The point now is that the first sum can be estimated as follows:

$$\begin{aligned} &\sum_n \langle W^*U|S|e_n, e_n \rangle \\ &= \sum_n \langle \sqrt{|S|}e_n, \sqrt{|S|}U^*W e_n \rangle \\ &\leq \sum_n \left\| \sqrt{|S|}e_n \right\| \cdot \left\| \sqrt{|S|}U^*W e_n \right\| \\ &\leq \sqrt{\sum_n \left\| \sqrt{|S|}e_n \right\|^2} \cdot \sqrt{\sum_n \left\| \sqrt{|S|}U^*W e_n \right\|^2} \end{aligned}$$

In order to estimate the terms on the right, we can proceed as follows:

$$\begin{aligned} \sum_n \left\| \sqrt{|S|}U^*W e_n \right\|^2 &= \sum_n \langle W^*U|S|U^*W e_n, e_n \rangle \\ &= \text{Tr}(W^*U|S|U^*W) \\ &\leq \text{Tr}(U|S|U^*) \\ &\leq \text{Tr}(|S|) \end{aligned}$$

The second sum in the above formula of $\text{Tr}|S + T|$ can be estimated in the same way, and in the end we obtain, as desired:

$$\text{Tr}|S + T| \leq \text{Tr}|S| + \text{Tr}|T|$$

(2) The estimate $\|T\| \leq \|T\|_1$ can be established as follows:

$$\begin{aligned} \|T\| &= \||T|\| \\ &= \sup_{\|x\|=1} \langle |T|x, x \rangle \\ &\leq \text{Tr}|T| \end{aligned}$$

(3) The fact that $B_1(H)$ is indeed a Banach space follows by constructing a limit for any Cauchy sequence, by using the singular value decomposition.

(4) The fact that $B_1(H)$ is indeed closed under the involution follows from:

$$\begin{aligned} \text{Tr}(T^*) &= \sum_n \langle T^* e_n, e_n \rangle \\ &= \sum_n \langle e_n, T e_n \rangle \\ &= \overline{\text{Tr}(T)} \end{aligned}$$

(5) In order to prove the ideal property of $B_1(H)$, we use the standard fact, that we already know from the above, that any bounded operator $T \in B(H)$ can be written as a linear combination of 4 unitary operators:

$$T = \lambda_1 U_1 + \lambda_2 U_2 + \lambda_3 U_3 + \lambda_4 U_4$$

Indeed, by taking the real and imaginary part we can first write T as a linear combination of 2 self-adjoint operators, and then by functional calculus each of these 2 self-adjoint operators can be written as a linear linear combination of 2 unitary operators.

(6) With this trick in hand, we can now prove the ideal property of $B_1(H)$. Indeed, it is enough to prove that $T \in B_1(H)$ and $U \in U(H)$ implies $UT, TU \in B_1(H)$. But this latter result follows by using the polar decomposition theorem.

(7) With a bit more care, we obtain from this the estimate $\|ST\|_1 \leq \|S\| \cdot \|T\|_1$ from the statement. As for the last assertion, this is clear as well. \square

This was for the basic theory of the trace class operators. More can be said, and we refer here to the literature. We will be back to this.

4d. Hilbert-Schmidt operators

As a last topic of this chapter, let us discuss yet another important class of operators, namely the Hilbert-Schmidt ones.

These operators, that we will need in what follows, are introduced as follows:

DEFINITION 4.25. *An operator $T \in B(H)$ is said to be Hilbert-Schmidt if:*

$$\text{Tr}(T^*T) < \infty$$

The set of such operators is denoted $B_2(H)$.

As before with other sets of operators, in finite dimensions we obtain in this way all the operators.

In general, we have the following result, regarding such operators:

THEOREM 4.26. *The space $B_2(H)$ of Hilbert-Schmidt operators, which appears as an intermediate space between the trace class operators and the compact operators,*

$$B_0(H) \subset B_1(H) \subset B_2(H) \subset B_\infty(H)$$

*is a two-sided *-ideal of $B_\infty(H)$. This ideal has the property*

$$S, T \in B_2(H) \implies ST \in B_1(H)$$

and conversely, each $T \in B_1(H)$ appears as product of two operators in $B_2(H)$. In terms of the singular values (λ_n) , the Hilbert-Schmidt operators are characterized by:

$$\sum_n \lambda_n^2 < \infty$$

Also, the following formula, whose output is finite by Cauchy-Schwarz,

$$\langle S, T \rangle = \text{Tr}(ST^*)$$

defines a scalar product of $B_2(H)$, making it a Hilbert space.

PROOF. All this is quite standard, from the results that we have already, and more specifically from the singular value decomposition theorem, and its applications. To be more precise, the proof of the various assertions goes as follows:

(1) First of all, the fact that the space of Hilbert-Schmidt operators $B_2(H)$ is stable under taking sums, and so is a vector space, follows from:

$$\begin{aligned} (S+T)^*(S+T) &\leq (S+T)^*(S+T) + (S-T)^*(S-T) \\ &= (S^*+T^*)(S+T) + (S^*-T^*)(S-T) \\ &= 2(S^*S+T^*T) \end{aligned}$$

Regarding now multiplicative properties, we can use here the following inequality:

$$\begin{aligned} (ST)^*(ST) &= T^*S^*ST \\ &\leq \|S\|^2 T^*T \end{aligned}$$

Thus, the space $B_2(H)$ is a two-sided *-ideal of $B_\infty(H)$, as claimed.

(2) In order to prove now that the product of any two Hilbert-Schmidt operators is a trace class operator, we can use the following formula, which is elementary:

$$S^*T = \sum_{k=1}^4 i^k (S-iT)^*(S-iT)$$

Conversely, given an arbitrary trace class operator $T \in B_1(H)$, we have:

$$\begin{aligned} T \in B_1(H) &\implies |T| \in B_1(H) \\ &\implies \sqrt{|T|} \in B_2(H) \end{aligned}$$

Thus, by using the polar decomposition $T = U|T|$, we obtain the following decomposition for T , with both components being Hilbert-Schmidt operators:

$$\begin{aligned} T &= U|T| \\ &= U\sqrt{|T|} \cdot \sqrt{|T|} \end{aligned}$$

- (3) The condition for the singular values is clear.
- (4) The fact that we have a scalar product is clear as well.
- (5) The proof of the completeness property is routine as well. □

In the above result, the key statement is the last one, allowing us to use Hilbert space techniques for the study of the operators $T \in B_2(H)$. We will be back to this.

This was for the basic theory of bounded operators on a Hilbert space, $T \in B(H)$. In the remainder of this book we will be rather interested in the operator algebras $A \subset B(H)$ that these operators can form. This is of course related to operator theory, because we can, at least in theory, take $A = \langle T \rangle$, and then study T via the properties of A . Actually, this is something that we already did a few times, when doing spectral theory, and notably when talking about functional calculus for normal operators.

For further operator theory, however, nothing beats a good operator theory book, and various ad-hoc methods, depending on the type of operators involved.

Let us mention as well that there is a lot of interesting theory regarding the unbounded operators $T \in \mathcal{L}(H)$, which is something quite technical, and here once again, we warmly recommend a good operator theory book. In addition, we recommend as well a good PDE book, and this because most of the questions making appear unbounded operators usually have PDE formulations as well, which are extremely efficient.

Finally, in what concerns these latter questions, many of them come from quantum mechanics, or at least from the quantum mechanics which is presently known. Whether some of these questions can be reformulated or not in terms of bounded operators remains to be seen, and no one really knows the answer here.

4e. Exercises

There has been a lot of theory in this chapter, with some of the things not really explained in great detail, and we have several exercises about all this. First comes:

EXERCISE 4.27. *Prove that we have the formula*

$$K^{\perp\perp} = \bar{K}$$

valid for any linear subspace $K \subset H$ of a Hilbert space.

This is something that we used in the above, with the comment that this comes as an easy application of the Hahn-Banach theorem. Time now to do this, in detail.

Another exercise now, which is more philosophical, is:

EXERCISE 4.28. *Try to find the best operator theoretic analogue of the usual formula*

$$z = re^{it}$$

for the complex numbers, telling us that any number is a real multiple of a unitary.

As explained in the above, a weak analogue of this holds, stating that any operator is a linear combination of 4 unitaries. The problem is that of improving this.

In relation now with the polar decomposition theorem, we have:

EXERCISE 4.29. *Work out a few explicit examples of the polar decomposition formula*

$$T = U\sqrt{T^*T}$$

with, if possible, a non-trivial computation for the square root.

This is actually something quite tricky, even for the usual matrices.

Regarding now the compact operators, of all types, we have here:

EXERCISE 4.30. *Look up the various extra general properties of the sets of finite rank, trace class, Hilbert-Schmidt and compact operators,*

$$B_0(H) \subset B_1(H) \subset B_2(H) \subset B_\infty(H)$$

coming in addition to what has been said above, about such operators.

This is of course quite vague, and, as good news, it is not indicated either if you should just come with a list of such properties, or with a list of such properties coming with complete proofs. Up to you here, and the more the better.

Finally, as already mentioned, there are so many other things that can be said about the operators. Here is, as an example, something that you should learn too:

EXERCISE 4.31. *Look up the Fredholm alternative theorem, and write a brief account of what you found and learned.*

There are actually many possible formulations here, and the first task is that of choosing a reference text that you like, and trust. Go your own way, of course.

Part II

Operator algebras

*There was something in the air that night
The stars were bright, Fernando
They were shining there for you and me
For liberty, Fernando*

CHAPTER 5

Operator algebras

5a. Basic theory

We have seen that the study of the bounded operators $T \in B(H)$ often leads to the consideration of the algebras $\langle T \rangle \subset B(H)$ generated by such operators, the idea being that the study of $A = \langle T \rangle$ can lead to results about T itself. This method is particularly efficient in the study of the normal operators, $TT^* = T^*T$, where the algebra $A = \langle T \rangle$ is commutative, with all this being related to the Spectral Theorem.

In this chapter, and in the remainder of this book, we will focus on the study of such algebras $A \subset B(H)$, which can be singly generated, $A = \langle T \rangle$, or multiply generated, $A = \langle T_i \rangle$, or simply abstract, $A \subset B(H)$. The main motivations will become clear later on, the idea being that the study of such abstract algebras $A \subset B(H)$ leads, after some work, into some interesting problems coming from quantum physics.

Let us start our discussion with the following broad definition:

DEFINITION 5.1. *An operator algebra is an algebra of bounded operators*

$$A \subset B(H)$$

which contains the unit, is closed under taking adjoints,

$$T \in A \implies T^* \in A$$

and is closed as well under the norm.

Here, as in the previous chapters, H is an arbitrary Hilbert space, with the case that we are mostly interested in being the separable one. By separable we mean having a countable orthonormal basis, $\{e_i\}_{i \in I}$ with I countable, and such a space is of course unique. The simplest model is the space $l^2(\mathbb{N})$, but in practice, we are particularly interested in the spaces of the form $H = L^2(X)$, which are separable too, but with the basis $\{e_i\}_{i \in \mathbb{N}}$ and the subsequent identification $H \simeq l^2(\mathbb{N})$ being not necessarily very explicit.

Also as in the previous chapters, $B(H)$ is the algebra of linear operators $T : H \rightarrow H$ which are bounded, in the sense that the norm $\|T\| = \sup_{\|x\|=1} \|Tx\|$ is finite. This algebra has an involution $T \rightarrow T^*$, with the adjoint operator $T^* \in B(H)$ being defined

by the formula $\langle Tx, y \rangle = \langle x, T^*y \rangle$, and in the above definition, the assumption $T \in A \implies T^* \in A$ refers to this involution. Thus, A must be a $*$ -algebra.

The assumption that A must be a $*$ -algebra comes from operator theory, or even from usual linear algebra. Indeed, the adjoint matrix or operator $T^* \in B(H)$ is a sort of “twin brother” to the original operator $T \in B(H)$, with these two operators being related by a lot of interesting mathematics, starting from the very basics, such as the equations $U^* = U^{-1}$ for the unitaries or $P^2 = P = P^*$ for the projections, and then going up to fairly advanced theory, such as the subtle normality condition $TT^* = T^*T$, and its consequences. Thus, it is unthinkable to assume $T \in A$, and $T^* \notin A$.

As for the assumption that A is norm closed, which is automatic in finite dimensions, this comes as well from operator theory. We have seen indeed in the previous chapters that in infinite dimensions things are quite tricky, and analytic. Thus, in order to get to something interesting, we must be able to do some analysis inside A , and the least of the things here is that of assuming that our Cauchy sequences inside A converge.

With the above definition given, in relation with the normal operators, where most of the non-trivial results that we have so far are, we have the following result:

THEOREM 5.2. *The operator algebra $\langle T \rangle \subset B(H)$ generated by a normal operator $T \in B(H)$ appears as an algebra of functions,*

$$\langle T \rangle = C(\sigma(T))$$

where $\sigma(T) \subset \mathbb{C}$ denotes as usual the spectrum of T .

PROOF. This is an abstract reformulation of the continuous functional calculus theorem for the normal operators, that we know from chapter 3 above. Indeed, that theorem tells us that we have a continuous morphism of $*$ -algebras, as follows:

$$\begin{aligned} C(\sigma(T)) &\rightarrow B(H) \\ f &\rightarrow f(T) \end{aligned}$$

Moreover, by the general properties of the continuous calculus, also established in chapter 3, this morphism is injective, and its image is the norm closed algebra $\langle T \rangle$ generated by T, T^* . Thus, we obtain the isomorphism in the statement. \square

It is possible to further build on the above result, as follows:

THEOREM 5.3. *The commutative operator algebras $A \subset B(H)$ are, when regarded as abstract Banach $*$ -algebras, exactly the algebras of the form*

$$A = C(X)$$

with the “spectrum” X being the space of Banach algebra characters $\chi : A \rightarrow \mathbb{C}$, with topology making each of the evaluation maps $ev_T : \chi \rightarrow \chi(T)$ continuous.

PROOF. We have several things to be proved, the idea being as follows:

(1) The first observation, which is elementary, coming from basic analysis, is that given a compact space X , the algebra $A = C(X)$ of complex continuous functions $f : X \rightarrow \mathbb{C}$ is indeed a Banach $*$ -algebra, with norm and involution as follows:

$$\|f\| = \sup_{x \in X} |f(x)|$$

$$f^*(x) = \overline{f(x)}$$

(2) Our next claim is that we can realize $A = C(X)$ as an algebra of operators, on a certain Hilbert space. To be more precise, let us pick a probability measure μ on the space X . Our claim is that we have an embedding as follows, with $H = L^2(X)$:

$$A \subset B(H)$$

$$f \rightarrow (g \rightarrow fg)$$

(3) Indeed, given a continuous function $f \in C(X)$, consider the following linear operator, acting on the Hilbert space $H = L^2(X)$:

$$T_f(g) = fg$$

Observe that T_f is well-defined, and bounded as well, because:

$$\|fg\|_2 = \sqrt{\int_X |f(x)|^2 |g(x)|^2 d\mu(x)}$$

$$\leq \|f\|_\infty \|g\|_2$$

The application $f \rightarrow T_f$ being linear, involutive, continuous, and injective as well, we obtain in this way a Banach $*$ -algebra embedding $A \subset B(H)$, as claimed.

(4) Conversely, given a commutative operator algebra $A \subset B(H)$, we can define X as in the statement. Then X is compact, and $T \rightarrow ev_T$ is a morphism of algebras:

$$ev : A \rightarrow C(X)$$

(5) We first prove that ev is involutive. We use the following formula, which is similar to the $z = Re(z) + iIm(z)$ formula for the usual complex numbers:

$$T = \frac{T + T^*}{2} - i \cdot \frac{i(T - T^*)}{2}$$

Thus it is enough to prove the equality $ev_{T^*} = ev_T^*$ for self-adjoint elements T . In other words, we would like to prove that $T = T^*$ implies that:

$$ev_T = ev_T^*$$

But this is the same as proving that $T = T^*$ implies that ev_T is a real function, which is in turn true, because $ev_T(\chi) = \chi(T)$ is an element of $\sigma(T)$, contained in \mathbb{R} .

(6) Since the algebra A is commutative, each element is normal, and so ev is isometric, due to the following norm computation:

$$\begin{aligned} \|ev_T\| &= \rho(T) \\ &= \|T\| \end{aligned}$$

(7) In order to finish the proof, it remains to prove that the evaluation map ev is surjective. But this follows from the Stone-Weierstrass theorem, because $ev(A)$ is a closed subalgebra of $C(X)$, which separates the points. \square

The above result is not the end of the story, because we still have the question of understanding the structure of the commutative operator algebras $A \subset B(H)$, up to spatial isomorphism. This is something quite technical, and we will be back to it.

As a second comment, a quick look at what we have so far raises the question of axiomatizing the operator algebras $A \subset B(H)$, as being the abstract Banach $*$ -algebras satisfying a number of extra conditions. This is an interesting question too, and we will be back to it later, with the answer that the only extra condition which is needed is the following compatibility formula, between the norm and the involution:

$$\|TT^*\| = \|T\|^2$$

As a third and final comment, Theorem 5.3 is something quite interesting, philosophically speaking, because it allows us to formulate the following fancy definition:

DEFINITION 5.4. *Given an arbitrary operator algebra $A \subset B(H)$, we write*

$$A = C(X)$$

and call the abstract object X a compact quantum space.

This might seem a bit abusive, but remember that we are mathematicians, not physicists, and so everything within our range of thoughts is possible. To be more precise, in correct and rigorous mathematical parlance, we can define the category of the compact quantum spaces to the category of the operator algebras $A \subset B(H)$, up to Banach $*$ -algebra isomorphism of such algebras, with the arrows reversed. And that's all.

When A is commutative, the space X considered above exists indeed, as shown by Theorem 5.3 above. In general, X is something rather abstract, and the idea can be simply that of studying of course A , but formulating our results in terms of X . For instance whenever we have a morphism $\Phi : A \rightarrow B$, we will write $A = C(X)$, $B = C(Y)$, and rather speak of the corresponding morphism $\phi : Y \rightarrow X$. And so on.

We will see later, after developing some more theory, that the above formalism has in fact certain limitations, and needs a fix. To be more precise, Definition 5.4 produces in fact more quantum spaces than actually “exist”, and this due to a subtle mathematical

phenomenon, called non-amenability. Thus, we will have to come up with a certain equivalence relation, in order to get rid of redundancies. We will be back to this.

Our purpose in what follows will be that of going beyond what can be done with the arbitrary operator algebras $A \subset B(H)$, by assuming that our algebra A is closed under a certain weaker topology on $B(H)$. As potential goals, we would like for instance our operator algebras $A \subset B(H)$ to be stable under polar decomposition, under taking spectral projections, and more generally, under measurable functional calculus.

At an even more ambitious level, we would like for instance to have a theorem stating that our commutative operator algebras are of the form $A = L^\infty(X)$, with X being a measured space. Indeed, with such a result in hand, we would be able to talk afterwards about the “quantum measured spaces”, simply by writing any operator algebra in our improved sense as $A = L^\infty(X)$, with X being such a quantum measured space.

All this sounds quite exciting, and getting into this program will be our next goal. For this purpose, we need a preliminary discussion on the notion of closedness inside $B(H)$. The result that we need, which is something elementary, is as follows:

PROPOSITION 5.5. *For a subalgebra $A \subset B(H)$, the following are equivalent:*

- (1) *A is closed under the weak operator topology, making each of the linear maps $T \rightarrow \langle Tx, y \rangle$ continuous.*
- (2) *A is closed under the strong operator topology, making each of the linear maps $T \rightarrow Tx$ continuous.*

In the case where these conditions are satisfied, A is closed under the norm topology.

PROOF. There are several statements here, the proof being as follows:

(1) It is clear from definitions that the norm topology is stronger than the strong operator topology, which is in turn stronger than the weak operator topology.

At the level of the subsets $S \subset B(H)$ which are closed things get reversed, in the sense that weakly closed implies strongly closed, which in turn implies norm closed. Thus, in order to establish the result, we are left with proving that for any algebra $A \subset B(H)$, strongly closed implies weakly closed.

(2) But this latter fact is something standard, which can be proved via an amplification trick. Consider the Hilbert space obtained by summing n times H with itself:

$$K = H \oplus \dots \oplus H$$

The operators over K can be regarded as being square matrices with entries in $B(H)$, and in particular, we have a representation $\pi : B(H) \rightarrow B(K)$, as follows:

$$\pi(T) = \begin{pmatrix} T & & \\ & \ddots & \\ & & T \end{pmatrix}$$

Assume now that we are given an operator $T \in \bar{A}$, with the bar denoting the weak closure. We have then, by using the Hahn-Banach theorem, for any $x \in K$:

$$\begin{aligned} T \in \bar{A} &\implies \pi(T) \in \overline{\pi(A)} \\ &\implies \pi(T)x \in \overline{\pi(A)x} \\ &\implies \pi(T)x \in \overline{\pi(A)x}^{\|\cdot\|} \end{aligned}$$

Now observe that the last formula tells us that for any $x = (x_1, \dots, x_n)$, and any $\varepsilon > 0$, we can find $S \in A$ such that the following holds, for any i :

$$\|Sx_i - Tx_i\| < \varepsilon$$

Thus T belongs to the strong operator closure of A , as desired. \square

Observe that in the above the terminology is a bit confusing, because the norm topology is stronger, and in fact considerably stronger, than the strong operator topology. As a solution to this problem, we agree to call the norm topology “strong”, and the weak and strong operator topologies “weak”, whenever these two topologies coincide.

With this convention made, the algebras $A \subset B(H)$ appearing in Proposition 5.5 are those which are weakly closed. Thus, we can now formulate:

DEFINITION 5.6. *A von Neumann algebra is an operator algebra*

$$A \subset B(H)$$

which is closed under the weak topology.

The von Neumann algebras will be our main objects of study, in the remainder of this chapter, and of this book. As basic examples, we have the algebra $B(H)$ itself, then the singly generated von Neumann algebras, $A = \langle T \rangle$ with $T \in B(H)$, and then the multiply generated von Neumann algebras, $A = \langle T_i \rangle$ with $T_i \in B(H)$.

We have as well a number of concrete and advanced examples coming from the operator theory developed in the previous chapters, and we will be back to this later.

For the moment, let us keep things simple, and build directly on Definition 5.6, by using basic functional analysis methods. We will need the following key result:

THEOREM 5.7. *For an operator algebra $A \subset B(H)$, we have*

$$A'' = \bar{A}$$

with A'' being the bicommutant inside $B(H)$, and \bar{A} being the weak closure.

PROOF. We can prove this by double inclusion, as follows:

“ \supset ” Since any operator commutes with the operators that it commutes with, we have a trivial inclusion $S \subset S''$, valid for any set $S \subset B(H)$. In particular, we have:

$$A \subset A''$$

Our claim now is that the algebra A'' is closed, with respect to the strong operator topology. Indeed, assuming that we have $T_i \rightarrow T$ in this topology, we have:

$$\begin{aligned} T_i \in A'' &\implies ST_i = T_iS, \forall S \in A' \\ &\implies ST = TS, \forall S \in A' \\ &\implies T \in A \end{aligned}$$

Thus our claim is proved, and together with Proposition 5.5, which allows us to pass from the strong to the weak operator topology, this gives the desired inclusion:

$$\bar{A} \subset A''$$

“ \subset ” Here we must prove that we have the following implication, valid for any $T \in B(H)$, with the bar denoting as usual the weak operator closure:

$$T \in A'' \implies T \in \bar{A}$$

For this purpose, we use the same amplification trick as in the proof of Proposition 5.5 above. Consider the Hilbert space obtained by summing n times H with itself:

$$K = H \oplus \dots \oplus H$$

The operators over K can be regarded as being square matrices with entries in $B(H)$, and in particular, we have a representation $\pi : B(H) \rightarrow B(K)$, as follows:

$$\pi(T) = \begin{pmatrix} T & & \\ & \ddots & \\ & & T \end{pmatrix}$$

The idea will be that of doing the computations in this representation. First, in this representation, the image of our algebra $A \subset B(H)$ is given by:

$$\pi(A) = \left\{ \begin{pmatrix} T & & \\ & \ddots & \\ & & T \end{pmatrix} \mid T \in A \right\}$$

We can compute the commutant of this image, exactly as in the usual scalar matrix case, and we obtain the following formula:

$$\pi(A)' = \left\{ \left(\begin{array}{ccc} S_{11} & \cdots & S_{1n} \\ \vdots & & \vdots \\ S_{n1} & \cdots & S_{nn} \end{array} \right) \mid S_{ij} \in A' \right\}$$

We conclude from this that, given an operator $T \in A''$ as above, we have:

$$\begin{pmatrix} T & & \\ & \ddots & \\ & & T \end{pmatrix} \in \pi(A)''$$

In other words, the conclusion of all this is that we have:

$$T \in A'' \implies \pi(T) \in \pi(A)''$$

Now given a vector $x \in K$, consider the orthogonal projection $P \in B(K)$ on the norm closure of the vector space $\pi(A)x \subset K$. Since the subspace $\pi(A)x \subset K$ is invariant under the action of $\pi(A)$, so is its norm closure inside K , and we obtain from this:

$$P \in \pi(A)'$$

By combining this with what we found above, we conclude that we have:

$$T \in A'' \implies \pi(T)P = P\pi(T)$$

Now since this holds for any $x \in K$, we conclude that any $T \in A''$ belongs to the strong operator closure of A . By using now Proposition 5.5, which allows us to pass from the strong to the weak operator closure, we conclude that we have:

$$A'' \subset \bar{A}$$

But this leads to the conclusion in the statement. □

Now getting back to the von Neumann algebras, from Definition 5.6, we have the following result, which is a reformulation of Theorem 5.7, by using this notion:

THEOREM 5.8. *For an operator algebra $A \subset B(H)$, the following are equivalent:*

- (1) *A is weakly closed, so it is a von Neumann algebra.*
- (2) *A equals its algebraic bicommutant A'' , taken inside $B(H)$.*

PROOF. This follows from the formula $A'' = \bar{A}$ from Theorem 5.7 above, along with the trivial fact that the commutants are automatically weakly closed. □

The above statement, called bicommutant theorem, and due to von Neumann, is quite interesting, philosophically speaking. Among others, it shows that the von Neumann algebras are exactly the commutants of the self-adjoint sets of operators:

PROPOSITION 5.9. *Given a subset $S \subset B(H)$ which is closed under $*$, the commutant*

$$A = S'$$

is a von Neumann algebra. Any von Neumann algebra appears in this way.

PROOF. We have two assertions here, the idea being as follows:

(1) Given $S \subset B(H)$ satisfying $S = S^*$, the commutant $A = S'$ satisfies $A = A^*$, and is also weakly closed. Thus, A is a von Neumann algebra. Note that this follows as well from the following “tricommutant formula”, which follows from Theorem 5.8:

$$S''' = S'$$

(2) Given a von Neumann algebra $A \subset B(H)$, we can take $S = A'$. Then S is closed under the involution, and we have $S' = A$, as desired. \square

As another interesting consequence, we have the following result:

PROPOSITION 5.10. *Given a von Neumann algebra $A \subset B(H)$, its center*

$$Z(A) = A \cap A'$$

regarded as an algebra $Z(A) \subset B(H)$, is a von Neumann algebra too.

PROOF. This follows from the fact that the commutants are weakly closed, that we know from the above, which shows that $A' \subset B(H)$ is a von Neumann algebra. Thus, the intersection $Z(A) = A \cap A'$ must be a von Neumann algebra too, as claimed. \square

We will see later on, throughout this book, that this latter result provides us with the key for the study of the von Neumann algebras, in general. Indeed, we will prove soon that any commutative von Neumann algebra must be of the form $L^\infty(X)$, with X being a measured space. Now in the case of a center algebra $Z(A)$, which is obviously commutative, this result applies, and provides us with a formula as follows:

$$Z(A) = L^\infty(X)$$

Starting from this, the idea will be then that of proving that the whole algebra A decomposes as a sum of “fibers”, called factors and denoted A_x , over the center space X , with this eventually providing us with a powerful decomposition result, as follows:

$$A = \int_X A_x dx$$

But more on this later, after developing the appropriate tools for this program.

Finally, as yet another consequence of the bicommutant theorem, this time of practical usefulness, for the presentation, this result allows us to write A'' for the weak closure of a $*$ -algebra $A \subset B(H)$. We will systematically use this convention, in what follows.

5b. Finite dimensions

In order to develop now some general theory for the von Neumann algebras, let us start by investigating the finite dimensional case. Here the ambient operator algebra is a usual matrix algebra, $B(H) = M_N(\mathbb{C})$, and any linear subspace $A \subset B(H)$ is automatically closed, for all 3 topologies from Proposition 5.5 above.

Thus, we are left with the question of investigating the $*$ -algebras of usual matrices $A \subset M_N(\mathbb{C})$. But this is a purely algebraic question, whose answer is as follows:

THEOREM 5.11. *The $*$ -algebras $A \subset M_N(\mathbb{C})$ are exactly the algebras of the form*

$$A = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

depending on parameters $k \in \mathbb{N}$ and $r_1, \dots, r_k \in \mathbb{N}$ satisfying

$$r_1 + \dots + r_k = N$$

embedded into $M_N(\mathbb{C})$ via the obvious block embedding, twisted by a unitary $U \in U_N$.

PROOF. We have two assertions to be proved, the idea being as follows:

(1) Given numbers $r_1, \dots, r_k \in \mathbb{N}$ satisfying $r_1 + \dots + r_k = N$, we have an obvious embedding of $*$ -algebras, via matrix blocks, as follows:

$$M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C}) \subset M_N(\mathbb{C})$$

In addition, we can twist this embedding by a unitary $U \in U_N$, as follows:

$$M \rightarrow U M U^*$$

Thus, we have proved one of the implications.

(2) In the other sense now, consider an arbitrary $*$ -algebra of the $N \times N$ matrices:

$$A \subset M_N(\mathbb{C})$$

Let us first look at the center of this algebra, which given by:

$$Z(A) = A \cap A'$$

It is elementary to prove that this center, as an algebra, is of the following form:

$$Z(A) \simeq \mathbb{C}^k$$

Consider now the standard basis $e_1, \dots, e_k \in \mathbb{C}^k$, and let $p_1, \dots, p_k \in Z(A)$ be the images of these vectors via the above identification. In other words, these elements $p_1, \dots, p_k \in A$ are central minimal projections, summing up to 1:

$$p_1 + \dots + p_k = 1$$

The idea is then that this partition of the unity will eventually lead to the block decomposition of A , as in the statement. We prove this in 4 steps, as follows:

Step 1. We first construct the matrix blocks, our claim here being that each of the following linear subspaces of A are non-unital $*$ -subalgebras of A :

$$A_i = p_i A p_i$$

But this is clear, with the fact that each A_i is closed under the various non-unital $*$ -subalgebra operations coming from the projection equations $p_i^2 = p_i = p_i^*$.

Step 2. We prove now that the above algebras $A_i \subset A$ are in a direct sum position, in the sense that we have a non-unital $*$ -algebra sum decomposition, as follows:

$$A = A_1 \oplus \dots \oplus A_k$$

As with any direct sum question, we have two things to be proved here. First, by using the formula $p_1 + \dots + p_k = 1$ and the projection equations $p_i^2 = p_i = p_i^*$, we conclude that we have the needed generation property, namely:

$$A_1 + \dots + A_k = A$$

As for the fact that the sum is indeed direct, this follows as well from the formula $p_1 + \dots + p_k = 1$, and from the projection equations $p_i^2 = p_i = p_i^*$.

Step 3. Our claim now, which will finish the proof, is that each of the $*$ -subalgebras $A_i = p_i A p_i$ constructed above is a full matrix algebra. To be more precise here, with $r_i = \text{rank}(p_i)$, our claim is that we have isomorphisms, as follows:

$$A_i \simeq M_{r_i}(\mathbb{C})$$

In order to prove this claim, recall that the projections $p_i \in A$ were chosen central and minimal. Thus, the center of each of the algebras A_i reduces to the scalars:

$$Z(A_i) = \mathbb{C}$$

But this shows, either via a direct computation, or via the bicommutant theorem, that each of the algebras A_i is a full matrix algebra, as claimed.

Step 4. We can now obtain the result, by putting together what we have. Indeed, by using the results from Step 2 and Step 3, we obtain an isomorphism as follows:

$$A \simeq M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

Moreover, a careful look at the isomorphisms established in Step 3 shows that at the global level, of the algebra A itself, the above isomorphism simply comes by twisting the following standard multimatrix embedding, discussed in the beginning of the proof, (1) above, by a certain unitary matrix $U \in U_N$:

$$M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C}) \subset M_N(\mathbb{C})$$

Now by putting everything together, we obtain the result. \square

In relation with the bicommutant theorem, we have the following result, which fully clarifies the situation, with a very explicit proof, in finite dimensions:

PROPOSITION 5.12. *Consider a $*$ -algebra $A \subset M_N(\mathbb{C})$, written as above:*

$$A = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

The commutant of this algebra is then, with respect with the block decomposition used,

$$A' = \mathbb{C} \oplus \dots \oplus \mathbb{C}$$

and by taking one more time the commutant we obtain A itself, $A = A''$.

PROOF. Let us decompose indeed our algebra A as in Theorem 5.11 above, as a direct sum of matrix algebras:

$$A = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

The center of each matrix algebra being reduced to the scalars, the commutant of this algebra is then as follows, with each copy of \mathbb{C} corresponding to a matrix block:

$$A' = \mathbb{C} \oplus \dots \oplus \mathbb{C}$$

By taking once again the commutant we obtain A itself, and we are done. \square

We will see later on, in chapters 9-12 below, that the above result, which is enlightening, can be generalized to the case of the arbitrary von Neumann algebras, by using some suitable integrals instead of direct sums. More on this later.

As another application of Theorem 5.11, clarifying this time the relation with operator theory, in finite dimensions, we have the following result:

THEOREM 5.13. *Given an operator $T \in B(H)$ in finite dimensions, $H = \mathbb{C}^N$, the von Neumann algebra $A = \langle T \rangle$ that it generates inside $B(H) = M_N(\mathbb{C})$ is*

$$A = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

with the sizes of the blocks $r_1, \dots, r_k \in \mathbb{N}$ coming from the spectral theory of the associated matrix $M \in M_N(\mathbb{C})$. In the normal case $TT^ = T^*T$, this decomposition comes from*

$$T = UDU^*$$

with $D \in M_N(\mathbb{C})$ diagonal, and with $U \in U_N$ unitary.

PROOF. This is routine, by using the linear algebra theory and spectral theory developed in chapter 1 above, for the usual matrices $M \in M_N(\mathbb{C})$. To be more precise:

(1) The fact that $A = \langle T \rangle$ decomposes into a direct sum of matrix algebras is something that we already know, coming from Theorem 5.11 above.

(2) By using standard linear algebra, we can compute the block sizes $r_1, \dots, r_k \in \mathbb{N}$, from the knowledge of the spectral theory of the associated matrix $M \in M_N(\mathbb{C})$.

(3) In the normal case, $TT^* = T^*T$, we can simply invoke the spectral theorem, and by suitably changing the basis, we are led to the conclusion in the statement. \square

5c. Commutative algebras

Let us get now to infinite dimensions, with Theorem 5.13 as our main source of inspiration. The same argument applies, provided that we are in the normal case, and we have the following result, summarizing our basic knowledge here:

THEOREM 5.14. *Given a bounded operator $T \in B(H)$ which is normal,*

$$TT^* = T^*T$$

the von Neumann algebra $A = \langle T \rangle$ that it generates inside $B(H)$ is

$$\langle T \rangle = L^\infty(\sigma(T))$$

with $\sigma(T)$ being its spectrum, formed of numbers $\lambda \in \mathbb{C}$ such that $T - \lambda$ is not invertible.

PROOF. The measurable functional calculus theorem for the normal operators tells us that we have a weakly continuous morphism of $*$ -algebras, as follows:

$$L^\infty(\sigma(T)) \rightarrow B(H)$$

$$f \rightarrow f(T)$$

Moreover, by the general properties of the measurable calculus, also established in chapter 3, this morphism is injective, and its image is the weakly closed algebra $\langle T \rangle$ generated by T, T^* . Thus, we obtain the isomorphism in the statement. \square

More generally now, along the same lines, we have the following result:

THEOREM 5.15. *Given operators $T_i \in B(H)$ which are normal, and which commute, the von Neumann algebra $A = \langle T_i \rangle$ that these operators generates inside $B(H)$ is*

$$\langle T_i \rangle = L^\infty(X)$$

with X being a certain measured space, associated to the family $\{T_i\}$.

PROOF. This is once again routine, by using the spectral theory for the families of commuting normal operators $T_i \in B(H)$ developed in chapter 3 above. \square

As an interesting abstract consequence of this, we have:

THEOREM 5.16. *The commutative von Neumann algebras are the algebras*

$$A = L^\infty(X)$$

with X being a measured space.

PROOF. We have two assertions to be proved, the idea being as follows:

(1) In one sense, we must prove that given a measured space X , we can realize the commutative algebra $A = L^\infty(X)$ as a von Neumann algebra, on a certain Hilbert space

H . But this is something that we already know, coming from the multiplicity operators $T_f(g) = fg$ discussed in chapter 2 above, the representation being as follows:

$$L^\infty(X) \subset B(L^2(X))$$

(2) In the other sense, given a commutative von Neumann algebra $A \subset B(H)$, we must construct a certain measured space X , and an identification $A = L^\infty(X)$. But this follows from Theorem 5.15, because we can write our von Neumann algebra as follows:

$$A = \langle T_i \rangle$$

To be more precise, A being commutative, any element $T \in A$ is normal. Thus, we can pick a basis $\{T_i\} \subset A$, and then we have $A = \langle T_i \rangle$ as above, with $T_i \in B(H)$ being commuting normal operators. Thus Theorem 5.15 applies, and gives the result. \square

The above result is not the end of the story with the commutative von Neumann algebras, because we still have to understand how a given algebra $A = L^\infty(X)$ can be represented as an operator algebra, $A \subset B(H)$, over the various Hilbert spaces H .

This latter question is something quite technical, and there are several approaches to it, one idea for instance being that of writing any commutative von Neumann algebra $A \subset B(H)$ in singly generated form, $A = \langle T \rangle$, and then using Theorem 5.14 above, which is something more explicit than Theorem 5.15.

In any case, the final result here, once again due to von Neumann, states that the commutative von Neumann algebras appear as follows, up to a certain multiplicity:

$$L^\infty(X) \subset B(L^2(X))$$

To be more precise, in one sense the idea is that of suitably amplifying the standard representation $L^\infty(X) \subset B(L^2(X))$, from the proof of Theorem 5.16 above. In the other sense, the idea is that of getting this by using advanced spectral theory methods.

As a more concrete consequence now of Theorem 5.16, we have:

THEOREM 5.17. *Given a von Neumann algebra $A \subset B(H)$, its center*

$$Z(A) = A \cap A'$$

can be written in the following form,

$$Z(A) = L^\infty(X)$$

with X being a certain measured space.

PROOF. We know from Proposition 5.10 above, which is something elementary, that if $A \subset B(H)$ is a von Neumann algebra, then the center $Z(A) \subset B(H)$ is a von Neumann algebra too. Thus Theorem 5.16 applies, and gives the result. \square

We will see later that this result provides us with the key for the study of the von Neumann algebras. The idea indeed, following some late work of von Neumann, and more specifically his “reduction theory” paper, will be that of proving that the whole algebra A decomposes as a sum of “fibers”, called factors and denoted A_x , over the center space X , with this eventually providing us with a powerful decomposition result, as follows:

$$A = \int_X A_x dx$$

But more on this later, after developing the appropriate tools for this program, which is something non-trivial. Among others, before getting into such things, we will have to study the von Neumann algebras A having trivial center, $Z(A) = \mathbb{C}$, called factors, and which include the fibers A_x in the above decomposition result. More on this later.

Summarizing, we already have many interesting open problems, which all look fundamental, coming from the von Neumann algebra theory that we developed so far. However, all this is quite non-trivial, and requires substantial preparations. We will be back to all this, and to a number of related questions, in the third part of the present book, chapters 9-12 below, when systematically doing functional analysis, of advanced type.

5d. Quantum spaces

In relation with our previous “quantum space” goals, Theorem 5.16 is something very interesting, philosophically speaking, suggesting us to formulate:

DEFINITION 5.18. *Given a von Neumann algebra $A \subset B(H)$, we write*

$$A = L^\infty(X)$$

and call X a quantum measured space.

This is of course something quite informal, at least at this stage of our presentation, and in the lack of any advanced tools in dealing with the von Neumann algebras. However, since talking about quantum spaces will be a main purpose of the present book, it is useful to have a discussion started on the subject, that we will fine-tune gradually later on.

As an example here, for the simplest noncommutative von Neumann algebra that we know, namely the usual matrix algebra $A = M_N(\mathbb{C})$, the formula that we want to write is as follows, with M_N being a certain mysterious quantum space:

$$M_N(\mathbb{C}) = L^\infty(M_N)$$

What can we say about this space M_N ? As a first observation, this is a finite space, with its cardinality being defined and computed as follows:

$$|M_N| = \dim_{\mathbb{C}} M_N(\mathbb{C}) = N^2$$

Now since this is the same as the cardinality of the set $\{1, \dots, N^2\}$, we are led to the conclusion that we should have a twisting result as follows, with the twisting operation $X \rightarrow X^\sigma$ being something that destroys the points, but keeps the cardinality:

$$M_N = \{1, \dots, N^2\}^\sigma$$

Abstract algebra can help here, and it is possible to prove that this is indeed the case, with a result stating that at the level of the associated algebras of L^∞ functions we have indeed a twisting result, as follows, with the algebraic twisting operation $A \rightarrow A^\sigma$ being something that destroys the commutativity of the multiplication:

$$M_N(\mathbb{C}) = L^\infty(1, \dots, N^2)^\sigma$$

From an analytic viewpoint, in relation with the writing $M_N(\mathbb{C}) = L^\infty(M_N)$ above, we would like of course to understand first what is the “uniform measure” on M_N , giving rise to the corresponding L^∞ functions. But this problem is obviously ill-posed, because the space M_N having no points, we cannot talk about measures on it.

However, we can talk about integration functionals with respect to such measures, and in answer to our question, the integration with respect to the uniform measure on M_N exists, and is given by the following formula:

$$\int_{M_N} A = tr(A)$$

To be more precise, on the left we have the integral of an arbitrary function on M_N , which according to our conventions, should be a usual matrix:

$$A \in L^\infty(M_N) = M_N(\mathbb{C})$$

As for the quantity on the right, the outcome of the computation, this can only be the trace of A . In addition, it is better to choose this trace to be normalized, by $tr(1) = 1$, and this in order for our measure on M_N to have mass 1, as it is ideal:

$$tr(A) = \frac{1}{N} Tr(A)$$

We can say even more about this. Indeed, since the traces of positive matrices are positive, we are led to the following formula, to be taken with the above conventions, which shows that the measure on M_N that we constructed is a probability measure:

$$A > 0 \implies \int_{M_N} A > 0$$

It is possible to have some further fun along these lines, and we will be back to this. Needless to say, all this won't bring anything new to this old mathematical field which is linear algebra. However, for more complicated von Neumann algebras, the philosophy from Definition 5.18, developed along the above lines, can be something very useful.

Before going further, let us record what we found, for future reference:

THEOREM 5.19. *The noncommutative measured space M_N given by*

$$M_N(\mathbb{C}) = L^\infty(M_N)$$

has cardinality N^2 , appears as a twist, in a purely algebraic sense,

$$M_N = \{1, \dots, N^2\}^\sigma$$

and is a probability space, its uniform integration being given by

$$\int_{M_N} A = \text{tr}(A)$$

where at right we have the normalized trace of matrices, $\text{tr} = \text{Tr}/N$.

PROOF. This is something half-informal, which basically follows from the above discussion, the details and missing details being as follows:

(1) In what regards the formula $|M_N| = N^2$, coming by computing the complex vector space dimension, as explained above, this is obviously something rock-solid.

(2) Regarding the twisting result now, as explained before, this should come by definition from a twisting result at the level of the algebras of functions. To be more precise, we would like to have a formula as follows, with the algebraic twisting operation $A \rightarrow A^\sigma$ being something that destroys the commutativity of the multiplication:

$$L^\infty(M_N) = L^\infty(1, \dots, N^2)^\sigma$$

In more familiar terms, of usual complex matrices on the left, and with a better-looking product of sets being used on the right, this formula reads:

$$M_N(\mathbb{C}) = L^\infty\left(\{1, \dots, N\} \times \{1, \dots, N\}\right)^\sigma$$

In order to establish this formula, consider the algebra on the right. As a complex vector space, this algebra has the standard basis $\{f_{ij}\}$ formed by the Dirac masses at the points (i, j) , and the multiplicative structure of this algebra is given by:

$$f_{ij}f_{kl} = \delta_{ij,kl}$$

Now let us twist this multiplication, according to the formula $e_{ij}e_{kl} = \delta_{jk}e_{il}$. We obtain in this way the usual combination formulae for the standard matrix units $e_{ij} : e_j \rightarrow e_i$ of the algebra $M_N(\mathbb{C})$, and so we have our twisting result, as claimed.

(3) In what regards the integration formula in the statement, with the conclusion that the underlying measure on M_N is a probability one, this is something that we fully explained before, and as for the result (1) above, it is something rock-solid.

(4) As a last technical comment, observe that the twisting operation performed in (2) destroys both the involution, and the trace of the algebra. This is something quite interesting, which cannot be fixed, and we will back to it, later on. \square

In order to advance now, based on the above result, the key point there is the construction and interpretation of the trace $tr : M_N(\mathbb{C}) \rightarrow \mathbb{C}$, as an integration functional. But this leads us into the following natural question: in the general context of Definition 5.18 above, what is the underlying integration functional $tr : A \rightarrow \mathbb{C}$?

This is a subtle question, and there are several possible answers here. For instance, we would like the integration functional to have the following property:

$$tr(ab) = tr(ba)$$

And the problem is that certain von Neumann algebras do not possess such traces. This is actually something quite advanced, that we do not know yet, but as usual, in what concerns the various counterexamples, we will take their existence, or even the vague news of their existence, for granted, and build our theory accordingly.

Summarizing, we must modify Definition 5.18, in the following way:

DEFINITION 5.20. *Given a von Neumann algebra $A \subset B(H)$, coming with a faithful positive unital form $tr : A \rightarrow \mathbb{C}$, satisfying the trace condition*

$$tr(ab) = tr(ba)$$

we formally write our von Neumann algebra as follows,

$$A = L^\infty(X)$$

and call X a noncommutative probability space. We also write the trace as

$$tr = \int_X$$

and call it integration with respect to the uniform measure on X .

Observe that we use here the mathematical notation for the expectation, as an integral. We will sometimes use the probabilistic notation \mathbb{E} too, depending on the context. In fact, the most often will we just use the original tr , which is the most practical.

At the level of examples, passed the classical spaces X , we know from Theorem 5.19 above that the space M_N , coming from the von Neumann algebra of matrices $M_N(\mathbb{C})$ via the formula $M_N(\mathbb{C}) = L^\infty(M_N)$, is a finite noncommutative probability space.

This raises the question of understanding what the finite noncommutative probability spaces are, in general. For this purpose, we need to examine the finite dimensional von Neumann algebras. And the result here, extending Theorem 5.11, is as follows:

THEOREM 5.21. *The finite dimensional von Neumann algebras $A \subset B(H)$ over an arbitrary Hilbert space H are exactly the direct sums of matrix algebras,*

$$A = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

embedded into $B(H)$ by using a partition of unity of $B(H)$ with rank 1 projections

$$1 = P_1 + \dots + P_k$$

with the “factors” $M_{r_i}(\mathbb{C})$ being each embedded into the algebra $P_i B(H) P_i$.

PROOF. This is standard, as in the case $A \subset M_N(\mathbb{C})$. Consider first the center of A , which is a finite dimensional commutative von Neumann algebra, of the following form:

$$Z(A) = \mathbb{C}^k$$

Now let P_i be the Dirac mass at $i \in \{1, \dots, k\}$. Then $P_i \in B(H)$ is an orthogonal projection, and these projections form a partition of unity, as follows:

$$1 = P_1 + \dots + P_k$$

With $A_i = P_i A P_i$, which is a non-unital $*$ -subalgebra of A , we have then a non-unital $*$ -algebra sum decomposition, as follows:

$$A = A_1 \oplus \dots \oplus A_k$$

On the other hand, it follows from the minimality of each of the projections $P_i \in Z(A)$ that we have unital $*$ -algebra isomorphisms, as follows:

$$A_i \simeq M_{r_i}(\mathbb{C})$$

Thus, we are led to the conclusion in the statement. \square

With the above result in hand, we have a partial answer to our question, in the sense that we can deduce what the finite noncommutative measured spaces are, in the sense of the old Definition 5.18. Indeed, we must solve the following equation:

$$L^\infty(X) = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

Now since the direct unions of sets correspond to the direct sums at the level of the associated algebras of functions, in the classical case, we can take the following formula as a definition for a direct union of sets, in the general, noncommutative case:

$$L^\infty(X_1 \sqcup \dots \sqcup X_r) = L^\infty(X_1) \oplus \dots \oplus L^\infty(X_r)$$

Now by remembering the definition of M_N , we are led to the conclusion that the solution to our noncommutative measured space equation above is as follows:

$$X = M_{r_1} \sqcup \dots \sqcup M_{r_k}$$

However, for fully solving the problem, in the spirit of Definition 5.20, we still have to discuss the traces on $L^\infty(X)$. Such a trace can only appear as a linear combination of the traces of the components, with certain weights $\lambda_i > 0$, summing up to 1:

$$tr = \lambda_1 tr_1 \oplus \dots \oplus \lambda_k tr_k$$

We are therefore done, or almost, the last question, which is a bit philosophical, being that of understanding whether, among these traces, there is a “canonical” one. The first thought here would go to the trace having equally distributed weights, as follows:

$$\lambda_i = \frac{1}{k}$$

This is motivated for instance by the fact that our trace $tr : A \rightarrow \mathbb{C}$ is uniquely determined by its restriction to the center of our algebra, $tr : Z(A) \rightarrow \mathbb{C}$, and since we have $Z(A) = \mathbb{C}^k$, we are therefore left with choosing a trace, as follows:

$$tr : \mathbb{C}^k \rightarrow \mathbb{C}$$

But common sense tells us that the good choice here can only be the standard trace, corresponding to the choice of the weights $\lambda_i = 1/k$ for the global trace, as above.

However, this is a not the correct choice. The point indeed is that, when using the pair (A, tr) in practice, in connection with various advanced questions, the different sized blocks $M_{r_i}(\mathbb{C})$ should correspond to different sized weights $\lambda_i > 0$. The solution to the problem, along with a summary of the above discussion, is as follows:

THEOREM 5.22. *The finite noncommutative measured spaces are the spaces*

$$X = M_{r_1} \sqcup \dots \sqcup M_{r_k}$$

according to the following formula, for the associated algebras of functions:

$$L^\infty(X) = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

The cardinality $|X|$ of such a space is the following number,

$$N = r_1^2 + \dots + r_k^2$$

and the possible traces are as follows, with $\lambda_i > 0$ summing up to 1:

$$tr = \lambda_1 tr_1 \oplus \dots \oplus \lambda_k tr_k$$

Among these traces, we have the canonical trace, appearing as

$$tr : L^\infty(X) \subset \mathcal{L}(L^\infty(X)) \rightarrow \mathbb{C}$$

via the left regular representation, having weights $\lambda_i = n_i^2/N$.

PROOF. We have many assertions here, which were for the most already discussed in the above, the details and missing details being as follows:

(1) The first assertion, regarding the structure of the spaces X and of the algebras $L^\infty(X)$ follows from Theorem 5.21, as explained above.

(2) The second assertion, regarding the cardinality, is clear from our convention for the cardinalities of noncommutative measured spaces, namely:

$$|X| = \dim_{\mathbb{C}} L^\infty(X)$$

(3) The third assertion, regarding the traces on $L^\infty(X)$, is clear from the fact that each matrix block has a unique trace, namely the normalized trace of matrices.

(4) Regarding now the last assertion, consider indeed the left regular representation of our algebra $A = L^\infty(X)$, which is given by the following formula:

$$\pi : A \subset \mathcal{L}(A)$$

$$\pi(a) : b \rightarrow ab$$

Observe that this representation is something purely algebraic, and this because we have not yet a trace, and hence a scalar product, on our algebra A . However, the algebra $\mathcal{L}(A)$ of linear operators $T : A \rightarrow A$ is isomorphic to a matrix algebra, and more specifically to the algebra $M_N(\mathbb{C})$, with $N = |X|$ being as before:

$$\mathcal{L}(A) \simeq M_N(\mathbb{C})$$

Thus, this algebra has a unique trace, as follows:

$$tr : \mathcal{L}(A) \rightarrow \mathbb{C}$$

Thus, we can compose this trace with the representation π , and we obtain a certain trace $tr : A \rightarrow \mathbb{C}$, that we can call “canonical”, for obvious reasons:

$$tr : A \subset \mathcal{L}(A) \rightarrow \mathbb{C}$$

We can compute the weights of this trace by using a multimatrix basis of A , formed by matrix units e_{ab}^i , with $i \in \{1, \dots, k\}$ and with $a, b \in \{1, \dots, r_i\}$, and we obtain:

$$\lambda_i = \frac{n_i^2}{N}$$

Thus, we are led to the conclusion in the statement, namely that we have indeed a canonical trace on $A = L^\infty(X)$, given by the formula there. \square

We will be back to quantum spaces on several occasions, in what follows, and in particular with further clarifications and extensions of the above discussion.

In fact, the present book is as much on operator algebras as it is on quantum spaces, and this because these two points of view are both very useful, and complementary to each other. We will get used to this, gradually, in what follows.

5e. Exercises

The theory in this chapter has been quite exciting, and we have already run into a number of difficult questions. As a first exercise on all this, we have:

EXERCISE 5.23. *Study the abstract Banach $*$ -algebras A having the property*

$$\|aa^*\| = \|a\|^2$$

for any $a \in A$, notably by proving that the commutative such algebras are those of type

$$A = C(X)$$

with X being a compact space.

Normally this is something quite subtle, called Gelfand theorem, and to be explained later on, in this book. However, we have seen in the above the proof of $A = C(X)$ for the operator algebras $A \subset B(H)$, and the proof in general is in fact not that much harder.

As a second exercise now, which is rather difficult as well, we have:

EXERCISE 5.24. *Find the bug with the abstract notation for the operator algebras*

$$A = C(X)$$

with X being a quantum space, by finding something which “obviously” goes wrong.

This is of course something quite subjective, and for the purposes of this exercise, you must convert yourself into a physicist. What you have to find is something defying common sense, such as two algebras $A \neq A'$ which “obviously” must correspond to the same quantum space X . As a hint, try tensor products, or group algebras.

In relation now with the bicommutant theorem, we have:

EXERCISE 5.25. *Find a simple proof for the bicommutant theorem, in finite dimensions.*

This is also something quite subjective, and try not to cheat. That is, not to convert the amplification proof that we have in general, by using matrix algebras everywhere, and nor by using the structure result for the finite dimensional algebras either.

As a somewhat bizarre theoretical exercise now, we have:

EXERCISE 5.26. *Given a commutative von Neumann algebra, $A = L^\infty(X)$, understand what happens when writing $A = C(Y)$, with Y being a compact space.*

To be more precise here, the problem is that of understanding, given a measured space X , what is the “compactification” Y making the formula $L^\infty(X) = C(Y)$ work. Of course, in case you’re stuck with this, don’t hesitate to look it up. And also don’t worry, about this, and the other exercises here: we will be back to all this, with more theory.

CHAPTER 6

Random matrices

6a. Random matrices

Our main results so far on the von Neumann algebras concern the finite dimensional case, where the algebra is of the form $A = \oplus_i M_{n_i}(\mathbb{C})$, and the commutative case, where the algebra is of the form $A = L^\infty(X)$. In order to unify these two constructions, the natural idea is that of looking at direct integrals of matrix algebras:

$$A = \int_X M_{n_x}(\mathbb{C}) dx$$

To be more precise, it is possible to introduce such algebras, which generalize both the finite dimensional ones, which correspond to the case $|X| < \infty$, and the commutative ones, which correspond to the case $n_x = 1$ for any $x \in X$. Moreover, it is possible to axiomatize this new class of algebras, via an abstract functional analysis definition, coupled with a new spectral theorem, providing a decomposition as above.

All this is quite tricky, and we will discuss this later. For the moment, let us discuss the “isotypic” case, where the fibers are all isomorphic. In this case our algebra is a random matrix algebra, which is something familiar in probability theory:

$$A = \int_X M_N(\mathbb{C}) dx$$

Although there is some functional analysis to be done with these algebras, the main questions regard the individual operators $T \in A$, called random matrices. Thus, we are basically back to good old operator theory. Let us begin our discussion with:

DEFINITION 6.1. *A random matrix algebra is a von Neumann algebra of the following type, with X being a probability space, and with $N \in \mathbb{N}$ being an integer:*

$$A = M_N(L^\infty(X))$$

In other words, A appears as a tensor product, as follows,

$$A = M_N(\mathbb{C}) \otimes L^\infty(X)$$

of a matrix algebra and a commutative von Neumann algebra.

As a first observation, our algebra can be written as well as follows, with this latter convention being quite standard in the probability literature:

$$A = L^\infty(X, M_N(\mathbb{C}))$$

Observe that this latter convention is closer to the fact that A appears as a direct integral of matrix algebras, with isotypic fibers, in the spirit of the discussion above:

$$A = \int_X M_N(\mathbb{C}) dx$$

In connection with the tensor product notation, which is often the most useful one for computations, we have as well the following possible writing, once again coming from the above second possible writing of A , and which is frequently used in probability:

$$A = L^\infty(X) \otimes M_N(\mathbb{C})$$

Importantly now, each random matrix algebra A is naturally endowed with a canonical von Neumann algebra trace $tr : A \rightarrow \mathbb{C}$, which appears as follows:

PROPOSITION 6.2. *Given a random matrix algebra $A = M_N(L^\infty(X))$, consider the linear form $tr : A \rightarrow \mathbb{C}$ given by:*

$$tr(T) = \frac{1}{N} \sum_{i=1}^N \int_X T_{ii}^x dx$$

In tensor product notation, $A = M_N(\mathbb{C}) \otimes L^\infty(X)$, we have then the formula

$$tr = \frac{1}{N} Tr \otimes \int_X$$

and this functional $tr : A \rightarrow \mathbb{C}$ is a faithful positive unital trace.

PROOF. The first assertion, regarding the tensor product writing of tr , is clear from definitions. As for the second assertion, regarding the various properties of tr , this follows from this, because these properties are stable under taking tensor products. \square

As before, there is a discussion here in connection with the other possible writings of A . With the probabilistic notation $A = L^\infty(X, M_N(\mathbb{C}))$, the trace appears as:

$$tr(T) = \int_X \frac{1}{N} Tr(T^x) dx$$

Also, with the probabilistic tensor notation $A = L^\infty(X) \otimes M_N(\mathbb{C})$, the trace appears exactly as in the second part of Proposition 6.2, with the order inverted:

$$tr = \int_X \otimes \frac{1}{N} Tr$$

As already mentioned, the main questions about the random matrix algebras are rather of operator theory nature, regarding the individual operators $T \in A$, called random

matrices. To be more precise, we are interested in computing the “laws” or “distributions” of such matrices, with the basic theory here being as follows:

THEOREM 6.3. *Any self-adjoint random matrix $T \in M_N(L^\infty(X))$ has a distribution, or law, meaning a real probability measure μ satisfying:*

$$\text{tr}(\varphi(T)) = \int_{\mathbb{R}} \varphi(x) d\mu(x)$$

The support of this distribution, or law μ is the spectrum $\sigma(T) \subset \mathbb{R}$.

PROOF. This result, which is the foundation for the random matrix theory, is clear from our spectral theory knowledge, from the previous chapters. \square

As a basic example, assume that our probability space is $X = \{.\}$, so that we are dealing with a usual scalar matrix, $T \in M_N(\mathbb{C})$. By changing the basis of \mathbb{C}^N , which won't affect any of our trace computations, we can assume that T is diagonal:

$$T \sim \begin{pmatrix} \lambda_1 & & \\ & \ddots & \\ & & \lambda_N \end{pmatrix}$$

Thus, the law of T is the average of the Dirac masses at the eigenvalues:

$$\mu = \frac{1}{N} (\delta_{\lambda_1} + \dots + \delta_{\lambda_N})$$

As a second example now, going in an opposite direction, assume that that matrix size is $N = 1$, so that we are dealing with a usual random variable $T \in L^\infty(X)$. In this case we obtain the usual law of T , because the equation to be satisfied by μ is:

$$\int_X \varphi(T) = \int_{\mathbb{R}} \varphi(x) d\mu(x)$$

In general, what we have is a kind of combination of these two situations. Things can get quite complicated here, and we will see explicit examples later on.

More generally, we can talk in the same way about laws of normal random matrices, as being complex measures supported on the spectrum $\sigma(T) \subset \mathbb{C}$:

THEOREM 6.4. *Any normal random matrix $T \in M_N(L^\infty(X))$ has a distribution, or law, meaning a complex probability measure μ satisfying:*

$$\text{tr}(\varphi(T)) = \int_{\mathbb{C}} \varphi(x) d\mu(x)$$

The support of this distribution, or law μ is the spectrum $\sigma(T) \subset \mathbb{C}$.

PROOF. Once again this is clear from our spectral theory knowledge, from the previous chapters, with the remark that the operation $\varphi \rightarrow \varphi(T)$ used in the statement should be taken in the general sense of functional calculus, with $z \rightarrow T$ and $\bar{z} \rightarrow T^*$. \square

The basic examples, with $X = \{.\}$ or with $N = 1$, are similar to those in the self-adjoint case, discussed above. With $X = \{.\}$ we have $T \in M_N(\mathbb{C})$, and we obtain once again the average of the Dirac masses at the eigenvalues, which are now complex, $\lambda_i \in \mathbb{C}$. With $N = 1$ we have $T \in L^\infty(X)$, complex random variable, and we obtain the usual law of T , which is a certain probability measure in the complex plane \mathbb{C} .

In the non-normal case the situation is quite tricky, and this even for the usual scalar matrices $T \in M_N(\mathbb{C})$, the simplest counterexample being as follows:

PROPOSITION 6.5. *The following matrix, which is not normal,*

$$T = \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix}$$

has no distribution $\mu \in \mathcal{P}(\mathbb{C})$ in the usual sense.

PROOF. We must prove that there is no measure $\mu \in \mathcal{P}(\mathbb{C})$ making hold the following formula, with $\varphi \rightarrow \varphi(T)$ being functional calculus, mapping $z \rightarrow T$ and $\bar{z} \rightarrow T^*$:

$$\text{tr}(\varphi(T)) = \int_{\mathbb{C}} \varphi(x) d\mu(x)$$

For this purpose, observe that we have the following computation:

$$\text{tr}(TT^*TT^*) = \text{tr}((TT^*)^2) = \text{tr} \begin{pmatrix} 1 & 0 \\ 0 & 0 \end{pmatrix} = \frac{1}{2}$$

On the other hand, we have as well the following computation:

$$\text{tr}(TTT^*T^*) = \text{tr}(T^2(T^*)^2) = \text{tr} \begin{pmatrix} 0 & 0 \\ 0 & 0 \end{pmatrix} = 0$$

Now since these numbers are different, we cannot obtain them by integrating $|z|^2$ with respect to a measure $\mu \in \mathcal{P}(\mathbb{C})$ as above, and so we have our counterexample. \square

We have in fact the following general result:

THEOREM 6.6. *Given an operator algebra $A \subset B(H)$ endowed with a faithful trace $\text{tr} : A \rightarrow \mathbb{C}$, any normal variable,*

$$TT^* = T^*T$$

has a “law”, which is by definition a complex probability measure $\mu \in \mathcal{P}(\mathbb{C})$ satisfying:

$$\text{tr}(T^k) = \int_{\mathbb{C}} z^k d\mu(z)$$

This law is unique, and is supported by the spectrum of the variable, $\sigma(T) \subset \mathbb{C}$. In the non-normal case, $TT^ \neq T^*T$, such a law does not exist.*

PROOF. We have two assertions here, the idea being as follows:

(1) In the normal case, $TT^* = T^*T$, the spectral theorem, or rather the subsequent continuous functional calculus theorem, tells us that we have:

$$\langle T \rangle = C(\sigma(T))$$

Thus the functional $f(T) \rightarrow \text{tr}(f(T))$ can be regarded as an integration functional on the algebra $C(\sigma(T))$, and by the Riesz theorem, this latter functional must come from a probability measure μ on the spectrum $\sigma(T)$, in the sense that we must have:

$$\text{tr}(f(T)) = \int_{\sigma(T)} f(z) d\mu(z)$$

We are therefore led to the conclusions in the statement, with the uniqueness assertion coming from the fact that the operators T^k , taken as usual with respect to colored integer exponents, $k = \circ \bullet \bullet \circ \dots$, generate the whole operator algebra $C(\sigma(T))$.

(2) In the non-normal case now, $TT^* \neq T^*T$, we must show that such a law does not exist. For this purpose, we can use a positivity trick, as follows:

$$\begin{aligned} & TT^* - T^*T \neq 0 \\ \implies & (TT^* - T^*T)^2 > 0 \\ \implies & TT^*TT^* - TT^*T^*T - T^*TTT^* + T^*TT^*T > 0 \\ \implies & \text{tr}(TT^*TT^* - TT^*T^*T - T^*TTT^* + T^*TT^*T) > 0 \\ \implies & \text{tr}(TT^*TT^* + T^*TT^*T) > \text{tr}(TT^*T^*T + T^*TTT^*) \\ \implies & \text{tr}(TT^*TT^*) > \text{tr}(TTT^*T^*) \end{aligned}$$

Now assuming that T has a law $\mu \in \mathcal{P}(\mathbb{C})$, in the sense that the moment formula in the statement holds, the above two different numbers would have to both appear by integrating $|z|^2$ with respect to this law μ , which is contradictory, as desired. \square

The main problem regarding the random matrices is that of computing the distribution of various classes of such matrices, coming in series:

$$T_N \in M_N(L^\infty(X))$$

The general strategy here is that of computing first the asymptotic law μ^0 , in the $N \rightarrow \infty$ limit, and then looking for the higher order terms as well, as to finally reach to a series in N^{-1} giving the law of T_N , as follows:

$$\mu_N = \mu^0 + N^{-1}\mu^1 + N^{-2}\mu^2 + \dots$$

As a basic example here, of particular interest are the random matrices having i.i.d. complex normal entries, under the constraint $T = T^*$. Here the asymptotic law μ^0 is the Wigner semicircle law on $[-2, 2]$. We will discuss this at the end of the present chapter, and then in chapter 8 below with more details, when doing free probability.

6b. Probability theory

In order to get started, let us set $N = 1$. Here our algebra is $A = L^\infty(X)$, an arbitrary commutative von Neumann algebra. The most interesting linear operators $T \in A$, that we will rather denote as complex functions $f : X \rightarrow \mathbb{C}$, and call random variables, as it is customary, are the normal variables, which are defined as follows:

DEFINITION 6.7. *A variable $f : X \rightarrow \mathbb{R}$ is called standard normal when its law is:*

$$g_1 = \frac{1}{\sqrt{2\pi}} e^{-x^2/2} dx$$

More generally, the normal law of parameter $t > 0$ is the following measure:

$$g_t = \frac{1}{\sqrt{2\pi t}} e^{-x^2/2t} dx$$

These are also called Gaussian distributions, with “g” standing for Gauss.

Observe that these laws have indeed mass 1, and this due to the Gauss formula, which can be established by using polar coordinates, as follows:

$$\begin{aligned} \left(\int_{\mathbb{R}} e^{-x^2} dx \right)^2 &= \int_{\mathbb{R}} \int_{\mathbb{R}} e^{-x^2-y^2} dx dy \\ &= \int_0^{2\pi} \int_0^\infty e^{-r^2} r dr dt \\ &= 2\pi \times \frac{1}{2} \\ &= \pi \end{aligned}$$

Let us start with some basic results regarding the normal laws. Regarding the variance, moments and Fourier transform, we have the following result:

PROPOSITION 6.8. *The normal law g_t with $t > 0$ has the following properties:*

- (1) *The variance of this law is its parameter:*

$$V = t$$

- (2) *The odd moments vanish, and the even moments are as follows, with the standard convention $k!! = (k-1)(k-3)(k-5)\dots$:*

$$M_k = t^{k/2} \times k!!$$

- (3) *The Fourier transform $F_f(x) = \mathbb{E}(e^{ixf})$ is given by:*

$$F(x) = e^{-tx^2/2}$$

Also, we have the convolution semigroup formula $g_s * g_t = g_{s+t}$, for any $s, t > 0$.

PROOF. We have four formulae to be proved, the idea being as follows:

(1) The normal law g_t being centered, its variance is the second moment, $V = M_2$. Thus the result follows from (2), proved below, which gives in particular:

$$M_2 = t^{2/2} \times 2!! = t$$

(2) We have indeed the following computation, by partial integration:

$$\begin{aligned} M_k &= \frac{1}{\sqrt{2\pi t}} \int_{\mathbb{R}} x^k e^{-x^2/2t} dx \\ &= \frac{1}{\sqrt{2\pi t}} \int_{\mathbb{R}} (tx^{k-1}) \left(-e^{-x^2/2t}\right)' dx \\ &= \frac{1}{\sqrt{2\pi t}} \int_{\mathbb{R}} t(k-1)x^{k-2} e^{-x^2/2t} dx \\ &= t(k-1) \times \frac{1}{\sqrt{2\pi t}} \int_{\mathbb{R}} x^{k-2} e^{-x^2/2t} dx \\ &= t(k-1)M_{k-2} \end{aligned}$$

As for the initial values, coming from (1), these are as follows:

$$M_0 = 1 \quad , \quad M_1 = 0$$

Thus, we obtain by recurrence the formula in the statement.

(3) The Fourier transform formula can be established as follows:

$$\begin{aligned} F(x) &= \frac{1}{\sqrt{2\pi t}} \int_{\mathbb{R}} e^{-y^2/2t+ixy} dy \\ &= \frac{1}{\sqrt{2\pi t}} \int_{\mathbb{R}} e^{-(y/\sqrt{2t}-\sqrt{t/2}ix)^2-tx^2/2} dy \\ &= \frac{1}{\sqrt{2\pi t}} \int_{\mathbb{R}} e^{-z^2-tx^2/2} \sqrt{2t} dz \\ &= \frac{1}{\sqrt{\pi}} e^{-tx^2/2} \int_{\mathbb{R}} e^{-z^2} dz \\ &= e^{-tx^2/2} \end{aligned}$$

(4) As for the last assertion, this follows from (3), because $\log F_{g_t}$ is linear in t . \square

We are now ready to establish the Central Limit Theorem (CLT), which is a key result, telling us why the normal laws appear a bit everywhere, in the real life:

THEOREM 6.9. *Given a sequence of real random variables $f_1, f_2, f_3, \dots \in L^\infty(X)$, which are i.i.d., centered, and with variance $t > 0$, we have*

$$\frac{1}{\sqrt{n}} \sum_{i=1}^n f_i \sim g_t$$

with $n \rightarrow \infty$, in moments.

PROOF. In terms of moments, the Fourier transform $F_f(x) = \mathbb{E}(e^{ixf})$ is given by:

$$\begin{aligned} F_f(x) &= \mathbb{E} \left(\sum_{k=0}^{\infty} \frac{(ixf)^k}{k!} \right) \\ &= \sum_{k=0}^{\infty} \frac{i^k M_k(f)}{k!} x^k \end{aligned}$$

Thus, the Fourier transform of the variable in the statement is:

$$\begin{aligned} F(x) &= \left[F_f \left(\frac{x}{\sqrt{n}} \right) \right]^n \\ &= \left[1 - \frac{tx^2}{2n} + O(n^{-2}) \right]^n \\ &\simeq \left[1 - \frac{tx^2}{2n} \right]^n \\ &\simeq e^{-tx^2/2} \end{aligned}$$

But this latter function being the Fourier transform of g_t , we obtain the result. \square

There are many other things that can be said about normal variables, but what we have so far, namely definition, basic properties and CLT is perfect as a starter pack.

Let us discuss now the “discrete” counterpart of the above results. The main result here will be the Poisson Limit Theorem (PLT), involving the Poisson laws:

DEFINITION 6.10. *The Poisson law of parameter 1 is the following measure,*

$$p_1 = \frac{1}{e} \sum_k \frac{\delta_k}{k!}$$

and the Poisson law of parameter $t > 0$ is the following measure,

$$p_t = e^{-t} \sum_k \frac{t^k}{k!} \delta_k$$

with the letter “ p ” standing for Poisson.

In analogy with the normal laws, the Poisson laws have the following basic properties:

PROPOSITION 6.11. *The Poisson law p_t with $t > 0$ has the following properties:*

- (1) *The variance is $V = t$.*
- (2) *The moments are $M_k = \sum_{\pi \in P(k)} t^{|\pi|}$, with $|\cdot|$ being the number of blocks.*
- (3) *The Fourier transform is $F(x) = \exp((e^{ix} - 1)t)$.*

Also, we have $p_s * p_t = p_{s+t}$, for any $s, t > 0$.

PROOF. We have four formulae to be proved, the idea being as follows:

(1) The variance is $V = M_2 - M_1^2$, and by using the formulae $M_1 = t$ and $M_2 = t + t^2$, coming from (2), proved below, we obtain as desired, $V = t$.

(2) We have indeed the following computation, using the binomial formula:

$$\begin{aligned}
 M_{k+1} &= e^{-t} \sum_s \frac{t^{s+1} (s+1)^{k+1}}{(s+1)!} \\
 &= e^{-t} \sum_s \frac{t^{s+1} s^k}{s!} \left(1 + \frac{1}{s}\right)^k \\
 &= e^{-t} \sum_s \frac{t^{s+1} s^k}{s!} \sum_r \binom{k}{r} s^{-r} \\
 &= \sum_r \binom{k}{r} \cdot e^{-t} \sum_s \frac{t^{s+1} s^{k-r}}{s!} \\
 &= t \sum_r \binom{k}{r} M_{k-r}
 \end{aligned}$$

The initial data being $M_0 = 1$, $M_1 = t$, we obtain the formula in the statement.

(3) The Fourier transform formula can be established as follows:

$$\begin{aligned}
 F_{p_t}(x) &= e^{-t} \sum_k \frac{t^k}{k!} F_{\delta_k}(x) \\
 &= e^{-t} \sum_k \frac{t^k}{k!} e^{ikx} \\
 &= e^{-t} \sum_k \frac{(e^{ix}t)^k}{k!} \\
 &= \exp(-t) \exp(e^{ix}t) \\
 &= \exp((e^{ix} - 1)t)
 \end{aligned}$$

(4) As for the last assertion, this follows from (3), because $\log F_{p_t}$ is linear in t . \square

Observe that at $t = 1$ we obtain the Bell numbers, as the moments of the Poisson law p_1 , with these numbers being impossible to compute explicitly. Thus, in a certain sense, discrete probability is more complicated than continuous probability.

We are now ready to establish the Poisson Limit Theorem (PLT), which is a key result, telling us why the normal laws appear a bit everywhere, in the real life:

THEOREM 6.12. *We have the following convergence, in moments,*

$$\left(\left(1 - \frac{t}{n} \right) \delta_0 + \frac{t}{n} \delta_1 \right)^{*n} \rightarrow p_t$$

for any $t > 0$.

PROOF. Let us denote by μ_n the measure under the convolution sign:

$$\mu_n = \left(1 - \frac{t}{n} \right) \delta_0 + \frac{t}{n} \delta_1$$

We have the following computation:

$$\begin{aligned} F_{\delta_r}(x) = e^{irx} &\implies F_{\mu_n}(x) = \left(1 - \frac{t}{n} \right) + \frac{t}{n} e^{ix} \\ &\implies F_{\mu_n^{*n}}(x) = \left(\left(1 - \frac{t}{n} \right) + \frac{t}{n} e^{ix} \right)^n \\ &\implies F_{\mu_n^{*n}}(x) = \left(1 + \frac{(e^{ix} - 1)t}{n} \right)^n \\ &\implies F(x) = \exp((e^{ix} - 1)t) \end{aligned}$$

Thus, we obtain the Fourier transform of p_t , as desired. \square

As a third and last topic from classical probability, let us discuss now the complex normal laws. To start with, we have the following definition:

DEFINITION 6.13. *The complex Gaussian law of parameter $t > 0$ is*

$$G_t = \text{law} \left(\frac{1}{\sqrt{2}}(a + ib) \right)$$

where a, b are independent, each following the law g_t .

As in the real case, these measures form convolution semigroups:

PROPOSITION 6.14. *The complex Gaussian laws have the property*

$$G_s * G_t = G_{s+t}$$

for any $s, t > 0$, and so they form a convolution semigroup.

PROOF. This follows indeed from the real result, namely $g_s * g_t = g_{s+t}$, established above, simply by taking real and imaginary parts. \square

We have the following complex analogue of the CLT:

THEOREM 6.15 (CCLT). *Given complex random variables $f_1, f_2, f_3, \dots \in L^\infty(X)$ which are i.i.d., centered, and with variance $t > 0$, we have, with $n \rightarrow \infty$, in moments,*

$$\frac{1}{\sqrt{n}} \sum_{i=1}^n f_i \sim G_t$$

where G_t is the complex Gaussian law of parameter t .

PROOF. This follows indeed from the real CLT, established above, simply by taking real and imaginary parts. \square

Regarding now the moments, the situation is more complicated than in the real case, because in order to have good results, we have to deal with both the complex variables, and their conjugates. Let us formulate the following definition:

DEFINITION 6.16. *The moments a complex variable $f \in L^\infty(X)$ are the numbers*

$$M_k = \mathbb{E}(f^k)$$

depending on colored integers $k = \circ \bullet \bullet \circ \dots$, with the conventions

$$f^\emptyset = 1 \quad , \quad f^\circ = f \quad , \quad f^\bullet = \bar{f}$$

and multiplicativity, in order to define the colored powers f^k .

Given a colored integer $k = \circ \bullet \bullet \circ \dots$, we say that a pairing $\pi \in P_2(k)$ is matching when it pairs $\circ - \bullet$ symbols. With these conventions, we have the following result:

THEOREM 6.17. *The moments of the complex normal law are the numbers*

$$M_k(G_t) = \sum_{\pi \in \mathcal{P}_2(k)} t^{|\pi|}$$

where $\mathcal{P}_2(k)$ are the matching pairings of $\{1, \dots, k\}$, and $|\cdot|$ is the number of blocks.

PROOF. This is something well-known, which can be done in several steps, as follows:

(1) We recall from the above that the moments of the real Gaussian law g_1 , with respect to integer exponents $k \in \mathbb{N}$, are the following numbers:

$$m_k = |P_2(k)|$$

Numerically, we have the following formula, explained as well in the above:

$$m_k = \begin{cases} k!! & (k \text{ even}) \\ 0 & (k \text{ odd}) \end{cases}$$

(2) We will show here that in what concerns the complex Gaussian law G_1 , similar results hold. Numerically, we will prove that we have the following formula, where a colored integer $k = \circ \bullet \bullet \circ \dots$ is called uniform when it contains the same number of \circ and \bullet , and where $|k| \in \mathbb{N}$ is the length of such a colored integer:

$$M_k = \begin{cases} (|k|/2)! & (k \text{ uniform}) \\ 0 & (k \text{ not uniform}) \end{cases}$$

Now since the matching partitions $\pi \in \mathcal{P}_2(k)$ are counted by exactly the same numbers, and this for trivial reasons, we will obtain the formula in the statement, namely:

$$M_k = |\mathcal{P}_2(k)|$$

(3) This was for the plan. In practice now, we must compute the moments, with respect to colored integer exponents $k = \circ \bullet \bullet \circ \dots$, of the variable in the statement:

$$c = \frac{1}{\sqrt{2}}(a + ib)$$

As a first observation, in the case where such an exponent $k = \circ \bullet \bullet \circ \dots$ is not uniform in \circ, \bullet , a rotation argument shows that the corresponding moment of c vanishes. To be more precise, the variable $c' = wc$ can be shown to be complex Gaussian too, for any $w \in \mathbb{C}$, and from $M_k(c) = M_k(c')$ we obtain $M_k(c) = 0$, in this case.

(4) In the uniform case now, where $k = \circ \bullet \bullet \circ \dots$ consists of p copies of \circ and p copies of \bullet , the corresponding moment can be computed as follows:

$$\begin{aligned} M_k &= \int (c\bar{c})^p \\ &= \frac{1}{2^p} \int (a^2 + b^2)^p \\ &= \frac{1}{2^p} \sum_s \binom{p}{s} \int a^{2s} \int b^{2p-2s} \\ &= \frac{1}{2^p} \sum_s \binom{p}{s} (2s)!! (2p-2s)!! \\ &= \frac{1}{2^p} \sum_s \frac{p!}{s!(p-s)!} \cdot \frac{(2s)!}{2^s s!} \cdot \frac{(2p-2s)!}{2^{p-s}(p-s)!} \\ &= \frac{p!}{4^p} \sum_s \binom{2s}{s} \binom{2p-2s}{p-s} \end{aligned}$$

(5) In order to finish now the computation, let us recall that we have the following formula, coming from the generalized binomial formula, or from the Taylor formula:

$$\frac{1}{\sqrt{1+t}} = \sum_{k=0}^{\infty} \binom{2k}{k} \left(\frac{-t}{4}\right)^k$$

By taking the square of this series, we obtain the following formula:

$$\begin{aligned} \frac{1}{1+t} &= \sum_{ks} \binom{2k}{k} \binom{2s}{s} \left(\frac{-t}{4}\right)^{k+s} \\ &= \sum_p \left(\frac{-t}{4}\right)^p \sum_s \binom{2s}{s} \binom{2p-2s}{p-s} \end{aligned}$$

Now by looking at the coefficient of t^p on both sides, we conclude that the sum on the right equals 4^p . Thus, we can finish the moment computation in (4), as follows:

$$M_p = \frac{p!}{4^p} \times 4^p = p!$$

(6) As a conclusion, if we denote by $|k|$ the length of a colored integer $k = \circ \bullet \bullet \circ \dots$, the moments of the variable c in the statement are given by:

$$M_k = \begin{cases} (|k|/2)! & (k \text{ uniform}) \\ 0 & (k \text{ not uniform}) \end{cases}$$

On the other hand, the numbers $|\mathcal{P}_2(k)|$ are given by exactly the same formula. Indeed, in order to have matching pairings of k , our exponent $k = \circ \bullet \bullet \circ \dots$ must be uniform, consisting of p copies of \circ and p copies of \bullet , with $p = |k|/2$. But then the matching pairings of k correspond to the permutations of the \bullet symbols, as to be matched with \circ symbols, and so we have $p!$ such matching pairings. Thus, we have the same formula as for the moments of c , and we are led to the conclusion in the statement. \square

This was for the basic probability theory, which in a certain sense is advanced operator theory, inside the commutative von Neumann algebras, $A = L^\infty(X)$. We will be back to this, with some further probability limiting theorems, in chapter 8 below.

6c. Wigner matrices

Let us exit now the classical world, of the von Neumann algebras $A = L^\infty(X)$, and do as promised some random matrix theory. We recall that a random matrix algebra is a von Neumann algebra of type $A = M_N(L^\infty(X))$, and that we are interested here in the computation of the laws of the operators $T \in A$, called random matrices.

Regarding the precise classes of matrices that we are interested in, first we have the complex Gaussian matrices, which are constructed as follows:

DEFINITION 6.18. *A complex Gaussian matrix is a random matrix of type*

$$Z \in M_N(L^\infty(X))$$

which has i.i.d. complex normal entries.

We will be interested as well in the Wigner random matrices, which are the self-adjoint versions of the above Gaussian matrices, constructed as follows:

DEFINITION 6.19. *A Wigner matrix is a random matrix of type*

$$Z \in M_N(L^\infty(X))$$

which has i.i.d. complex normal entries, up to the following constraint:

$$Z = Z^*$$

In other words, a Wigner matrix must be as follows, with f_i being real normal variables, g_{ij} being complex normal variables, and all these variables being independent:

$$W = \begin{pmatrix} f_1 & g_{12} & \cdots & \cdots & g_{1N} \\ \bar{g}_{12} & f_2 & \ddots & & \vdots \\ \vdots & \ddots & \ddots & \ddots & \vdots \\ \vdots & & \ddots & f_{N-1} & g_{N-1,N} \\ \bar{g}_{1N} & \cdots & \cdots & \bar{g}_{N-1,N} & f_N \end{pmatrix}$$

Finally, we will be interested as well in the complex Wishart random matrices, which are the positive versions of these matrices, which are constructed as follows:

DEFINITION 6.20. *A complex Wishart matrix is a random matrix of type*

$$Z = YY^* \in M_N(L^\infty(X))$$

with Y being a complex Gaussian matrix.

Summarizing, we have a whole menagerie of random matrices, and the list can be of course continued, the above 3 classes being just the main ones. See [57], [96].

In order to compute the asymptotic laws of the above matrices, we will use the moment method. We have the following result, which will be our main tool here:

THEOREM 6.21. *Given independent variables X_i , each following the complex normal law G_t , with $t > 0$ being a fixed parameter, we have the Wick formula*

$$\mathbb{E} (X_{i_1}^{k_1} \cdots X_{i_s}^{k_s}) = t^{s/2} \# \left\{ \pi \in \mathcal{P}_2(k) \mid \pi \leq \ker(i) \right\}$$

where $k = k_1 \cdots k_s$ and $i = i_1 \cdots i_s$, for the joint moments of these variables.

PROOF. This is something well-known, and the basis for all possible computations with complex normal variables, which can be proved in two steps, as follows:

(1) Let us first discuss the case where we have a single complex normal variable X , which amounts in taking $X_i = X$ for any i in the formula in the statement. What we have to compute here are the moments of X , with respect to colored integer exponents $k = \circ \bullet \bullet \circ \dots$, and the formula in the statement tells us that these moments must be:

$$\mathbb{E}(X^k) = t^{|k|/2} |\mathcal{P}_2(k)|$$

But this is something that we know well from the above, the idea being that at $t = 1$ this follows by doing some combinatorics and calculus, in analogy with the combinatorics and calculus from the real case, where the moment formula is identical, save for the matching pairings \mathcal{P}_2 being replaced by the usual pairings P_2 , and then that the general case $t > 0$ follows from this, by rescaling. Thus, we are done with this case.

(2) In general now, the point is that we obtain the formula in the statement. Indeed, when expanding the product $X_{i_1}^{k_1} \dots X_{i_s}^{k_s}$ and rearranging the terms, we are left with doing a number of computations as in (1), and then making the product of the expectations that we found. But this amounts precisely in counting the partitions in the statement, with the condition $\pi \leq \ker(i)$ there standing precisely for the fact that we are doing the various type (1) computations independently, and then making the product. \square

Now by getting back to the Gaussian matrices, we have the following result:

THEOREM 6.22. *Given a sequence of Gaussian random matrices*

$$Z_N \in M_N(L^\infty(X))$$

which by definition have i.i.d. centered complex normal entries, we have

$$\frac{1}{N} \int_X \text{Tr}(Z_N^k) \simeq C_k$$

for any colored integer $k = \circ \bullet \bullet \circ \dots$, in the $N \rightarrow \infty$ limit, up to a normalization, where

$$C_k = |\mathcal{NC}_2(k)|$$

with \mathcal{NC}_2 standing for noncrossing matching pairings.

PROOF. We can compute the moments in a straightforward way, by expanding and using the Wick formula. In the $N \rightarrow \infty$ limit the combinatorics simplifies, and we are led, up to some normalizations, to the numbers in the statement.

To be more precise, the computation of the moments is as follows:

$$\begin{aligned}
M_k(Z_N) &= \frac{1}{N} \int_X \text{Tr}(Z_N^k) \\
&= \frac{1}{N} \int_X \sum_{i_1=1}^N \cdots \sum_{i_s=1}^N Z_{i_1 i_2}^{k_1} Z_{i_2 i_3}^{k_2} \cdots Z_{i_s i_1}^{k_s} \\
&= \frac{t^{s/2}}{N} \sum_{i_1=1}^N \cdots \sum_{i_s=1}^N \# \left\{ \pi \in \mathcal{P}_2(k) \mid \pi \leq \ker(i_1 i_2, i_2 i_3, \dots, i_s i_1) \right\}
\end{aligned}$$

The point now is that with $N \rightarrow \infty$ the combinatorics simplifies, and we see that only the noncrossing partitions will contribute to the multiple sum on the right, with each of them contributing precisely with a N factor. In other words, we have:

$$\begin{aligned}
M_k(Z_N) &= \frac{t^{s/2}}{N} \sum_{i_1=1}^N \cdots \sum_{i_s=1}^N \# \left\{ \pi \in \mathcal{P}_2(k) \mid \pi \leq \ker(i_1 i_2, i_2 i_3, \dots, i_s i_1) \right\} \\
&= t^{s/2} \sum_{\pi \in \mathcal{P}_2(k)} \frac{1}{N} \# \left\{ i \in \{1, \dots, N\}^s \mid \pi \leq \ker(i_1 i_2, i_2 i_3, \dots, i_s i_1) \right\} \\
&\simeq t^{s/2} \sum_{\pi \in \mathcal{P}_2(k)} \delta_{\pi \in NC_2(k)} \\
&= t^{s/2} |NC_2(k)|
\end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

Regarding now the Wigner matrices, we have here the following result, coming as a consequence of Theorem 6.22 above, via some simple algebraic manipulations:

THEOREM 6.23. *Given a sequence of Wigner random matrices*

$$Z_N \in M_N(L^\infty(X))$$

which by definition have i.i.d. centered complex normal entries, up to $Z_N = Z_N^$, we have*

$$\frac{1}{N} \int_X \text{Tr}(Z_N^k) \simeq C_{k/2}$$

for any integer $k \in \mathbb{N}$, in the $N \rightarrow \infty$ limit, up to a normalization, where

$$C_{k/2} = |NC_2(k)|$$

with NC_2 standing as usual for noncrossing pairings.

PROOF. We know from Theorem 6.22 that for Gaussian matrices $Y_N \in M_N(L^\infty(X))$, we have the following formula, valid for any colored integer $k = \circ \bullet \bullet \circ \dots$, in the $N \rightarrow \infty$

limit, up to a normalization, with \mathcal{NC}_2 standing for noncrossing matching pairings:

$$\frac{1}{N} \int_X \text{Tr}(Y_N^k) \simeq |\mathcal{NC}_2(k)|$$

By doing some combinatorics, we deduce from this that we have the following formula for the moments of the matrices $\text{Re}(Y_N)$, with respect to usual exponents, $k \in \mathbb{N}$:

$$\frac{1}{N} \int_X \text{Tr}(\text{Re}(Y_N)^k) \simeq |\mathcal{NC}_2(k)|$$

Now since the matrices $Z_N = \text{Re}(Y_N)$ are of Wigner type, this gives the result. \square

Summarizing, all this brings us into counting noncrossing pairings. So, let us start with some preliminaries here. We first have the following well-known result:

THEOREM 6.24. *The Catalan numbers, which are by definition given by*

$$C_k = |\mathcal{NC}_2(2k)|$$

satisfy the following recurrence formula,

$$C_{k+1} = \sum_{a+b=k} C_a C_b$$

their generating series $f(z) = \sum_{k \geq 0} C_k z^k$ satisfies the equation

$$z f^2 - f + 1 = 0$$

and is given by the following explicit formula,

$$f(z) = \frac{1 - \sqrt{1 - 4z}}{2z}$$

and we have the following explicit formula for these numbers:

$$C_k = \frac{1}{k+1} \binom{2k}{k}$$

PROOF. We must count the noncrossing pairings of $\{1, \dots, 2k\}$. Now observe that such a pairing appears by pairing 1 to an odd number, $2a+1$, and then inserting a noncrossing pairing of $\{2, \dots, 2a\}$, and a noncrossing pairing of $\{2a+2, \dots, 2l\}$. We conclude that we have the following recurrence formula for the Catalan numbers:

$$C_k = \sum_{a+b=k-1} C_a C_b$$

Consider now generating series of the Catalan numbers:

$$f(z) = \sum_{k \geq 0} C_k z^k$$

In terms of this generating series, the above recurrence gives:

$$\begin{aligned}
 z f^2 &= \sum_{a,b \geq 0} C_a C_b z^{a+b+1} \\
 &= \sum_{k \geq 1} \sum_{a+b=k-1} C_a C_b z^k \\
 &= \sum_{k \geq 1} C_k z^k \\
 &= f - 1
 \end{aligned}$$

Thus f satisfies the following degree 2 equation:

$$z f^2 - f + 1 = 0$$

By solving this equation, and choosing the solution which is bounded at $z = 0$, we obtain the following formula:

$$f(z) = \frac{1 - \sqrt{1 - 4z}}{2z}$$

By using now the Taylor formula for \sqrt{x} , we obtain the following formula:

$$f(z) = \sum_{k \geq 0} \frac{1}{k+1} \binom{2k}{k} z^k$$

It follows that the Catalan numbers are given by:

$$C_k = \frac{1}{k+1} \binom{2k}{k}$$

Thus, we are led to the conclusion in the statement. □

In order to recapture now the measure from its moments, we can use:

PROPOSITION 6.25. *The Catalan numbers are the even moments of*

$$\gamma_1 = \frac{1}{2\pi} \sqrt{4 - x^2} dx$$

called standard semicircle law. As for the odd moments of γ_1 , these all vanish.

PROOF. The even moments of the Wigner law can be computed with the change of variable $x = 2 \cos t$, and we are led to the following formula:

$$\begin{aligned}
M_{2k} &= \frac{1}{\pi} \int_0^2 \sqrt{4-x^2} x^{2k} dx \\
&= \frac{1}{\pi} \int_0^{\pi/2} \sqrt{4-4\cos^2 t} (2\cos t)^{2k} 2\sin t dt \\
&= \frac{4^{k+1}}{\pi} \int_0^{\pi/2} \cos^{2k} t \sin^2 t dt \\
&= \frac{4^{k+1}}{\pi} \cdot \frac{\pi}{2} \cdot \frac{(2k)!!2!!}{(2k+3)!!} \\
&= 2 \cdot 4^k \cdot \frac{(2k)!/2^k k!}{2^{k+1}(k+1)!} \\
&= C_k
\end{aligned}$$

As for the odd moments, these all vanish, because the density of γ_1 is an even function. Thus, we are led to the conclusion in the statement. \square

More generally, we have the following result, involving a parameter $t > 0$:

PROPOSITION 6.26. *The Catalan numbers are the even moments of*

$$\gamma_t = \frac{1}{2\pi t} \sqrt{4t^2 - x^2} dx$$

called standard semicircle law. As for the odd moments of γ_t , these all vanish.

PROOF. This follows indeed from Proposition 6.25, with a change of variables. \square

Now by putting everything together, we obtain the Wigner theorem, as follows:

THEOREM 6.27. *Given a sequence of Wigner random matrices*

$$Z_N \in M_N(L^\infty(X))$$

which by definition have i.i.d. complex normal entries, up to $Z_N = Z_N^$, we have*

$$Z_N \sim \gamma_t$$

in the $N \rightarrow \infty$ limit, up to a normalization, where

$$\gamma_t = \frac{1}{2\pi t} \sqrt{4t^2 - x^2} dx$$

is the Wigner standard semicircle law.

PROOF. This follows by combining Theorem 6.23 and Proposition 6.26 above. \square

6d. Wishart matrices

Let us discuss now the Wishart matrices, which are the positive analogues of the Wigner matrices. Quite surprisingly, the computation here leads to the Catalan numbers, but not in the same way as for the Wigner matrices, the result being as follows:

THEOREM 6.28. *Given a sequence of complex Wishart random matrices*

$$W_N = \frac{1}{N} \cdot Y_N Y_N^* \in M_N(L^\infty(X))$$

with Y_N being $N \times N$ complex Gaussian of parameter 1, we have

$$\frac{1}{N} \int_X \text{Tr}(W_N^k) \simeq C_k$$

for any exponent $k \in \mathbb{N}$, in the $N \rightarrow \infty$ limit.

PROOF. There are several possible proofs for this result, as follows:

(1) A first method is by using the result that we have from the above, for the Gaussian matrices. Indeed, we know from there that we have the following formula, for any colored integer $K = \circ \bullet \bullet \circ \dots$, in the $N \rightarrow \infty$ limit, up to a normalization:

$$\frac{1}{N} \int_X \text{Tr}(Y_N^K) \simeq |\mathcal{NC}_2(K)|$$

Now with the choice $K = \circ \bullet \circ \bullet \dots$, alternating word of length $2k$, with $k \in \mathbb{N}$, for our exponent, this formula becomes, as before up to a normalization factor:

$$\frac{1}{N} \int_X \text{Tr}((Y_N Y_N^*)^k) \simeq |\mathcal{NC}_2(K)|$$

In terms of the Wishart matrix $W_N = \frac{1}{N} \cdot Y_N Y_N^*$ in the statement, we therefore obtain the following moment formula, once again up to a normalization factor:

$$\frac{1}{N} \int_X \text{Tr}(W_N^k) \simeq |\mathcal{NC}_2(K)|$$

The point now is that, by doing some combinatorics, we have, up to a certain factor:

$$|\mathcal{NC}_2(K)| = |\mathcal{NC}_2(2k)| = C_k$$

Thus, up to a normalization, we are led to the formula in the statement, namely:

$$\frac{1}{N} \int_X \text{Tr}(W_N^k) \simeq C_k$$

(2) A second method, that we will explain now as well, is by proving the result directly, starting from definitions. The matrix entries of our Wishart matrix are given by:

$$W_{ij} = \frac{1}{N} \sum_{r=1}^N Y_{ir} \bar{Y}_{jr}$$

Thus, the normalized traces of powers of W are given by the following formula:

$$\begin{aligned} \text{tr}(W^k) &= \frac{1}{N^{k+1}} \sum_{r_1, \dots, r_k=1}^N \prod_{s=1}^k W_{r_s, r_{s+1}} \\ &= \frac{1}{N^{k+1}} \sum_{r_1, \dots, r_k=1}^N \prod_{s=1}^k \sum_{p_1, \dots, p_k=1}^N Y_{r_s, p_s} \bar{Y}_{r_{s+1}, p_s} \end{aligned}$$

After interchanging the product with the last sum, the average of the general term can be computed by using the Wick rule, which gives:

$$\mathbb{E} \left(\prod_{s=1}^k Y_{r_s, p_s} \bar{Y}_{r_{s+1}, p_s} \right) = \sum_{\pi \in S_k} \prod_{s=1}^k \delta_{r_s, r_{\pi(s)+1}} \delta_{p_s, p_{\pi(s)}}$$

Now let $\gamma \in S_k$ be the full cycle, which is by definition the following permutation:

$$\gamma = (1 \ 2 \ \dots \ k)^{-1}$$

The general factor in the product computed above is then 1 precisely when following two conditions are simultaneously satisfied:

$$\begin{aligned} \gamma^{-1}\pi &\leq \ker r \\ \pi &\leq \ker p \end{aligned}$$

Counting the number of free parameters in our expectation formula, we obtain:

$$\begin{aligned} (\mathbb{E} \circ \text{tr})(W^k) &= \frac{1}{N^{k+1}} \sum_{\pi \in S_k} N^{|\pi| + |\gamma^{-1}\pi|} \\ &= \sum_{\pi \in S_k} N^{|\pi| + |\gamma^{-1}\pi| - k - 1} \end{aligned}$$

The point now is that the last exponent is well-known to be ≤ 0 , with equality precisely when the permutation $\pi \in S_k$ is geodesic, which in practice means that π must come from a noncrossing partition. Thus we obtain, in the $N \rightarrow \infty$ limit:

$$(\mathbb{E} \circ \text{tr})(W^k) \simeq C_k$$

Thus, we are led to the conclusion in the statement. \square

As a consequence of the above result, we have a new look on the Catalan numbers, which is more adapted to our present Wishart matrix considerations, as follows:

PROPOSITION 6.29. *The Catalan numbers $C_k = |NC_2(2k)|$ appear as well as*

$$C_k = |NC(k)|$$

where $NC(k)$ is the set of all noncrossing partitions of $\{1, \dots, k\}$.

PROOF. This follows indeed from the proof of Theorem 6.28 above. Observe that we obtain as well a formula in terms of matching pairings of alternating colored integers. \square

The direct explanation for the above formula, relating noncrossing partitions and pairings, comes from the following result, which is very useful, and good to know:

PROPOSITION 6.30. *We have a bijection between noncrossing partitions and pairings*

$$NC(k) \simeq NC_2(2k)$$

which is constructed as follows:

- (1) *The application $NC(k) \rightarrow NC_2(2k)$ is the “fattening” one, obtained by doubling all the legs, and doubling all the strings as well.*
- (2) *Its inverse $NC_2(2k) \rightarrow NC(k)$ is the “shrinking” application, obtained by collapsing pairs of consecutive neighbors.*

PROOF. The fact that the two operations in the statement are indeed inverse to each other is clear, by computing the corresponding two compositions, with the remark that the construction of the fattening operation requires the partitions to be noncrossing. \square

Getting back now to probability, we are led to the question of finding the law having the Catalan numbers as moments. The result here is as follows:

PROPOSITION 6.31. *The real measure having the Catalan numbers as moments is*

$$\pi_1 = \frac{1}{2\pi} \sqrt{4x^{-1} - 1} dx$$

called *Marchenko-Pastur law of parameter 1*.

PROOF. The moments of the law π_1 in the statement can be computed with the change of variable $x = 4 \cos^2 t$, as follows:

$$\begin{aligned} M_k &= \frac{1}{2\pi} \int_0^4 \sqrt{4x^{-1} - 1} x^k dx \\ &= \frac{1}{2\pi} \int_0^{\pi/2} \frac{\sin t}{\cos t} \cdot (4 \cos^2 t)^k \cdot 2 \cos t \sin t dt \\ &= \frac{4^{k+1}}{\pi} \int_0^{\pi/2} \cos^{2k} t \sin^2 t dt \\ &= \frac{4^{k+1}}{\pi} \cdot \frac{\pi}{2} \cdot \frac{(2k)!!2!!}{(2k+3)!!} \\ &= 2 \cdot 4^k \cdot \frac{(2k)!/2^k k!}{2^{k+1}(k+1)!} \\ &= C_k \end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

Now back to the Wishart matrices, we are led to the following result:

THEOREM 6.32. *Given a sequence of complex Wishart random matrices*

$$Z_N = \frac{1}{N} Y_N Y_N^* \in M_N(L^\infty(X))$$

with Y_N being $N \times N$ complex Gaussian of parameter 1, we have

$$Z_N \sim \frac{1}{2\pi} \sqrt{4x^{-1} - 1} dx$$

with $N \rightarrow \infty$, with the limiting measure being the Marchenko-Pastur law π_1 .

PROOF. This follows indeed from Theorem 6.28 and Proposition 6.31. \square

Let us discuss as well the parametric version of the above results. Things here are quite technical, and the result here, generalizing Theorem 6.32, is as follows:

THEOREM 6.33. *Given a sequence of general complex Wishart random matrices*

$$Z_N = \frac{1}{M} Y_N Y_N^* \in M_N(L^\infty(X))$$

with Y_N being $N \times M$ complex Gaussian of parameter 1, we have

$$Z_N \sim \max(1-t, 0) \delta_0 + \frac{\sqrt{4t - (x-1-t)^2}}{2\pi x} dx$$

with $M = tN \rightarrow \infty$, with the limiting measure being the Marchenko-Pastur law π_t .

PROOF. This follows once again by using the moment method, the limiting moments being as follows, after doing the combinatorics:

$$\lim_{N \rightarrow \infty} M_k = \sum_{\pi \in NC(k)} t^{|\pi|}$$

But these numbers are the moments of the Marchenko-Pastur law π_t , which in addition has the density given by the formula in the statement, and we are done. \square

As a conclusion to all this, we have 4 main laws in what we have been doing so far, namely the Gaussian laws g_t , the Poisson laws p_t , the Wigner laws γ_t and the Marchenko-Pastur laws π_t . These laws naturally form a diagram, as follows:

$$\begin{array}{ccc} \pi_t & \text{-----} & \gamma_t \\ | & & | \\ p_t & \text{-----} & g_t \end{array}$$

We will see in chapters 7-8 that π_t, γ_t appear as “free analogues” of p_t, g_t , and that a full theory can be developed, with central limiting theorems for all 4 laws, convolution

semigroup results for all 4 laws too, and character results for all 4 laws too. And also, we will be back to the random matrices as well, with further results about them.

6e. Exercises

There has been a lot of non-trivial combinatorics and calculus in this chapter, sometimes only briefly explained, and as an exercise on all this, we have:

EXERCISE 6.34. *Clarify all the details in connection with the Wigner and Marchenko-Pastur computations, first at $t = 1$, and then for general $t > 0$.*

As before, these are things discussed in the above, but only briefly, this whole chapter having been just a brief and modest introduction to this exciting subject which are the random matrices. In the hope that you'll find some time, and fully do the exercise.

CHAPTER 7

Quantum spaces

7a. Group duals

According to the general results from chapter 5 above, the commutative von Neumann algebras are the algebras of the form $A = L^\infty(X)$, with X being a measured space. In view of this, an arbitrary von Neumann algebra A can be thought of as being of the form $A = L^\infty(X)$, with X being a “quantum measured space”.

In practice, all this is quite tricky, and has its limitations. As a first observation, what would help much our intuition would be the existence of an integration functional $tr : A \rightarrow \mathbb{C}$, having the trace property $tr(ab) = tr(ba)$, exactly as in the classical case, $A = L^\infty(X)$. That is, we would like to study “noncommutative functions”, under the assumption however that these functions commute under the integration.

As explained in chapters 5-6, such traces exist, and are extremely useful, in the type I case. For general von Neumann algebras, however, such traces can exist or not. Thus, we must impose the existence of the trace, as an axiom, as follows:

DEFINITION 7.1. *Given a von Neumann algebra $A \subset B(H)$, coming with a faithful positive unital form $tr : A \rightarrow \mathbb{C}$, satisfying the trace condition*

$$tr(ab) = tr(ba)$$

we formally write our von Neumann algebra $A = L^\infty(X)$, and call X a quantum probability space. We also write the trace as

$$tr = \int_X$$

and call it integration with respect to the uniform measure on X .

Observe that we use here the mathematical notation for the expectation, as an integral. We will sometimes use the probabilistic notation \mathbb{E} too, depending on the context. In fact, the most often will we just use the original tr , which is the most practical.

At the level of examples now, we know that both the algebras $L^\infty(X)$ and $M_N(\mathbb{C})$ have traces. More generally, the random matrix algebras $M_N(L^\infty(X))$ have traces. And even more generally, one can talk about traces on type I algebras.

We will be looking here for more general examples, beyond type I. In order to get started, let us look at the classical measured spaces X . We have been talking about such spaces all the time, for instance when considering Hilbert spaces of type $H = L^2(X)$, or commutative von Neuman algebras $A = L^\infty(X)$. But, the following question appears: in practice, how to find measured spaces X , in the real life?

In what regards “spaces” in general, there are plenty of them, because it suffices to consider the solutions of suitable algebraic or differential equations. However, in what regards spaces coming with a measure, there are not so many of them. The measure in question cannot come out of “nowhere”, and is usually a Haar measure. Thus, if we really want explicit examples of measured spaces X , for the various theories that we want to develop with them, we must look into Lie groups and their quotient spaces, and use:

THEOREM 7.2. *The following happen:*

- (1) *Any compact Lie group G has a uniform, or Haar measure, meaning a probability measure μ satisfying the following condition, for any $E \subset G$:*

$$\mu(gE) = \mu(Eg) = \mu(E)$$

- (2) *More generally, any quotient space G/H of such compact Lie groups has a Haar measure, which must be by definition invariant under the action of G .*

PROOF. This is something well-known, the idea being as follows:

(1) We can indeed pick any probability measure ν on our group G , and then start convolving with itself. The more we convolve, the more ν becomes invariant, and so in the $n \rightarrow \infty$ limit we will obtain an invariant measure μ . This is the general idea, but in practice, in order for things to converge, we must use a Cesàro limit, as follows:

$$\mu = \lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^n \nu^{*k}$$

The fact that this Cesàro limit converges indeed, and is left and right invariant under the action of G , is standard measure theory and functional analysis.

(2) Assume indeed that $X = G/H$ is an homogeneous space, coming from a closed subgroup $H \subset G$, as in the statement. The Haar measure on X appears then as the push-forward of the Haar measure of G , via the canonical quotient map:

$$\pi : G \rightarrow G/H$$

We can recover this as well via functional analysis. Indeed, at the level of algebras of functions, the above quotient map π produces an embedding, as follows:

$$i : C(X) \subset C(G)$$

Thus, we can define the Haar integration over X as being the restriction of the Haar integration over G , via this embedding i , and this gives the result. \square

Our idea in what follows will be that of finding “noncommutative” analogues of Theorem 7.2 above, providing us with explicit formulae of type $A = L^\infty(X)$, involving noncommutative von Neumann algebras A , and quantum measured spaces X .

In order to get started, let us forget about the general manifolds $X = G/H$, which are quite complicated, and focus on the compact groups G . The simplest such groups are the abelian ones, and in connection with these abelian groups, there is one important result to be known, namely the Pontrjagin duality result, which is as follows:

THEOREM 7.3. *The compact abelian groups G are in correspondence with the discrete abelian groups Γ , via Pontrjagin duality,*

$$G = \widehat{\Gamma} \quad , \quad \Gamma = \widehat{G}$$

with the dual of a locally compact group L being the locally compact group \widehat{L} consisting of the continuous group characters $\chi : L \rightarrow \mathbb{T}$.

PROOF. This is something very standard, the idea being that, given a group L as above, its continuous characters $\chi : L \rightarrow \mathbb{T}$ form indeed a group, that we can call \widehat{L} . The correspondence $L \rightarrow \widehat{L}$ constructed in this way has then the following properties:

(1) We have $\widehat{\mathbb{Z}_N} = \mathbb{Z}_N$. This is the basic computation to be performed, before anything else, and which is something algebraic, with roots of unity.

(2) More generally, the dual of a finite abelian group $G = \mathbb{Z}_{N_1} \times \dots \times \mathbb{Z}_{N_k}$ is the group G itself. This comes indeed from (1) and from $\widehat{G \times H} = \widehat{G} \times \widehat{H}$.

(3) At the opposite end now, that of the locally compact groups which are not compact, nor discrete, the main example, which is standard, is $\widehat{\mathbb{R}} = \mathbb{R}$.

(4) Getting now to what we are interested in, it follows from the definition of the correspondence $L \rightarrow \widehat{L}$ that when L is compact \widehat{L} is discrete, and vice versa.

(5) Finally, in order to best understand this latter phenomenon, the best is to work out the main pair of examples, which are $\widehat{\mathbb{T}} = \mathbb{Z}$ and $\widehat{\mathbb{Z}} = \mathbb{T}$. \square

Observe that the above result reminds Fourier analysis, which can be of three types, namely the usual one, on \mathbb{R} , then the theory of Fourier series, on \mathbb{T}/\mathbb{Z} , and then the discrete Fourier transform, over \mathbb{Z}_N . This analogy is of course not accidental, and in fact quite the opposite happens, in the sense that when looking for abstract “Fourier transform theories” we are naturally led into groups, and into Theorem 7.3 above.

This is something very interesting, for general mathematical knowledge, and also for our purposes here. Indeed, the explanation for the above phenomenon must involve some kind of convolution algebras, or more general operator algebras, depending on the type of functions involved. And if we want our Fourier analysis result to be for L^∞ functions, we must use von Neumann algebras. The result here is as follows:

THEOREM 7.4. *Given a discrete group Γ , we can construct its von Neumann algebra*

$$L(\Gamma) \subset B(l^2(\Gamma))$$

by using the left regular representation. This algebra has a faithful positive trace, $tr(g) = \delta_{g,1}$, and when Γ is abelian we have an isomorphism of tracial von Neumann algebras

$$L(\Gamma) \simeq L^\infty(G)$$

given by a Fourier type transform, where $G = \widehat{\Gamma}$ is the compact dual of Γ .

PROOF. There are many assertions here, the idea being as follows:

(1) The first part is standard, with the left regular representation of Γ working as expected, and being a unitary representation, as follows:

$$\Gamma \subset B(l^2(\Gamma)) \quad , \quad \pi(g) : h \rightarrow gh$$

(2) The positivity of the trace comes from the following alternative formula for it, with the equivalence with the definition in the statement being clear:

$$tr(T) = \langle T1, 1 \rangle$$

(3) The third part is standard as well, because when Γ is abelian the algebra $L(\Gamma)$ is commutative, and its spectral decomposition leads by delinearization to the group characters $\chi : \Gamma \rightarrow \mathbb{T}$, and so the dual group $G = \widehat{\Gamma}$, as indicated.

(4) Finally, the fact that our isomorphism transforms the trace of $L(\Gamma)$ into the Haar integration functional of $L^\infty(G)$ is clear. Moreover, the study of various examples show that what we constructed is in fact the Fourier transform, in its various incarnations. \square

All the above might seem a bit mysterious, and as for Theorem 7.3, working out the main examples is an excellent exercise, in order to best understand the von Neumann algebra theory to follow, and also mathematics in general. Always remember here the saying “there is only one tool in mathematics, and this is the Fourier transform”.

Getting back now to our quantum space questions, we have a beginning of answer here, because based on the above, we can formulate the following definition:

DEFINITION 7.5. *Given a discrete group Γ , not necessarily abelian, we can construct its abstract dual $G = \widehat{\Gamma}$ as a quantum measured space, via the following formula:*

$$L^\infty(G) = L(\Gamma)$$

In the case where Γ happens to be abelian, this quantum space $G = \widehat{\Gamma}$ is a classical space, namely the usual Pontrjagin dual of Γ , endowed with its Haar measure.

This is very nice, and we will keep building on this. Observe however that the theory developed above is incomplete, still missing a discussion of the “group-type structure” of the abstract dual $G = \widehat{\Gamma}$ that we constructed. We will come later with a solution to this problem, stating that G is a “compact quantum group”, in an appropriate sense.

7b. Compact spaces

Getting back now to our general quantum space program, one problem is that we cannot use von Neumann algebra theory, because this assumes the existence of a Hilbert space $H = L^2(X)$, and constructing X is precisely what we want.

In short, we are led here into the well-known quantum mechanical dilemma: does the operator algebra A come from the Hilbert space H , or vice versa?

Obviously, we need some kind of “vice versa” theory, in order to get started, and fortunately, such a theory exists, the starting definition being as follows:

DEFINITION 7.6. *A C^* -algebra is a complex algebra A , given with*

- (1) *A complex algebra involution $a \rightarrow a^*$,*
- (2) *A Banach algebra norm $a \rightarrow \|a\|$,*

which are related by the formula $\|aa^\| = \|a\|^2$.*

As basic examples, we have the operator algebra $B(H)$, for any Hilbert space H , and more generally, the norm closed $*$ -subalgebras $A \subset B(H)$. It is possible to prove that any C^* -algebra appears in this way, but this is a non-trivial result, and more on this later.

As a second class of basic examples, which are of interest for us, we have:

PROPOSITION 7.7. *If X is an abstract compact space, the algebra $C(X)$ of continuous functions $f : X \rightarrow \mathbb{C}$ is a C^* -algebra, with structure as follows:*

- (1) *The norm is the usual sup norm:*

$$\|f\| = \sup_{x \in X} |f(x)|$$

- (2) *The involution is the usual involution:*

$$f^*(x) = \overline{f(x)}$$

This algebra is commutative, in the sense that $fg = gf$, for any f, g .

PROOF. Almost everything here is trivial. Observe that we have indeed:

$$\begin{aligned} \|ff^*\| &= \sup_{x \in X} |f(x)\overline{f(x)}| \\ &= \sup_{x \in X} |f(x)|^2 \\ &= \|f\|^2 \end{aligned}$$

Thus, the axioms are satisfied, and finally $fg = gf$ is clear. \square

In general, the C^* -algebras can be thought of as being algebras of operators, over some Hilbert space which is not present. By using this philosophy, one can emulate spectral theory in this setting, with extensions of the various results from chapters 2-3 above:

THEOREM 7.8. *Given a C^* -algebra element $a \in A$, define its spectrum as:*

$$\sigma(a) = \left\{ \lambda \in \mathbb{C} \mid a - \lambda \notin A^{-1} \right\}$$

The following spectral theory results hold, exactly as in the $A = B(H)$ case:

- (1) *We have $\sigma(ab) \cup \{0\} = \sigma(ba) \cup \{0\}$.*
- (2) *We have polynomial, rational and holomorphic calculus.*
- (3) *As a consequence, the spectra are compact and non-empty.*
- (4) *The spectra of unitaries ($u^* = u^{-1}$) and self-adjoints ($a = a^*$) are on \mathbb{T}, \mathbb{R} .*
- (5) *The spectral radius of normal elements ($aa^* = a^*a$) is given by $\rho(a) = \|a\|$.*
- (6) *As an application, we have the continuous functional calculus.*

PROOF. The statements and proofs here are similar to those for the full operator algebra $A = B(H)$. All this is standard material, and in fact, things in chapters 2-3 above were written in such a way as for their extension now, to the general C^* -algebra setting, to be obvious. This material can be found as well in any operator algebra book. \square

We can now get back to the commutative C^* -algebras, and we have the following result, due to Gelfand, which will be of crucial importance for us:

THEOREM 7.9. *The commutative C^* -algebras are exactly the algebras of the form*

$$A = C(X)$$

with the “spectrum” X being the space of characters $\chi : A \rightarrow \mathbb{C}$, with the topology making each of the evaluation maps $ev_a : \chi \rightarrow \chi(a)$ continuous.

PROOF. This is a more or less a reformulation of Theorem 7.8 (6), but here is however a brief complete proof, based only on the spectral radius formula, from Theorem 7.8 (5). Given a commutative C^* -algebra A , we can define X as in the statement. Then X is compact, and $a \rightarrow ev_a$ is a morphism of algebras:

$$ev : A \rightarrow C(X)$$

(1) We first prove that ev is involutive. We use the following formula, which is similar to the $z = Re(z) + iIm(z)$ formula for the usual complex numbers:

$$a = \frac{a + a^*}{2} - i \cdot \frac{i(a - a^*)}{2}$$

Thus it is enough to prove the equality $ev_{a^*} = ev_a^*$ for self-adjoint elements a . But this is the same as proving that $a = a^*$ implies that ev_a is a real function, which is in turn true, because $ev_a(\chi) = \chi(a)$ is an element of $\sigma(a)$, contained in \mathbb{R} .

(2) Since A is commutative, each element is normal, so ev is isometric:

$$\|ev_a\| = \rho(a) = \|a\|$$

(3) It remains to prove that ev is surjective. But this follows from the Stone-Weierstrass theorem, because $ev(A)$ is a closed subalgebra of $C(X)$, which separates the points. \square

In analogy with what happens in the von Neumann algebra case, let us formulate:

DEFINITION 7.10. *Given a C^* -algebra A , we write*

$$A = C(X)$$

and call X a compact quantum space.

As an illustration, let us discuss what happens for the group algebras. The situation here is quite tricky, more complicated than in the von Neumann algebra setting:

DEFINITION 7.11. *Associated to any discrete group Γ are several group C^* -algebras,*

$$C^*(\Gamma) \rightarrow C_\pi^*(\Gamma) \rightarrow C_{red}^*(\Gamma)$$

which are constructed as follows:

- (1) $C^*(\Gamma)$ is the closure of the group algebra $\mathbb{C}[\Gamma]$, with involution $g^* = g^{-1}$, with respect to the maximal C^* -seminorm on this $*$ -algebra, which is a C^* -norm.
- (2) $C_{red}^*(\Gamma)$ is the norm closure of the group algebra $\mathbb{C}[\Gamma]$ in the left regular representation, on the Hilbert space $l^2(\Gamma)$, given by $\lambda(g)(h) = gh$ and linearity.
- (3) $C_\pi^*(\Gamma)$ can be a priori any intermediate C^* -algebra, but for best results, the indexing object π must be a unitary group representation, satisfying $\pi \otimes \pi \subset \pi$.

This is something quite technical, with (2) being very similar to the von Neumann algebra construction, from Theorem 7.4 above, with (1) being something new, with the norm property there coming from (2), and finally with (3) being an informal statement.

When Γ is finite, or abelian, or more generally amenable, all the above group algebras coincide. In the abelian case, that we are particularly interested in here, the precise result is as follows, complementing the L^∞ analysis from Theorem 7.4 above:

THEOREM 7.12. *When Γ is abelian all its group C^* -algebras coincide, and we have an isomorphism as follows, given by a Fourier type transform,*

$$C^*(\Gamma) \simeq C(G)$$

where $G = \widehat{\Gamma}$ is the compact dual of Γ . Moreover, this isomorphism transforms the standard group algebra trace $tr(g) = \delta_{g,1}$ into the Haar integration of G .

PROOF. Since Γ is abelian, any of its group C^* -algebras $A = C_\pi^*(\Gamma)$ is commutative. Thus, we can apply the Gelfand theorem, and we obtain $A = C(X)$, with $X = Spec(A)$. But the spectrum $X = Spec(A)$, consisting of the characters $\chi : A \rightarrow \mathbb{C}$, can be identified by delinearizing with the Pontrjagin dual $G = \widehat{\Gamma}$, and this gives the results. \square

As already mentioned, the various group C^* -algebras coincide when Γ is more generally amenable. However, when Γ is not amenable, the full and reduced group algebras are known to be distinct. And this is something that can really happen, a well-known example of non-amenable group being the free group on two generators F_2 .

Philosophically speaking, we keep from the above discussion the conclusion that, while the quantum space $G = \widehat{\Gamma}$ is definitely unique, and the L^∞ functions on it are uniquely determined, via the formula $L^\infty(G) = L(\Gamma)$, in what concerns the continuous functions, “some of them are more continuous than some other”, in the non-amenable setting.

This might seem quite surprising, at a first glance, but with a bit of experience, and some further functional analysis knowledge, not only this becomes a non-issue in the long term, but rather something quite natural. It is in fact part of the quantum mechanics philosophy that the classical world that we live in is a kind of “wonderland”, appearing via mysterious $N \rightarrow \infty$ limits which simplify everything, and the continuity of functions in the classical setting, which is a luminous, unique notion, appears like this.

Let us clarify now the relation between C^* -algebras and von Neumann algebras, among others in order to shed more light on the above issues. We are led to the question of proving that any C^* -algebra appears as an operator algebra, and we first have here:

PROPOSITION 7.13. *Let A be a commutative C^* -algebra, write $A = C(X)$, with X being a compact space, and let μ be a positive measure on X . We have then an embedding*

$$A \subset B(H)$$

where $H = L^2(X)$, with $f \in A$ corresponding to the operator $g \rightarrow fg$.

PROOF. Given a continuous function $f \in C(X)$, consider the operator $T_f(g) = fg$, on $H = L^2(X)$. Observe that T_f is indeed well-defined, and bounded as well, because:

$$\begin{aligned} \|fg\|_2 &= \sqrt{\int_X |f(x)|^2 |g(x)|^2 d\mu(x)} \\ &\leq \|f\|_\infty \|g\|_2 \end{aligned}$$

The application $f \rightarrow T_f$ being linear, involutive, continuous, and injective as well, we obtain in this way a C^* -algebra embedding $A \subset B(H)$, as claimed. \square

In order to extend the above construction, let us start with:

DEFINITION 7.14. *Consider a C^* -algebra A .*

- (1) $a \in A$ is called *positive* when $a = a^*$ and $\sigma(a) \subset [0, \infty)$.
- (2) $\varphi : A \rightarrow \mathbb{C}$ is called *positive* when $a \geq 0 \implies \varphi(a) \geq 0$.
- (3) $\varphi : A \rightarrow \mathbb{C}$ is called *faithful and positive* when $a \geq 0, a \neq 0 \implies \varphi(a) > 0$.

In the commutative case, $A = C(X)$, the positive elements are the positive functions, $f : X \rightarrow [0, \infty)$. As for the positive linear forms $\varphi : A \rightarrow \mathbb{C}$, these appear as follows, with μ being positive, and strictly positive if we want φ to be faithful and positive:

$$\varphi(f) = \int_X f(x) d\mu(x)$$

In general, the positive linear forms can be thought of as being integration functionals with respect to some underlying “positive measures”. We can use them as follows:

PROPOSITION 7.15. *Let $\varphi : A \rightarrow \mathbb{C}$ be a positive linear form.*

- (1) $\langle a, b \rangle = \varphi(ab^*)$ defines a generalized scalar product on A .
- (2) By separating and completing we obtain a Hilbert space H .
- (3) $\pi(a) : b \rightarrow ab$ defines a representation $\pi : A \rightarrow B(H)$.
- (4) If φ is faithful in the above sense, then π is faithful.

PROOF. Almost everything here is straightforward, as follows:

(1) This is clear from definitions, and from the basic properties of the positive elements $a \geq 0$, which can be established exactly as in the $A = B(H)$ case.

(2) This is a standard procedure, which works for any scalar product, the idea being that of dividing by the vectors satisfying $\langle x, x \rangle = 0$, then completing.

(3) All the verifications here are standard algebraic computations, in analogy with what we have seen many times, for multiplication operators, or group algebras.

(4) Assuming that we have $a \neq 0$, we have then $\pi(aa^*) \neq 0$, which in turn implies by faithfulness that we have $\pi(a) \neq 0$, which gives the result. \square

In order to establish the embedding theorem, it remains to prove that any C^* -algebra has a faithful positive linear form $\varphi : A \rightarrow \mathbb{C}$. This is something more technical:

PROPOSITION 7.16. *Let A be a C^* -algebra.*

- (1) Any positive linear form $\varphi : A \rightarrow \mathbb{C}$ is continuous.
- (2) A linear form φ is positive iff there is a norm one $h \in A_+$ such that $\|\varphi\| = \varphi(h)$.
- (3) For any $a \in A$ there exists a positive norm one form φ such that $\varphi(aa^*) = \|a\|^2$.
- (4) If A is separable there is a faithful positive form $\varphi : A \rightarrow \mathbb{C}$.

PROOF. The proof here is quite technical, inspired from the existence proof of the probability measures on abstract compact spaces, the idea being as follows:

(1) This follows from Proposition 7.15, via the following estimate:

$$\begin{aligned} |\varphi(a)| &\leq \|\pi(a)\|\varphi(1) \\ &\leq \|a\|\varphi(1) \end{aligned}$$

(2) In one sense we can take $h = 1$. Conversely, let $a \in A_+$, $\|a\| \leq 1$. We have:

$$\begin{aligned} |\varphi(h) - \varphi(a)| &\leq \|\varphi\| \cdot \|h - a\| \\ &\leq \varphi(h)1 \\ &= \varphi(h) \end{aligned}$$

Thus $\operatorname{Re}(\varphi(a)) \geq 0$, and with $a = 1 - h$ we obtain:

$$\operatorname{Re}(\varphi(1 - h)) \geq 0$$

But this gives $\operatorname{Re}(\varphi(1)) \geq \|\varphi\|$, and so $\varphi(1) = \|\varphi\|$. Thus we can assume $h = 1$. Now observe that for any self-adjoint element a , and any $t \in \mathbb{R}$ we have, with $\varphi(a) = x + iy$:

$$\begin{aligned} \varphi(1)^2(1 + t^2\|a\|^2) &\geq \varphi(1)^2\|1 + t^2a^2\| \\ &= \|\varphi\|^2 \cdot \|1 + ita\|^2 \\ &\geq |\varphi(1 + ita)|^2 \\ &= |\varphi(1) - ty + itx|^2 \\ &\geq (\varphi(1) - ty)^2 \end{aligned}$$

Thus we have $y = 0$, and this finishes the proof of our remaining claim.

(3) We can set $\varphi(\lambda aa^*) = \lambda\|a\|^2$ on the linear space spanned by aa^* , then extend this functional by Hahn-Banach, to the whole A . The positivity follows from (2).

(4) This is standard, by starting with a dense sequence (a_n) , and taking the Cesàro limit of the functionals constructed in (3). We have $\varphi(aa^*) > 0$, and we are done. \square

With these ingredients in hand, we can now state and prove:

THEOREM 7.17. *Any C^* -algebra appears as a norm closed $*$ -algebra of operators*

$$A \subset B(H)$$

over a certain Hilbert space H . When A is separable, H can be taken to be separable.

PROOF. This result, called GNS representation theorem after Gelfand, Naimark and Segal, follows indeed by combining the construction from Proposition 7.15 above, called GNS construction, with the existence result from Proposition 7.16. \square

As a first application, let us get back to the bad functoriality properties of the Gelfand correspondence, in relation with group algebras. We can fix these issues as follows:

THEOREM 7.18. *We can talk about compact quantum measured spaces, as follows:*

- (1) *The category of compact quantum measured spaces (X, μ) is the category of the C^* -algebras with faithful traces (A, φ) , with the arrows reversed.*
- (2) *In the case where we have a non-faithful trace φ , we can still talk about the corresponding space (X, μ) , by performing the GNS construction.*
- (3) *By taking the weak closure in the GNS representation, we obtain the von Neumann algebra $A'' = L^\infty(X)$, in the previous general measured space sense.*
- (4) *In the particular case of the group algebras, all the group algebras $C_\pi^*(\Gamma)$ give rise to the same space, namely the usual abstract dual $G = \widehat{\Gamma}$.*

PROOF. All this follows from Theorem 7.17, and from the other things that we already know, with the whole result itself being something rather philosophical. \square

In short, we have now our fix for the bad functoriality properties of the Gelfand correspondence, at least at the theoretical level.

7c. Spheres and tori

The theory developed above is still not satisfactory, because it eventually takes us back to von Neumann algebras, and to Hilbert spaces, and this is precisely the place where we don't what to get. Thus, our functoriality “fix” is not a very good fix, and we must invent something else. In order to do so, our idea will be that of restricting the attention to certain special classes of algebraic manifolds. Let us begin with:

DEFINITION 7.19. *We have compact quantum spaces, constructed as follows,*

$$C(S_{\mathbb{R},+}^{N-1}) = C^* \left(x_1, \dots, x_N \mid x_i = x_i^*, \sum_i x_i^2 = 1 \right)$$

$$C(S_{\mathbb{C},+}^{N-1}) = C^* \left(x_1, \dots, x_N \mid \sum_i x_i x_i^* = \sum_i x_i^* x_i = 1 \right)$$

called respectively the free real sphere, and the free complex sphere.

Here the C^* symbols on the right stand for “universal C^* -algebra generated by”. The fact that such universal C^* -algebras exist indeed follows by considering the corresponding universal $*$ -algebras, and then completing with respect to the biggest C^* -norm. Observe that this biggest C^* -norm exists indeed, because the quadratic conditions give:

$$\begin{aligned} \|x_i\|^2 &= \|x_i x_i^*\| \\ &\leq \left\| \sum_i x_i x_i^* \right\| \\ &= 1 \end{aligned}$$

Given a compact quantum space X , its classical version is the space X_{class} obtained by dividing $C(X)$ by its commutator ideal, and then applying the Gelfand theorem:

$$C(X_{class}) = C(X)/I \quad , \quad I = \langle [a, b] \rangle$$

Observe that we have an embedding of compact quantum spaces $X_{class} \subset X$. In this situation, we also say that X appears as a “liberation” of X . We have:

PROPOSITION 7.20. *We have embeddings of compact quantum spaces*

$$\begin{array}{ccc} S_{\mathbb{C}}^{N-1} & \longrightarrow & S_{\mathbb{C},+}^{N-1} \\ \uparrow & & \uparrow \\ S_{\mathbb{R}}^{N-1} & \longrightarrow & S_{\mathbb{R},+}^{N-1} \end{array}$$

and the spaces on the right appear as liberations of the spaces of the left.

PROOF. The first assertion is clear. For the second one, we must establish the following isomorphisms, where C_{comm}^* stands for “universal commutative C^* -algebra”:

$$C(S_{\mathbb{R}}^{N-1}) = C_{comm}^* \left(x_1, \dots, x_N \mid x_i = x_i^*, \sum_i x_i^2 = 1 \right)$$

$$C(S_{\mathbb{C}}^{N-1}) = C_{comm}^* \left(x_1, \dots, x_N \mid \sum_i x_i x_i^* = \sum_i x_i^* x_i = 1 \right)$$

But these isomorphisms are both clear, by using the Gelfand theorem. \square

We can now introduce a broad class of quantum manifolds, as follows:

DEFINITION 7.21. *A real algebraic submanifold $X \subset S_{\mathbb{C},+}^{N-1}$ is a closed quantum space defined, at the level of the corresponding C^* -algebra, by a formula of type*

$$C(X) = C(S_{\mathbb{C},+}^{N-1}) / \langle f_i(x_1, \dots, x_N) = 0 \rangle$$

for certain noncommutative polynomials $f_i \in \mathbb{C} \langle x_1, \dots, x_N \rangle$.

Observe that such manifolds exist indeed, because the free complex spheres themselves exist, and this due to the fact that the quadratic conditions give:

$$\|x_i\| \leq 1$$

This estimate, explained before, is something extremely important, and any attempt of further extending Definition 7.21, beyond the sphere level, stumbles into this. There are no such things as free analogues of \mathbb{R}^N or \mathbb{C}^N , and the problem comes from this.

In practice now, while our assumption $X \subset S_{\mathbb{C},+}^{N-1}$ is definitely something technical, we are not losing much when imposing it, and we have the following list of examples:

PROPOSITION 7.22. *The following are algebraic submanifolds $X \subset S_{\mathbb{C},+}^{N-1}$:*

- (1) *The spheres $S_{\mathbb{R}}^{N-1} \subset S_{\mathbb{C}}^{N-1}$, $S_{\mathbb{R},+}^{N-1} \subset S_{\mathbb{C},+}^{N-1}$.*
- (2) *Any compact Lie group, $G \subset U_n$, with $n^2 = N$.*
- (3) *The duals $\widehat{\Gamma}$ of finitely generated groups, $\Gamma = \langle g_1, \dots, g_N \rangle$.*

PROOF. These facts are all well-known, the proof being as follows:

(1) This is true by definition of our various spheres.

(2) Given a closed subgroup $G \subset U_n$, we have indeed an embedding $G \subset S_{\mathbb{C}}^{N-1}$, with $N = n^2$, given in double indices by $x_{ij} = \frac{u_{ij}}{\sqrt{n}}$, that we can further compose with the standard embedding $S_{\mathbb{C}}^{N-1} \subset S_{\mathbb{C},+}^{N-1}$. As for the fact that we obtain indeed a real algebraic manifold, this is well-known, coming either from Lie theory or from Tannakian duality.

(3) This follows from the fact that the variables $x_i = \frac{g_i}{\sqrt{N}}$ satisfy the quadratic relations $\sum_i x_i x_i^* = \sum_i x_i^* x_i = 1$, with the algebraicity claim of the manifold being clear. \square

At the level of the general theory, we have the following version of the Gelfand theorem, which is something very useful, and that we will use several times in what follows:

THEOREM 7.23. *When $X \subset S_{\mathbb{C},+}^{N-1}$ is an algebraic manifold, given by*

$$C(X) = C(S_{\mathbb{C},+}^{N-1}) / \left\langle f_i(x_1, \dots, x_N) = 0 \right\rangle$$

for certain noncommutative polynomials $f_i \in \mathbb{C} \langle x_1, \dots, x_N \rangle$, we have

$$X_{class} = \left\{ x \in S_{\mathbb{C}}^{N-1} \mid f_i(x_1, \dots, x_N) = 0 \right\}$$

and X appears as a liberation of X_{class} .

PROOF. This is something that already met, in the context of the free spheres. In general, the proof is similar, by using the Gelfand theorem. Indeed, if we denote by X'_{class} the manifold constructed in the statement, then we have a quotient map of C^* -algebras as follows, mapping standard coordinates to standard coordinates:

$$C(X_{class}) \rightarrow C(X'_{class})$$

Conversely now, from $X \subset S_{\mathbb{C},+}^{N-1}$ we obtain $X_{class} \subset S_{\mathbb{C}}^{N-1}$, and since the relations defining X'_{class} are satisfied by X_{class} , we obtain an inclusion of subspaces $X_{class} \subset X'_{class}$. Thus, at the level of algebras of continuous functions, we have a quotient map of C^* -algebras as follows, mapping standard coordinates to standard coordinates:

$$C(X'_{class}) \rightarrow C(X_{class})$$

Thus, we have constructed a pair of inverse morphisms, and we are done. \square

Getting back now to the examples, the above formalism allows us to have a new, more geometric look at the group duals. Let us formulate indeed:

DEFINITION 7.24. *Given a closed subspace $S \subset S_{\mathbb{C},+}^{N-1}$, the subspace $T \subset S$ given by*

$$C(T) = C(S) / \left\langle x_i x_i^* = x_i^* x_i = \frac{1}{N} \right\rangle$$

is called associated torus. In the real case, $S \subset S_{\mathbb{R},+}^{N-1}$, we also call T cube.

As a basic example here, for $S = S_{\mathbb{C}}^{N-1}$ the corresponding submanifold $T \subset S$ appears by imposing the relations $|x_i| = \frac{1}{\sqrt{N}}$ to the coordinates, so we obtain a torus:

$$S = S_{\mathbb{C}}^{N-1} \implies T = \left\{ x \in \mathbb{C}^N \mid |x_i| = \frac{1}{\sqrt{N}} \right\}$$

As for the case of the real sphere, $S = S_{\mathbb{R}}^{N-1}$, here the submanifold $T \subset S$ appears by imposing the relations $x_i = \pm \frac{1}{\sqrt{N}}$ to the coordinates, so we obtain a cube:

$$S = S_{\mathbb{R}}^{N-1} \implies T = \left\{ x \in \mathbb{R}^N \mid x_i = \pm \frac{1}{\sqrt{N}} \right\}$$

Observe that we have a relation here with groups, because the complex torus computed above is the group \mathbb{T}^N , and the cube is the finite group \mathbb{Z}_2^N . In fact, we have:

THEOREM 7.25. *The tori of the basic spheres are all group duals, as follows,*

$$\begin{array}{ccc} \mathbb{T}^N & \longrightarrow & \widehat{F}_N \\ \uparrow & & \uparrow \\ \mathbb{Z}_2^N & \longrightarrow & \widehat{\mathbb{Z}_2^{*N}} \end{array}$$

where F_N is the free group on N generators, and $*$ is a group-theoretical free product.

PROOF. In order to prove this, let us get back to the general setting of Definition 7.23, and assume that the subspace there $S \subset S_{\mathbb{C},+}^{N-1}$ is an algebraic manifold, as follows:

$$C(S) = C(S_{\mathbb{C},+}^{N-1}) / \langle f_i(x_1, \dots, x_N) = 0 \rangle$$

In order to get to group algebras, let us rescale the coordinates, $u_i = \frac{x_i}{\sqrt{N}}$. Consider as well the corresponding rescalings of the polynomials f_i , given by:

$$g_i(u_1, \dots, u_N) = f_i(\sqrt{N}u_1, \dots, \sqrt{N}u_N)$$

Now since the relations defining $T \subset S$ from Definition 7.23 correspond to the fact that the rescaled coordinates u_i must be unitaries, we obtain:

$$C(T) = C^* \left(u_1, \dots, u_N \mid u_i^* = u_i^{-1}, g_i(u_1, \dots, u_N) = 0 \right)$$

Now in the case of the 4 main spheres, from Proposition 7.20 above, we obtain from this that the diagram formed by the corresponding algebras $C(T)$ is as follows:

$$\begin{array}{ccc} C^*(\mathbb{Z}^N) & \longleftarrow & C^*(\mathbb{Z}^{*N}) \\ \downarrow & & \downarrow \\ C^*(\mathbb{Z}_2^N) & \longleftarrow & C^*(\mathbb{Z}_2^{*N}) \end{array}$$

We are therefore led to the conclusion in the statement. □

All this is very nice, and not using the GNS theorem, or any kind of advanced mathematics. As a conclusion to these considerations, the Gelfand theorem alone produces out of “nothing”, or at least out of some basic common sense, some potentially interesting mathematics. We will be back later on to the above objects, on several occasions.

With these results in hand, we are now ready for formulating our second “fix” for the functoriality issues of the Gelfand correspondence, as follows:

DEFINITION 7.26. *The category of the real algebraic submanifolds $X \subset S_{\mathbb{C},+}^{N-1}$ is formed by the compact quantum spaces appearing as follows,*

$$\mathcal{C}(X) = C(S_{\mathbb{C},+}^{N-1}) / \langle f_i(x_1, \dots, x_N) = 0 \rangle$$

with $f_i \in \mathbb{C} \langle x_1, \dots, x_N \rangle$ being noncommutative polynomials, and with the arrows $X \rightarrow Y$ being the $*$ -algebra morphisms between the $*$ -algebras of coordinates

$$\mathcal{C}(Y) \rightarrow \mathcal{C}(X)$$

mapping standard coordinates to standard coordinates.

In other words, what we are doing here is that of proposing a definition for the morphisms between the compact quantum spaces, in the particular case where these compact quantum spaces are algebraic submanifolds of the free complex sphere $S_{\mathbb{C},+}^{N-1}$.

The point is that Definition 7.26 works fine for the group duals, as follows:

THEOREM 7.27. *The category of the finitely generated groups*

$$\Gamma = \langle g_1, \dots, g_N \rangle$$

with the morphisms mapping generators to generators, embeds contravariantly via

$$\Gamma \rightarrow \widehat{\Gamma}$$

into the category of real algebraic submanifolds $X \subset S_{\mathbb{C},+}^{N-1}$.

PROOF. We know from Proposition 7.22 above that, given an arbitrary finitely generated group $\Gamma = \langle g_1, \dots, g_N \rangle$, we have an embedding $\widehat{\Gamma} \subset S_{\mathbb{C},+}^{N-1}$ given by:

$$x_i = \frac{g_i}{\sqrt{N}}$$

Now since a morphism of $*$ -algebras of coordinates $\mathbb{C}[\Gamma] \rightarrow \mathbb{C}[\Lambda]$ mapping coordinates to coordinates corresponds to a morphism of groups $\Gamma \rightarrow \Lambda$ mapping generators to generators, our notion of isomorphism is indeed the correct one, as claimed. \square

Summarizing, we have now a second fix for the bad functoriality of the Gelfand correspondence, adding to the standard von Neumann algebra fix from Theorem 7.18, but using this time $*$ -algebras, in a tricky way, with free complex spheres, as above.

Technically speaking, we will see later on that Theorem 7.27 has extensions to the various quantum groups and quantum homogeneous spaces that we will be interested in, which all fit into the general framework of Definition 7.26. Also, we will discuss as well the relation with von Neumann algebras, after constructing Haar functionals for the free real and complex spheres, and for the other manifolds that we are interested in.

7d. Quantum groups

With these ingredients in hand, we can now develop our “quantum algebraic manifold” program. Let us start with the following key definition, due to Woronowicz:

DEFINITION 7.28. *A Woronowicz algebra is a C^* -algebra A , given with a unitary matrix $u \in M_N(A)$ whose coefficients generate A , such that the formulae*

$$\begin{aligned}\Delta(u_{ij}) &= \sum_k u_{ik} \otimes u_{kj} \\ \varepsilon(u_{ij}) &= \delta_{ij} \\ S(u_{ij}) &= u_{ji}^*\end{aligned}$$

define morphisms of C^* -algebras $\Delta : A \rightarrow A \otimes A$, $\varepsilon : A \rightarrow \mathbb{C}$, $S : A \rightarrow A^{opp}$.

We say that A is cocommutative when $\Sigma\Delta = \Delta$, where $\Sigma(a \otimes b) = b \otimes a$ is the flip. We have the following result, which justifies the terminology and axioms:

PROPOSITION 7.29. *The following are Woronowicz algebras:*

(1) $C(G)$, with $G \subset U_N$ compact Lie group. Here the structural maps are:

$$\begin{aligned}\Delta(\varphi) &= (g, h) \rightarrow \varphi(gh) \\ \varepsilon(\varphi) &= \varphi(1) \\ S(\varphi) &= g \rightarrow \varphi(g^{-1})\end{aligned}$$

(2) $C^*(\Gamma)$, with $F_N \rightarrow \Gamma$ finitely generated group. Here the structural maps are:

$$\begin{aligned}\Delta(g) &= g \otimes g \\ \varepsilon(g) &= 1 \\ S(g) &= g^{-1}\end{aligned}$$

Moreover, we obtain in this way all the commutative/cocommutative algebras.

PROOF. In both cases, we have to exhibit a certain matrix u . For the first assertion, we can use the matrix $u = (u_{ij})$ formed by matrix coordinates of G , given by:

$$g = \begin{pmatrix} u_{11}(g) & \dots & u_{1N}(g) \\ \vdots & & \vdots \\ u_{N1}(g) & \dots & u_{NN}(g) \end{pmatrix}$$

For the second assertion, we can use the diagonal matrix formed by generators:

$$u = \begin{pmatrix} g_1 & & 0 \\ & \ddots & \\ 0 & & g_N \end{pmatrix}$$

Finally, the last assertion follows from the Gelfand theorem, in the commutative case, and in the cocommutative case, we will be back to this later. \square

In general now, the structural maps Δ, ε, S have the following properties:

PROPOSITION 7.30. *Let (A, u) be a Woronowicz algebra.*

(1) Δ, ε satisfy the usual axioms for a comultiplication and a counit, namely:

$$\begin{aligned} (\Delta \otimes id)\Delta &= (id \otimes \Delta)\Delta \\ (\varepsilon \otimes id)\Delta &= (id \otimes \varepsilon)\Delta = id \end{aligned}$$

(2) S satisfies the antipode axiom, on the $*$ -subalgebra generated by entries of u :

$$m(S \otimes id)\Delta = m(id \otimes S)\Delta = \varepsilon(\cdot)1$$

(3) In addition, the square of the antipode is the identity, $S^2 = id$.

PROOF. Observe first that the result holds in the case where A is commutative. Indeed, by using Proposition 7.29 we can write:

$$\Delta = m^t \quad , \quad \varepsilon = u^t \quad , \quad S = i^t$$

The above 3 conditions come then by transposition from the basic 3 group theory conditions satisfied by m, u, i , which are as follows, with $\delta(g) = (g, g)$:

$$\begin{aligned} m(m \times id) &= m(id \times m) \\ m(id \times u) &= m(u \times id) = id \\ m(id \times i)\delta &= m(i \times id)\delta = 1 \end{aligned}$$

Observe that $S^2 = id$ is satisfied as well, coming from $i^2 = id$, which is a consequence of the group axioms. In general now, the proof goes as follows:

(1) This follows indeed from the following two computations:

$$\begin{aligned} (\Delta \otimes id)\Delta(u_{ij}) &= \sum_{kl} u_{ik} \otimes u_{kl} \otimes u_{lj} \\ (id \otimes \Delta)\Delta(u_{ij}) &= \sum_{kl} u_{ik} \otimes u_{kl} \otimes u_{lj} \end{aligned}$$

(2) The proof here is similar, coming from the following computations:

$$\begin{aligned} (id \otimes \varepsilon)\Delta(u_{ij}) &= u_{ij} \\ (\varepsilon \otimes id)\Delta(u_{ij}) &= u_{ij} \end{aligned}$$

(3) By using the fact that the matrix $u = (u_{ij})$ is unitary, we obtain indeed:

$$\begin{aligned} m(id \otimes S)\Delta(u_{ij}) &= \delta_{ij} \\ m(S \otimes id)\Delta(u_{ij}) &= \delta_{ij} \end{aligned}$$

Finally, the formula $S^2 = id$ holds as well on the generators, and we are done. \square

Summarizing, the Woronowicz algebras appear to have very nice properties. In view of Proposition 7.29 above, we can now formulate the following definition:

DEFINITION 7.31. *Given a Woronowicz algebra A , we formally write*

$$A = C(G) = C^*(\Gamma)$$

and call G compact quantum group, and Γ discrete quantum group.

When A is both commutative and cocommutative, G is a compact abelian group, Γ is a discrete abelian group, and these groups are dual to each other:

$$\begin{aligned} G &= \widehat{\Gamma} \\ \Gamma &= \widehat{G} \end{aligned}$$

In general, we still agree to write $G = \widehat{\Gamma}$, $\Gamma = \widehat{G}$, but in a formal sense, for instance by talking about covariant and contravariant isomorphisms of the corresponding categories.

Finally, for the theory to in tune with our previous considerations regarding the quantum manifolds, we must complement Definition 7.28 and Definition 7.31 with:

DEFINITION 7.32. *Given two Woronowicz algebras (A, u) and (B, v) , we write*

$$A \simeq B$$

and we identify as well the corresponding compact and discrete quantum groups, when we have an isomorphism between the $$ -algebras $\mathcal{A} = \langle u_{ij} \rangle$ and $\mathcal{B} = \langle v_{ij} \rangle$,*

$$\mathcal{A} \simeq \mathcal{B}$$

mapping standard coordinates to standard coordinates.

With this in hand, the functoriality problem for the compact and discrete quantum groups is fixed. To be more precise, any compact or discrete quantum group corresponds to a unique Woronowicz algebra, up to equivalence. We will be back to this later, with a number of supplementary comments, when talking about amenability.

In order to develop some theory, let us call corepresentation of A any unitary matrix $v \in M_n(\mathcal{A})$, with $\mathcal{A} = \langle u_{ij} \rangle$, satisfying the same conditions as u , namely:

$$\begin{aligned} \Delta(v_{ij}) &= \sum_k v_{ik} \otimes v_{kj} \\ \varepsilon(v_{ij}) &= \delta_{ij} \\ S(v_{ij}) &= v_{ji}^* \end{aligned}$$

These corepresentations can be thought of as corresponding to the unitary representations of the underlying compact quantum group G . As main examples, we have $u = (u_{ij})$ itself, its conjugate $\bar{u} = (u_{ij}^*)$, as well as any tensor product between u, \bar{u} .

We have the following key result, due to Woronowicz [98]:

THEOREM 7.33. *Any Woronowicz algebra has a unique Haar integration functional,*

$$\left(\int_G \otimes id \right) \Delta = \left(id \otimes \int_G \right) \Delta = \int_G (\cdot) 1$$

which can be constructed by starting with any faithful positive form $\varphi \in A^*$, and setting

$$\int_G = \lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^n \varphi^{*k}$$

where $\phi * \psi = (\phi \otimes \psi)\Delta$. Moreover, for any corepresentation $v \in M_n(\mathbb{C}) \otimes A$ we have

$$\left(id \otimes \int_G \right) v = P$$

where P is the orthogonal projection onto $Fix(v) = \{\xi \in \mathbb{C}^n | v\xi = \xi\}$.

PROOF. Following [98], this can be done in 3 steps, as follows:

(1) Given $\varphi \in A^*$, our claim is that the following limit converges, for any $a \in A$:

$$\int_{\varphi} a = \lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^n \varphi^{*k}(a)$$

Indeed, by linearity we can assume that a is the coefficient of corepresentation, $a = (\tau \otimes id)v$. But in this case, an elementary computation shows that we have the following formula, where P_{φ} is the orthogonal projection onto the 1-eigenspace of $(id \otimes \varphi)v$:

$$\left(id \otimes \int_{\varphi} \right) v = P_{\varphi}$$

(2) Since $v\xi = \xi$ implies $[(id \otimes \varphi)v]\xi = \xi$, we have $P_{\varphi} \geq P$, where P is the orthogonal projection onto the space $Fix(v) = \{\xi \in \mathbb{C}^n | v\xi = \xi\}$. The point now is that when $\varphi \in A^*$ is faithful, by using a positivity trick, one can prove that we have $P_{\varphi} = P$. Thus our linear form \int_{φ} is independent of φ , and is given on coefficients $a = (\tau \otimes id)v$ by:

$$\left(id \otimes \int_{\varphi} \right) v = P$$

(3) With the above formula in hand, the left and right invariance of $\int_G = \int_{\varphi}$ is clear on coefficients, and so in general, and this gives all the assertions. See [98]. \square

As a main application, we can develop a Peter-Weyl type theory for the corepresentations of A . Consider the dense *-subalgebra $\mathcal{A} \subset A$ generated by the coefficients of the fundamental corepresentation u , and endow it with the following scalar product:

$$\langle a, b \rangle = \int_G ab^*$$

With this convention, we have the following fundamental result:

THEOREM 7.34. *We have the following Peter-Weyl type results:*

- (1) *Any corepresentation decomposes as a sum of irreducible corepresentations.*
- (2) *Each irreducible corepresentation appears inside a certain $u^{\otimes k}$.*
- (3) $\mathcal{A} = \bigoplus_{v \in \text{Irr}(A)} M_{\dim(v)}(\mathbb{C})$, *the summands being pairwise orthogonal.*
- (4) *The characters of irreducible corepresentations form an orthonormal system.*

PROOF. All these results are from [98], the idea being as follows:

(1) Given $v \in M_n(A)$, its intertwiner algebra $\text{End}(v) = \{T \in M_n(\mathbb{C}) | Tv = vT\}$ is a finite dimensional C^* -algebra, and so decomposes as $\text{End}(v) = M_{n_1}(\mathbb{C}) \oplus \dots \oplus M_{n_r}(\mathbb{C})$. But this gives a decomposition of type $v = v_1 + \dots + v_r$, as desired.

(2) Consider indeed the Peter-Weyl corepresentations, $u^{\otimes k}$ with k colored integer, defined by $u^{\otimes 0} = 1$, $u^{\otimes \circ} = u$, $u^{\otimes \bullet} = \bar{u}$ and multiplicativity. The coefficients of these corepresentations span the dense algebra \mathcal{A} , and by using (1), this gives the result.

(3) Here the direct sum decomposition, which is technically a $*$ -coalgebra isomorphism, follows from (2). As for the second assertion, this follows from the fact that $(\text{id} \otimes \int_G)v$ is the orthogonal projection P_v onto the space $\text{Fix}(v)$, for any corepresentation v .

(4) Let us define indeed the character of $v \in M_n(A)$ to be the matrix trace, $\chi_v = \text{Tr}(v)$. Since this character is a coefficient of v , the orthogonality assertion follows from (3). As for the norm 1 claim, this follows once again from $(\text{id} \otimes \int_G)v = P_v$. \square

We can now solve a problem that we left open before, namely:

PROPOSITION 7.35. *The cocommutative Woronowicz algebras appear as the quotients*

$$C^*(\Gamma) \rightarrow A \rightarrow C_{\text{red}}^*(\Gamma)$$

given by $A = C_{\pi}^(\Gamma)$ with $\pi \otimes \pi \subset \pi$, with Γ being a discrete group.*

PROOF. This follows indeed from the Peter-Weyl theory. \square

As another consequence of the above results, once again by basically following [98], we have the following result, dealing with functional analysis aspects:

THEOREM 7.36. *Let A_{full} be the enveloping C^* -algebra of \mathcal{A} , and let A_{red} be the quotient of A by the null ideal of the Haar integration. The following are then equivalent:*

- (1) *The Haar functional of A_{full} is faithful.*
- (2) *The projection map $A_{\text{full}} \rightarrow A_{\text{red}}$ is an isomorphism.*
- (3) *The counit map $\varepsilon : A_{\text{full}} \rightarrow \mathbb{C}$ factorizes through A_{red} .*
- (4) *We have $N \in \sigma(\text{Re}(\chi_u))$, the spectrum being taken inside A_{red} .*

If this is the case, we say that the underlying discrete quantum group Γ is amenable.

PROOF. This is well-known in the group dual case, $A = C^*(\Gamma)$, with Γ being a usual discrete group. In general, the result follows by adapting the group dual case proof:

(1) \iff (2) This simply follows from the fact that the GNS construction for the algebra A_{full} with respect to the Haar functional produces the algebra A_{red} .

(2) \iff (3) Here \implies is trivial, and conversely, a counit map $\varepsilon : A_{red} \rightarrow \mathbb{C}$ produces an isomorphism $A_{red} \rightarrow A_{full}$, via a formula of type $(\varepsilon \otimes id)\Phi$. See [98].

(3) \iff (4) Here \implies is clear, coming from $\varepsilon(N - Re(\chi(u))) = 0$, and the converse can be proved by doing some functional analysis. Once again, we refer here to [98]. \square

Summarizing, we have a nice theory of compact and discrete quantum groups, generalizing the basic theory of the usual compact and discrete groups. For the moment this is all that we need, but we will be back later, on several occasions, to the compact and discrete quantum groups, which more general theory.

Let us discuss now some “new” examples. Following Wang, we have:

THEOREM 7.37. *The following universal algebras are Woronowicz algebras,*

$$\begin{aligned} C(O_N^+) &= C^* \left((u_{ij})_{i,j=1,\dots,N} \mid u = \bar{u}, u^t = u^{-1} \right) \\ C(U_N^+) &= C^* \left((u_{ij})_{i,j=1,\dots,N} \mid u^* = u^{-1}, u^t = \bar{u}^{-1} \right) \end{aligned}$$

so the underlying spaces O_N^+, U_N^+ and O_N^*, U_N^* are compact quantum groups.

PROOF. This follows from the elementary fact that if a matrix $u = (u_{ij})$ is orthogonal or biunitary, then so must be the following matrices:

$$u_{ij}^\Delta = \sum_k u_{ik} \otimes u_{kj} \quad , \quad u_{ij}^\varepsilon = \delta_{ij} \quad , \quad u_{ij}^S = u_{ji}^*$$

Thus, we can indeed define morphisms Δ, ε, S as in Definition 7.28, by using the universal properties of $C(O_N^+)$, $C(U_N^+)$, and this gives the result. \square

Following Wang [92], further interesting examples can be introduced as follows:

THEOREM 7.38. *The following construction, where “magic” means formed of projections, which sum up to 1 on each row and column,*

$$C(S_N^+) = C^* \left((u_{ij})_{i,j=1,\dots,N} \mid u = \text{magic} \right)$$

produces a quantum group liberation of S_N . Moreover, the inclusion $S_N \subset S_N^+$ is an isomorphism at $N \leq 3$, but not at $N \geq 4$, where S_N^+ is not classical, nor finite.

PROOF. The quantum group assertion follows as in the proof of Theorem 7.37, from the fact that if $u = (u_{ij})$ is magic, then so must be the following matrices:

$$u_{ij}^{\Delta} = \sum_k u_{ik} \otimes u_{kj} \quad , \quad u_{ij}^{\varepsilon} = \delta_{ij} \quad , \quad u_{ij}^S = u_{ji}^*$$

Regarding now the last assertion, the study here is as follows:

Case $N = 2$. The result here is trivial, the 2×2 magic matrices being by definition as follows, with p being a projection:

$$U = \begin{pmatrix} p & 1-p \\ 1-p & p \end{pmatrix}$$

Indeed, this shows that the entries of a 2×2 magic matrix must pairwise commute, and so the algebra $C(S_2^+)$ follows to be commutative, which gives the result.

Case $N = 3$. This is more tricky, and we present here a short proof from [55]. By using the same abstract argument as in the $N = 2$ case, and by permuting rows and columns, it is enough to check that u_{11}, u_{22} commute. But this follows from:

$$\begin{aligned} u_{11}u_{22} &= u_{11}u_{22}(u_{11} + u_{12} + u_{13}) \\ &= u_{11}u_{22}u_{11} + u_{11}u_{22}u_{13} \\ &= u_{11}u_{22}u_{11} + u_{11}(1 - u_{21} - u_{23})u_{13} \\ &= u_{11}u_{22}u_{11} \end{aligned}$$

Indeed, by applying the involution to this, we obtain as well $u_{22}u_{11} = u_{11}u_{22}u_{11}$. Thus we have $u_{11}u_{22} = u_{22}u_{11}$, as desired.

Case $N = 4$. In order to prove our various claims about S_4^+ , consider the following matrix, with p, q being projections, on some infinite dimensional Hilbert space:

$$U = \begin{pmatrix} p & 1-p & 0 & 0 \\ 1-p & p & 0 & 0 \\ 0 & 0 & q & 1-q \\ 0 & 0 & 1-q & q \end{pmatrix}$$

This matrix is magic, and if we choose p, q as for the algebra $\langle p, q \rangle$ to be not commutative, and infinite dimensional, we conclude that $C(S_4^+)$ is not commutative and infinite dimensional as well, and in particular is not isomorphic to $C(S_4)$.

Case $N \geq 5$. Here we can use the standard embedding $S_4^+ \subset S_N^+$, obtained at the level of the corresponding magic matrices in the following way:

$$u \rightarrow \begin{pmatrix} u & 0 \\ 0 & 1_{N-4} \end{pmatrix}$$

Indeed, with this embedding in hand, the fact that S_4^+ is a non-classical, infinite compact quantum group implies that S_N^+ with $N \geq 5$ has these two properties as well. \square

The above result came as a surprise at the time of [92], and there has been a lot of work since then, in order to understand what the quantum permutations really are.

In what concerns us here, let us rather construct more examples of quantum groups. With the above result in hand, we can now introduce the quantum reflections:

THEOREM 7.39. *The following constructions produce compact quantum groups,*

$$C(H_N^+) = C^* \left((u_{ij})_{i,j=1,\dots,N} \mid u_{ij} = u_{ij}^*, (u_{ij}^2) = \text{magic} \right)$$

$$C(K_N^+) = C^* \left((u_{ij})_{i,j=1,\dots,N} \mid [u_{ij}, u_{ij}^*] = 0, (u_{ij}u_{ij}^*) = \text{magic} \right)$$

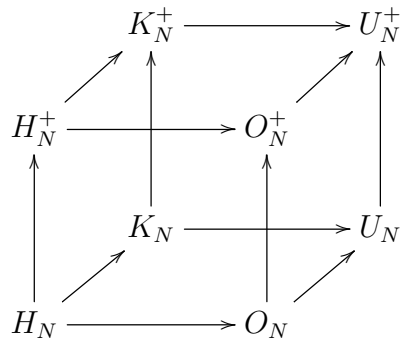
which appear as liberations of the reflection groups $H_N = \mathbb{Z}_2 \wr S_N$ and $K_N = \mathbb{T} \wr S_N$.

PROOF. This can be proved in the usual way, with the first assertion coming from the fact that if u satisfies the relations in the statement, then so do the matrices $u^\Delta, u^\varepsilon, u^S$, and with the second assertion coming as in the sphere case. See [7], [10]. \square

As before with the quantum permutations, there are many things that can be said about the quantum reflections. For instance the free hyperoctahedral group has a free wreath product decomposition, of type $H_N^+ = \mathbb{Z}_2 \wr_* S_N^+$, and the free complex reflection group decomposes as well, as $K_N^+ = \mathbb{T} \wr_* S_N^+$. We refer here to the literature.

As a conclusion to all this, we have a nice cubic diagram, as follows:

THEOREM 7.40. *We have basic quantum unitary and reflection groups,*



with the quantum groups on top appearing as liberations of those on the bottom.

PROOF. This follows indeed by putting together all the above results, and with the remark that we have ditched the symmetric group S_N and its free analogue S_N^+ , due to some technical reasons, that will become later on. \square

Summarizing, we have a beginning of theory here. It is possible to go well beyond this, first with a Haar measure result for the free spheres, and then with a generalization of all this to more general quotient spaces, of type G/H , eventually unifying all the quantum

spaces, spheres or quantum groups, that we have talked about in the above. However, this is quite technical, and in what follows, for an introduction, we have what we need.

7e. Exercises

There have been many things in this chapter, with several twists performed, around the obstacles met on the way. Generally speaking, all this is quite complicated, and fully understanding it would require doing lots of exercises, such as trying to go in directions which do not work, for making sure that these directions do not work indeed. For keeping things short, however, we will have only constructive exercises. Let us start with:

EXERCISE 7.41. *Work out the proof of the existence result for the Haar measure on a compact group G , as a particular case of the result proved for quantum groups.*

This is of course something very standard, the problem being that of eliminating algebras, linear forms and other functional analysis notions from the proof for the quantum groups, as to have in the end something talking about spaces, and measures on them.

Along the same lines, dealing this time with discrete groups, we have:

EXERCISE 7.42. *Clarify the construction of the discrete group algebras $C_\pi^*(\Gamma)$, using representations satisfying $\pi \otimes \pi \subset \pi$, again by adapting our quantum group knowledge.*

As before with the previous exercise, this should be something quite standard.

As a key exercise now, regarding the quantum permutations, we have:

EXERCISE 7.43. *Prove that, with a suitable formalism for quantum group actions, S_N^+ is indeed the quantum symmetry group of the set $X = \{1, \dots, N\}$.*

This is something quite tricky, and as a hint, the keyword is “counting measure”.

As another nice exercise now, regarding the free quantum groups, we have:

EXERCISE 7.44. *Clarify the definition of H_N, K_N , and the decomposition results*

$$H_N = \mathbb{Z}_2 \wr S_N \quad , \quad K_N = \mathbb{T} \wr S_N$$

and then do the same for H_N^+, K_N^+ , by establishing decomposition results of type

$$H_N^+ = \mathbb{Z}_2 \wr_* S_N^+ \quad , \quad K_N^+ = \mathbb{T} \wr_* S_N^+$$

with \wr_ being a suitable free wreath product operation.*

Finally, as a bonus exercise, try developing a kind of “noncommutative geometry” theory, converging all the spaces, classical and free, discussed in the present chapter, namely the classical and free spheres, the classical and free tori, and the classical and free quantum groups. Or at least start developing such a theory, and report on what you found.

CHAPTER 8

Integration theory

8a. Independence, freeness

We know that the commutative von Neumann algebras are the algebras of the form $A = L^\infty(X)$, with X being a measured space. In view of this, an arbitrary von Neumann algebra A can be thought of as being of the form $A = L^\infty(X)$, with X being a “quantum measured space”. In this picture, the integration over X corresponds to a functional $\varphi : A \rightarrow \mathbb{C}$, and of particular interest is the case where this functional is a trace:

$$\text{tr} : A \rightarrow \mathbb{C}$$

We have many examples of such algebras, and notably the group algebras $L(\Gamma)$, more generally the algebras $L^\infty(G)$ associated to the compact quantum groups G , which can be taken to be group duals, $G = \widehat{\Gamma}$, and even more generally, the algebras $L^\infty(G/H)$ associated to the subgroups $H \subset G$ of the compact quantum groups G .

We discuss here the integration theory for the von Neumann algebras coming with traces, $\text{tr} : A \rightarrow \mathbb{C}$, following [83], [84], [85], [86]. Let us start our discussion with:

DEFINITION 8.1. *Let A be a von Neumann algebra, given with a trace $\text{tr} : A \rightarrow \mathbb{C}$.*

- (1) *The elements $a \in A$ are called random variables.*
- (2) *The moments of such a variable are the numbers $M_k(a) = \text{tr}(a^k)$.*
- (3) *The law of such a variable is the functional $\mu : P \rightarrow \text{tr}(P(a))$.*

Here $k = \circ \bullet \bullet \circ \dots$ is by definition a colored integer, and the powers a^k are defined by multiplicativity and the usual formulae, namely:

$$a^\emptyset = 1 \quad , \quad a^\circ = a \quad , \quad a^\bullet = a^*$$

As for the polynomial P , this is a noncommuting $*$ -polynomial in one variable:

$$P \in \mathbb{C} \langle X, X^* \rangle$$

Observe that the law is uniquely determined by the moments, because:

$$P(X) = \sum_k \lambda_k X^k \implies \mu(P) = \sum_k \lambda_k M_k(a)$$

Generally speaking, the above definition is something quite abstract, but there is no other way of doing things, at least at this level of generality.

However, in the special case where our variable $a \in A$ is self-adjoint, or more generally normal, the theory simplifies, and we recover more familiar objects, as follows:

PROPOSITION 8.2. *The law of a normal variable $a \in A$ can be identified with the corresponding spectral measure $\mu \in \mathcal{P}(\mathbb{C})$, according to the following formula,*

$$\text{tr}(f(a)) = \int_{\sigma(a)} f(x) d\mu(x)$$

valid for any $f \in L^\infty(\sigma(a))$, coming from the measurable functional calculus. In the self-adjoint case the spectral measure is real, $\mu \in \mathcal{P}(\mathbb{R})$.

PROOF. This is something that we know well, either from chapter 5 above, or simply from chapter 3 above, from the spectral theorem for the normal operators. \square

Summarizing, we have so far a beginning of a theory, generalizing that of the usual probability spaces (X, μ) . A quantum probability space corresponds by definition to a pair (A, tr) , according to the formulae $A = L^\infty(X)$ and $\text{tr}(f) = \int_X f(x) d\mu(x)$. We can talk about moments and laws, although a bit abstractly. And when A is commutative, or at least our variables are normal, we recover in this way the usual probability theory.

Let us discuss now the independence, and its noncommutative versions. As a starting point, we have the following notion, generalizing the classical independence notion:

DEFINITION 8.3. *We call two subalgebras $B, C \subset A$ independent when the following condition is satisfied, for any $x \in B$ and $y \in C$:*

$$\text{tr}(xy) = \text{tr}(x)\text{tr}(y)$$

Equivalently, the following condition must be satisfied, for any $x \in B$ and $y \in C$:

$$\text{tr}(x) = \text{tr}(y) = 0 \implies \text{tr}(xy) = 0$$

Also, two variables $b, c \in A$ are called independent when the algebras that they generate $B = \langle b \rangle$ and $C = \langle c \rangle$ are independent, in the above sense.

Observe that the above two conditions for independence are indeed equivalent, as shown by the following computation, with the convention $x' = x - \text{tr}(x)$:

$$\begin{aligned} \text{tr}(xy) &= \text{tr}[(x' + \text{tr}(x))(y' + \text{tr}(y))] \\ &= \text{tr}(x'y') + \text{tr}(x')\text{tr}(y) + \text{tr}(x)\text{tr}(y') + \text{tr}(x)\text{tr}(y) \\ &= \text{tr}(x'y') + \text{tr}(x)\text{tr}(y) \\ &= \text{tr}(x)\text{tr}(y) \end{aligned}$$

It is possible to develop some theory here, but all this is ultimately not very interesting. As a much more interesting notion now, we have the freeness:

DEFINITION 8.4. *Given a pair (A, tr) , we call two subalgebras $B, C \subset A$ free when the following condition is satisfied, for any $x_i \in B$ and $y_i \in C$:*

$$tr(x_i) = tr(y_i) = 0 \implies tr(x_1 y_1 x_2 y_2 \dots) = 0$$

Also, two noncommutative random variables $b, c \in A$ are called free when the C^ -algebras $B = \langle b \rangle, C = \langle c \rangle$ that they generate inside A are free, in the above sense.*

In short, freeness appears by definition as a kind of “free analogue” of independence, taking into account the fact that the variables do not necessarily commute.

As a first observation, there is a certain lack of symmetry between Definition 8.3 and Definition 8.4, because the latter does not include an explicit formula for:

$$tr(x_1 y_1 x_2 y_2 \dots)$$

This is simply due to the fact that the formula in the free case is something quite complicated, and not really explicit, the precise result being as follows:

PROPOSITION 8.5. *If $B, C \subset A$ are free, the restriction of tr to $\langle B, C \rangle$ can be computed in terms of the restrictions of tr to B, C . To be more precise, we have*

$$tr(x_1 y_1 x_2 y_2 \dots) = P\left(\{tr(x_{i_1} x_{i_2} \dots)\}_i, \{tr(y_{j_1} y_{j_2} \dots)\}_j\right)$$

where P is certain polynomial, depending on the length of $x_1 y_1 x_2 y_2 \dots$, having as variables the traces of products $x_{i_1} x_{i_2} \dots$ and $y_{j_1} y_{j_2} \dots$, with $i_1 < i_2 < \dots$ and $j_1 < j_2 < \dots$.

PROOF. As a first illustration, we have the following formula, with proof exactly as in the independence case, by using the computation performed after Definition 8.3:

$$tr(xy) = tr(x)tr(y)$$

In general, we can start our computation of $tr(x_1 y_1 x_2 y_2 \dots)$ as follows:

$$tr(x_1 y_1 x_2 y_2 \dots) = tr[(x_1 + (tr(x_1) - x_1))(y_1 + (tr(y_1) - y_1)) \\ (x_2 + (tr(x_2) - x_2))(y_2 + (tr(y_2) - y_2)) \dots]$$

Thus, we are led to a kind of recurrence, and this gives the result. \square

Let us discuss now some examples of independence and freeness. We first have:

THEOREM 8.6. *Given two algebras (A, tr) and (B, tr) , the following hold:*

- (1) *A, B are independent inside their tensor product $A \otimes B$, endowed with its canonical tensor product trace, given on basic tensors by $tr(a \otimes b) = tr(a)tr(b)$.*
- (2) *A, B are free inside their free product $A * B$, endowed with its canonical free product trace, given by the formulae in Proposition 8.5.*

PROOF. Both the assertions are clear from definitions, as follows:

(1) This is something elementary, which is clear with either of the definitions of the independence given above.

(2) This is clear from definitions, the only point being that of showing that the notion of freeness, or the recurrence formulae in Proposition 8.5, can be used in order to construct a canonical free product trace, on the free product of the two algebras involved:

$$tr : A * B \rightarrow \mathbb{C}$$

But this can be checked for instance by using a GNS construction. Indeed, consider the GNS constructions for the algebras (A, tr) and (B, tr) :

$$A \rightarrow B(l^2(A))$$

$$B \rightarrow B(l^2(B))$$

By taking the free product of these representations, we obtain a representation as follows, with the $*$ symbol on the right being a free product of pointed Hilbert spaces:

$$A * B \rightarrow B(l^2(A) * l^2(B))$$

Now by composing with the linear form $T \rightarrow \langle T\xi, \xi \rangle$, where $\xi = 1_A = 1_B$ is the common pointed vector of $l^2(A)$ and $l^2(B)$, we obtain a linear form, as follows:

$$tr : A * B \rightarrow \mathbb{C}$$

It is routine then to check that tr is indeed a trace, and this is the “canonical free product trace” from the statement. Then, an elementary computation shows that A, B are indeed free inside $A * B$, with respect to this trace, and this finishes the proof. \square

More concretely now, both the independence and the freeness are very nicely modelled inside group von Neumann algebras. We have here the following result:

THEOREM 8.7. *We have the following results, valid for group algebras:*

- (1) $L(\Gamma), L(\Lambda)$ are independent inside $L(\Gamma \times \Lambda)$.
- (2) $L(\Gamma), L(\Lambda)$ are free inside $L(\Gamma * \Lambda)$.

PROOF. In order to prove these results, we have two possible methods:

(1) We can either use the general results in Theorem 8.6 above, along with the following two isomorphisms, which are both standard:

$$L(\Gamma \times \Lambda) = L(\Lambda) \otimes L(\Gamma)$$

$$L(\Gamma * \Lambda) = L(\Lambda) * L(\Gamma)$$

(2) Or, we can prove this directly, by using the fact that each group algebra is spanned by the corresponding group elements. Indeed, this shows that it is enough to check the independence and freeness formulae on group elements, which is in turn trivial. \square

8b. Limiting theorems

We have already seen limiting theorems in classical probability, in chapter 6 above. In order to deal now with freeness, let us develop some tools. First, we have:

PROPOSITION 8.8. *We have a well-defined operation \boxplus , given by*

$$\mu_a \boxplus \mu_b = \mu_{a+b}$$

with a, b being free, called free convolution.

PROOF. The only verification which is needed is that of the fact that if a, b are free, then the distribution μ_{a+b} depends only on the distributions μ_a, μ_b . But for this purpose, we can use the general formula from Proposition 8.5 above, namely:

$$\text{tr}(x_1 y_1 x_2 y_2 \dots) = P\left(\{\text{tr}(x_{i_1} x_{i_2} \dots)\}_i, \{\text{tr}(y_{j_1} y_{j_2} \dots)\}_j\right)$$

Here P is certain polynomial, depending on the length of $x_1 y_1 x_2 y_2 \dots$, having as variables the traces of products $x_{i_1} x_{i_2} \dots$ and $y_{j_1} y_{j_2} \dots$, with $i_1 < i_2 < \dots$ and $j_1 < j_2 < \dots$.

Now by plugging in arbitrary powers of a, b as variables x_i, y_j , we obtain a family of formulae of the following type, with Q being certain polynomials:

$$\text{tr}(a^{k_1} b^{l_1} a^{k_2} b^{l_2} \dots) = P\left(\{\text{tr}(a^k)\}_k, \{\text{tr}(b^l)\}_l\right)$$

Thus the moments of $a+b$ depend only on the moments of a, b , and the same argument shows that the same holds for $*$ -moments, and this gives the result. \square

In order to advance now, we would need an analogue of the Fourier transform, or rather of the log of the Fourier transform. Quite remarkably, such a transform exists indeed, the precise result here, due to Voiculescu [83], being as follows:

THEOREM 8.9. *Given a probability measure μ , define its R -transform as follows:*

$$G_\mu(\xi) = \int_{\mathbb{R}} \frac{d\mu(t)}{\xi - t} \implies G_\mu\left(R_\mu(\xi) + \frac{1}{\xi}\right) = \xi$$

The free convolution operation is then linearized by the R -transform.

PROOF. This is something quite tricky, the idea being as follows:

(1) In order to model the free convolution, the best is to use creation operators on free Fock spaces, corresponding to the monoid von Neumann algebras $L(\mathbb{N}^{*k})$. Indeed, we have some freeness here, a bit in the same way as in the free group algebras $L(F_k)$.

(2) The point now, motivating this choice, is that the variables of type $S^* + f(S)$, with $S \in L(\mathbb{N})$ being the shift, and with $f \in \mathbb{C}[X]$ being an arbitrary polynomial, are easily seen to model in moments all the possible distributions $\mu : \mathbb{C}[X] \rightarrow \mathbb{C}$.

(2) Now let $f, g \in \mathbb{C}[X]$ and consider the variables $S^* + f(S)$ and $T^* + g(T)$, where $S, T \in L(\mathbb{N} * \mathbb{N})$ are the shifts corresponding to the generators of $\mathbb{N} * \mathbb{N}$. These variables are free, and by using a 45° argument, their sum has the same law as $S^* + (f + g)(S)$.

(3) Thus the operation $\mu \rightarrow f$ linearizes the free convolution. We are therefore left with a computation inside $L(\mathbb{N})$, which is elementary, and whose conclusion is that $R_\mu = f$ can be recaptured from μ via the Cauchy transform G_μ , as in the statement. \square

With the above linearization technology in hand, we can now establish the following remarkable free analogue of the CLT, also due to Voiculescu [83]:

THEOREM 8.10. *Given self-adjoint variables x_1, x_2, x_3, \dots , which are f.i.d., centered, with variance $t > 0$, we have, with $n \rightarrow \infty$, in moments,*

$$\frac{1}{\sqrt{n}} \sum_{i=1}^n x_i \sim \gamma_t$$

where γ_t is the Wigner semicircle law of parameter t , namely

$$\gamma_t = \frac{1}{2\pi t} \sqrt{4t^2 - x^2} dx$$

also called free Gaussian law of parameter t .

PROOF. We follow the same idea as in the proof of the CLT:

(1) At $t = 1$, the R -transform of the variable in the statement can be computed by using the linearization property from Theorem 8.9, and is given by:

$$R(\xi) = nR_x \left(\frac{\xi}{\sqrt{n}} \right) \simeq \xi$$

(2) On the other hand, some standard computations show that the Cauchy transform of the Wigner law γ_1 satisfies the following equation:

$$G_{\gamma_1} \left(\xi + \frac{1}{\xi} \right) = \xi$$

Thus, by using Theorem 8.9, we have the following formula:

$$R_{\gamma_1}(\xi) = \xi$$

(3) We conclude that the laws in the statement have the same R -transforms, and so they are equal. The passage to the general case, $t > 0$, is routine, by dilation. \square

Let us discuss as well the complex versions of the limiting theorems that we have. In the classical case, the starting definition is as follows:

DEFINITION 8.11. *The complex Gaussian law of parameter $t > 0$ is*

$$G_t = \text{law} \left(\frac{1}{\sqrt{2}}(a + ib) \right)$$

where a, b are independent, each following the law g_t .

As in the real case, these measures form convolution semigroups:

PROPOSITION 8.12. *The complex Gaussian laws have the property*

$$G_s * G_t = G_{s+t}$$

for any $s, t > 0$, and so they form a convolution semigroup.

PROOF. This follows indeed from the real result, namely $g_s * g_t = g_{s+t}$, established above, simply by taking real and imaginary parts. \square

We have the following complex analogue of the CLT:

THEOREM 8.13 (CCLT). *Given complex random variables $f_1, f_2, f_3, \dots \in L^\infty(X)$ which are i.i.d., centered, and with variance $t > 0$, we have, with $n \rightarrow \infty$, in moments,*

$$\frac{1}{\sqrt{n}} \sum_{i=1}^n f_i \sim G_t$$

where G_t is the complex Gaussian law of parameter t .

PROOF. This follows indeed from the real CLT, established above, simply by taking real and imaginary parts. \square

In the free case now, we can develop a similar theory. First, we have:

DEFINITION 8.14. *The Voiculescu circular law of parameter $t > 0$ is given by*

$$\Gamma_t = \text{law} \left(\frac{1}{\sqrt{2}}(a + ib) \right)$$

where a, b are free, each following the Wigner semicircle law γ_t .

In other words, the passage $\gamma_t \rightarrow \Gamma_t$ is by definition entirely similar to the passage $g_t \rightarrow G_t$ from the classical case, by taking real and imaginary parts. As before in other similar situations, the fact that Γ_t is indeed well-defined is clear from definitions.

Let us start with a number of straightforward results, obtained by complexifying the free probability theory that we have. As a first result, we have:

PROPOSITION 8.15. *The Voiculescu circular laws have the property*

$$\Gamma_s \boxplus \Gamma_t = \Gamma_{s+t}$$

so they form a 1-parameter semigroup with respect to free convolution.

PROOF. This follows from our previous result stating that the Wigner laws γ_t have the free semigroup convolution property, by taking real and imaginary parts. \square

We have the following free analogue of the complex CLT:

THEOREM 8.16 (Free CCLT). *Given random variables x_1, x_2, x_3, \dots which are f.i.d., centered, with variance $t > 0$, we have, with $n \rightarrow \infty$, in moments,*

$$\frac{1}{\sqrt{n}} \sum_{i=1}^n x_i \sim \Gamma_t$$

where Γ_t is the Voiculescu circular law of parameter t .

PROOF. This follows indeed from the free CLT, established before, by taking real and imaginary parts. Indeed, let us write:

$$x_i = \frac{1}{\sqrt{2}}(y_i + iz_i)$$

The variables y_i satisfy then the assumptions of the free CLT, and so their rescaled averages converge to a semicircle law γ_t , and the same happens for the variables z_i :

$$\frac{1}{\sqrt{n}} \sum_{i=1}^n y_i \sim \gamma_t \quad , \quad \frac{1}{\sqrt{n}} \sum_{i=1}^n z_i \sim \gamma_t$$

Now since the two limiting semicircle laws that we obtain in this way are free, their rescaled sum is circular, in the sense of Definition 8.14, and this gives the result. \square

We should mention that both the complex Gaussian variables and the Voiculescu circular laws have a quite interesting combinatorics. We will be back to this later.

Now that we are done with the basic results in continuous case, let us discuss the discrete case. We can establish a free version of the PLT, as follows:

THEOREM 8.17. *The following limit converges, for any $t > 0$,*

$$\lim_{n \rightarrow \infty} \left(\left(1 - \frac{t}{n} \right) \delta_0 + \frac{t}{n} \delta_1 \right)^{\boxplus n}$$

and we obtain the Marchenko-Pastur law of parameter t ,

$$\pi_t = \max(1 - t, 0) \delta_0 + \frac{\sqrt{4t - (x - 1 - t)^2}}{2\pi x} dx$$

also called free Poisson law of parameter t .

PROOF. Let μ be the measure in the statement, appearing under the convolution sign. The Cauchy transform of this measure is elementary to compute, given by:

$$G_\mu(\xi) = \left(1 - \frac{t}{n} \right) \frac{1}{\xi} + \frac{t}{n} \cdot \frac{1}{\xi - 1}$$

By using Theorem 8.9 above, we want to compute the following R -transform:

$$R = R_{\mu^{\boxplus n}}(y) = nR_{\mu}(y)$$

We know that the equation for this function R is as follows:

$$\left(1 - \frac{t}{n}\right) \frac{1}{y^{-1} + R/n} + \frac{t}{n} \cdot \frac{1}{y^{-1} + R/n - 1} = y$$

By multiplying by n/y , this equation can be written as follows:

$$\frac{t + yR}{1 + yR/n} = \frac{t}{1 + yR/n - y}$$

With $n \rightarrow \infty$ we obtain from this the following formula:

$$t + yR = \frac{t}{1 - y}$$

Thus we have the following formula, in the $n \rightarrow \infty$ limit:

$$R = \frac{t}{1 - y}$$

But this being the R -transform of π_t , via some calculus, we are done. \square

Let us record as well the following result:

THEOREM 8.18. *The Wigner and Marchenko-Pastur laws have the semigroup property*

$$\gamma_{s+t} = \gamma_s \boxplus \gamma_t$$

$$\pi_{s+t} = \pi_s \boxplus \pi_t$$

with respect to the Voiculescu free convolution operation \boxplus .

PROOF. This follows either from the CLT and PLT, or from Theorem 8.9 and the explicit formulae of the R -transforms of γ_t, π_t found in the proofs above, namely:

$$R_{\gamma_t}(\xi) = t\xi$$

$$R_{\pi_t}(\xi) = \frac{t}{1 - \xi}$$

Indeed, in each case the R -transform is linear in t , and this gives the result. \square

Observe that the above results give a more conceptual explanation for the random matrix results obtained in chapter 6 above, or rather for the limiting laws obtained there. It is possible to further build on all this, and we refer here to the literature.

Let us develop now a number of further limiting theorems, classical and free. We have the following definition, extending the Poisson limit theory developed above:

DEFINITION 8.19. Associated to any compactly supported positive measure ρ on \mathbb{R} are the probability measures

$$p_\rho = \lim_{n \rightarrow \infty} \left(\left(1 - \frac{c}{n}\right) \delta_0 + \frac{1}{n} \rho \right)^{*n}$$

$$\pi_\rho = \lim_{n \rightarrow \infty} \left(\left(1 - \frac{c}{n}\right) \delta_0 + \frac{1}{n} \rho \right)^{\boxplus n}$$

where $c = \text{mass}(\rho)$, called compound Poisson and compound free Poisson laws.

In what follows we will be interested in the case where ρ is discrete, as is for instance the case for $\rho = \delta_t$ with $t > 0$, which produces the Poisson and free Poisson laws. The following result allows one to detect compound Poisson/free Poisson laws:

PROPOSITION 8.20. For $\rho = \sum_{i=1}^s c_i \delta_{z_i}$ with $c_i > 0$ and $z_i \in \mathbb{R}$, we have

$$F_{p_\rho}(y) = \exp \left(\sum_{i=1}^s c_i (e^{iyz_i} - 1) \right)$$

$$R_{\pi_\rho}(y) = \sum_{i=1}^s \frac{c_i z_i}{1 - yz_i}$$

where F, R denote respectively the Fourier transform, and Voiculescu's R -transform.

PROOF. Let μ_n be the measure appearing in Definition 8.19, under the convolution signs. In the classical case, we have the following computation:

$$\begin{aligned} F_{\mu_n}(y) &= \left(1 - \frac{c}{n}\right) + \frac{1}{n} \sum_{i=1}^s c_i e^{iyz_i} \\ \implies F_{\mu_n^{*n}}(y) &= \left(\left(1 - \frac{c}{n}\right) + \frac{1}{n} \sum_{i=1}^s c_i e^{iyz_i} \right)^n \\ \implies F_{p_\rho}(y) &= \exp \left(\sum_{i=1}^s c_i (e^{iyz_i} - 1) \right) \end{aligned}$$

In the free case now, we use a similar method. The Cauchy transform of μ_n is:

$$G_{\mu_n}(\xi) = \left(1 - \frac{c}{n}\right) \frac{1}{\xi} + \frac{1}{n} \sum_{i=1}^s \frac{c_i}{\xi - z_i}$$

Consider now the R -transform of the measure $\mu_n^{\boxplus n}$, which is given by:

$$R_{\mu_n^{\boxplus n}}(y) = n R_{\mu_n}(y)$$

The above formula of G_{μ_n} shows that the equation for $R = R_{\mu_n^{\boxplus}}$ is as follows:

$$\begin{aligned} & \left(1 - \frac{c}{n}\right) \frac{1}{y^{-1} + R/n} + \frac{1}{n} \sum_{i=1}^s \frac{c_i}{y^{-1} + R/n - z_i} = y \\ \implies & \left(1 - \frac{c}{n}\right) \frac{1}{1 + yR/n} + \frac{1}{n} \sum_{i=1}^s \frac{c_i}{1 + yR/n - yz_i} = 1 \end{aligned}$$

Now multiplying by n , rearranging the terms, and letting $n \rightarrow \infty$, we get:

$$\begin{aligned} & \frac{c + yR}{1 + yR/n} = \sum_{i=1}^s \frac{c_i}{1 + yR/n - yz_i} \\ \implies & c + yR_{\pi_\rho}(y) = \sum_{i=1}^s \frac{c_i}{1 - yz_i} \\ \implies & R_{\pi_\rho}(y) = \sum_{i=1}^s \frac{c_i z_i}{1 - yz_i} \end{aligned}$$

This finishes the proof in the free case, and we are done. \square

We also have the following result, providing an alternative to Definition 8.19:

THEOREM 8.21. *For $\rho = \sum_{i=1}^s c_i \delta_{z_i}$ with $c_i > 0$ and $z_i \in \mathbb{R}$, we have*

$$p_\rho / \pi_\rho = \text{law} \left(\sum_{i=1}^s z_i \alpha_i \right)$$

where the variables α_i are Poisson/free Poisson(c_i), independent/free.

PROOF. Let α be the sum of Poisson/free Poisson variables in the statement. We will show that the Fourier/ R -transform of α is given by the formulae in Proposition 8.20.

Indeed, by using some well-known Fourier transform formulae, we have:

$$\begin{aligned} F_{\alpha_i}(y) = \exp(c_i(e^{iy} - 1)) & \implies F_{z_i \alpha_i}(y) = \exp(c_i(e^{iyz_i} - 1)) \\ & \implies F_\alpha(y) = \exp \left(\sum_{i=1}^s c_i (e^{iyz_i} - 1) \right) \end{aligned}$$

Also, by using some well-known R -transform formulae, we have:

$$\begin{aligned} R_{\alpha_i}(y) = \frac{c_i}{1 - y} & \implies R_{z_i \alpha_i}(y) = \frac{c_i z_i}{1 - yz_i} \\ & \implies R_\alpha(y) = \sum_{i=1}^s \frac{c_i z_i}{1 - yz_i} \end{aligned}$$

Thus we have indeed the same formulae as those in Proposition 8.20. \square

In what follows we will be interested in the main examples of classical and free compound Poisson laws, which are constructed as follows:

DEFINITION 8.22. *The Bessel and free Bessel laws are the compound Poisson laws*

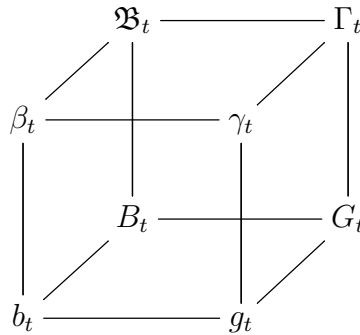
$$b_t^s = p_{t\varepsilon_s} \quad , \quad \beta_t^s = \pi_{t\varepsilon_s}$$

where ε_s is the uniform measure on the s -th roots unity. In particular:

- (1) At $s = 1$ we obtain the usual Poisson and free Poisson laws, p_t, π_t .
- (2) At $s = 2$ we obtain the “real” Bessel and free Bessel laws, denoted b_t, β_t .
- (3) At $s = \infty$ we obtain the “complex” Bessel and free Bessel laws, denoted B_t, \mathfrak{B}_t .

There is a lot of theory regarding these laws, and we refer here to [7], where these laws were introduced. Let us just record here the following result, that we will need:

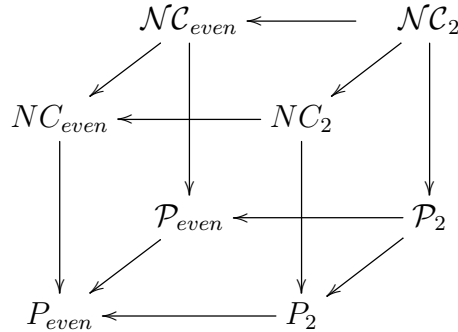
THEOREM 8.23. *The moments of the various central limiting measures, namely*



are always given by the same formula, involving partitions, namely

$$M_k = \sum_{\pi \in D(k)} t^{|\pi|}$$

with the sets of partitions $D(k)$ in question being respectively



where 2 stands for pairings, even stands for “with even blocks”, NC stands for noncrossing, calligraphic stands for matching, and with $|\cdot|$ being as usual the number of blocks.

PROOF. This is something that we know well for the Gaussian laws g_t , and from this we can deduce the result for the complex Gaussian laws G_t too. We also know this for the Poisson laws p_t , with these laws actually not appearing in the above diagram, with the corresponding set of partitions being the set P of all partitions. Finally, we met the formula in the statement for the Wigner laws γ_t and the Marchenko-Pastur laws π_t , in chapter 6 above, in the random matrix context. The proof for the remaining laws is similar, by using calculus and combinatorics, and for full details, we refer to [7]. \square

There is an obvious similarity between the quantum group cube at the end of chapter 7, and the cube from Theorem 8.23, and understanding this will be our next objective.

8c. Diagrams, easiness

Let us go back to quantum groups, and to the Peter-Weyl theory from chapter 7. In order to further build on this, and reach to combinatorics, the idea will be, as in the Lie group case, to “linearize”. In the present setting we cannot really do geometry and talk about Lie algebras, but we can instead talk about tensor categories, as follows:

DEFINITION 8.24. *The Tannakian category associated to a Woronowicz algebra (A, u) is the collection $C = (C(k, l))$ of vector spaces*

$$C(k, l) = \text{Hom}(u^{\otimes k}, u^{\otimes l})$$

where the corepresentations $u^{\otimes k}$ with $k = \circ \bullet \bullet \circ \dots$ colored integer, defined by

$$u^{\otimes \emptyset} = 1 \quad , \quad u^{\otimes \circ} = u \quad , \quad u^{\otimes \bullet} = \bar{u}$$

and multiplicativity, $u^{\otimes kl} = u^{\otimes k} \otimes u^{\otimes l}$, are the Peter-Weyl corepresentations.

As a key remark, the fact that $u \in M_N(A)$ is biunitary translates into the following conditions, where $R: \mathbb{C} \rightarrow \mathbb{C}^N \otimes \mathbb{C}^N$ is the linear map given by $R(1) = \sum_i e_i \otimes e_i$:

$$R \in \text{Hom}(1, u \otimes \bar{u}) \quad , \quad R \in \text{Hom}(1, \bar{u} \otimes u)$$

$$R^* \in \text{Hom}(u \otimes \bar{u}, 1) \quad , \quad R^* \in \text{Hom}(\bar{u} \otimes u, 1)$$

We are therefore led to the following abstract definition:

DEFINITION 8.25. *Let H be a finite dimensional Hilbert space. A tensor category over H is a collection $C = (C(k, l))$ of subspaces*

$$C(k, l) \subset \mathcal{L}(H^{\otimes k}, H^{\otimes l})$$

satisfying the following conditions:

- (1) $S, T \in C$ implies $S \otimes T \in C$.
- (2) If $S, T \in C$ are composable, then $ST \in C$.
- (3) $T \in C$ implies $T^* \in C$.
- (4) Each $C(k, k)$ contains the identity operator.
- (5) $C(\emptyset, \circ \bullet)$ and $C(\emptyset, \bullet \circ)$ contain the operator $R: 1 \rightarrow \sum_i e_i \otimes e_i$.

We have the following result, summarizing our knowledge so far:

PROPOSITION 8.26. *To any Woronowicz algebra (A, u) , with fundamental corepresentation $u \in M_N(A)$, we can associate the following tensor category over \mathbb{C}^N :*

$$C_A = \text{Hom}(u^{\otimes k}, u^{\otimes l})$$

Conversely, to any tensor category $C = (C(k, l))$ over \mathbb{C}^N , in the above sense, we can associate the following Woronowicz algebra:

$$A_C = C^* \left((u_{ij})_{i,j=1,\dots,N} \middle| T \in \text{Hom}(u^{\otimes k}, u^{\otimes l}), \forall k, l, \forall T \in C(k, l) \right)$$

PROOF. Here the first assertion follows from the above discussion, and the second assertion is standard too, because the relations $T \in \text{Hom}(u^{\otimes k}, u^{\otimes l})$ determine a Hopf ideal, in the sense that they allow the construction of Δ, ε, S as in chapter 7 above. \square

With the above constructions in hand, we have the following result:

THEOREM 8.27. *The Tannakian duality constructions*

$$C \rightarrow A_C \quad , \quad A \rightarrow C_A$$

are inverse to each other, modulo identifying full and reduced versions.

PROOF. The idea is that we have an obvious inclusion $C \subset C_{A_C}$, for any algebra A , and so we are left with proving that we have $C_{A_C} \subset C$, for any Tannakian category C . But this follows from a long series of algebraic manipulations, including in particular an application of the von Neumann bicommutant theorem, and for details here we refer to [56], and also to [99], where this result was first proved, by using other methods. \square

As a first application, solving a theoretical problem that we have left open in chapter 7, let us go back to the standard embedding $G \subset S_{\mathbb{C},+}^{N^2-1}$, given by $x_{ij} = u_{ij}/\sqrt{N}$. By using Theorem 8.27, we see that G is an algebraic manifold, as previously claimed.

More concretely now, and back to integration questions, we can go take our basic geometric objects, from chapter 7 above, and their various extensions, and write a Weingarten integration formula for them. Let us start with the quantum group case. The first result here, coming from Peter-Weyl theory and basic linear algebra, is as follows:

THEOREM 8.28. *Assuming that $A = C(G)$ has Tannakian category $C = (C(k, l))$, the Haar integration over G is given by the Weingarten type formula*

$$\int_G u_{i_1 j_1}^{e_1} \dots u_{i_k j_k}^{e_k} = \sum_{\pi, \sigma \in D_k} \delta_\pi(i) \delta_\sigma(j) W_k(\pi, \sigma)$$

for any colored integer $k = e_1 \dots e_k$ and any multi-indices i, j , where D_k is a linear basis of $C(\emptyset, k)$, $\delta_\pi(i) = \langle \pi, e_{i_1} \otimes \dots \otimes e_{i_k} \rangle$, and $W_k = G_k^{-1}$, with $G_k(\pi, \sigma) = \langle \pi, \sigma \rangle$.

PROOF. We know from chapter 7 above that the integrals in the statement form altogether the orthogonal projection P^k onto the following space:

$$\text{Fix}(u^{\otimes k}) = \text{span}(D_k)$$

Consider now the following linear map, with $D_k = \{\xi_k\}$ being as in the statement:

$$E(x) = \sum_{\pi \in D_k} \langle x, \xi_\pi \rangle \xi_\pi$$

By a standard linear algebra computation, it follows that we have $P = WE$, where W is the inverse on $\text{span}(T_\pi | \pi \in D_k)$ of the restriction of E . But this restriction is the linear map given by G_k , and so W is the linear map given by W_k , and this gives the result. \square

In practice now, we must combine the above result with Brauer type results, for the specific quantum groups that we are interested in. Let us start with:

DEFINITION 8.29. *Let $P(k, l)$ be the set of partitions between an upper colored integer k , and a lower colored integer l . A collection of subsets*

$$D = \bigsqcup_{k,l} D(k, l)$$

with $D(k, l) \subset P(k, l)$ is called a category of partitions when it has the following properties:

- (1) *Stability under the horizontal concatenation, $(\pi, \sigma) \rightarrow [\pi\sigma]$.*
- (2) *Stability under vertical concatenation $(\pi, \sigma) \rightarrow \begin{bmatrix} \sigma \\ \pi \end{bmatrix}$, with matching middle symbols.*
- (3) *Stability under the upside-down turning $*$, with switching of colors, $\circ \leftrightarrow \bullet$.*
- (4) *Each set $P(k, k)$ contains the identity partition $|| \dots ||$.*
- (5) *The sets $P(\emptyset, \circ\bullet)$ and $P(\emptyset, \bullet\circ)$ both contain the semicircle \cap .*

Observe the similarity with Definition 8.25. In fact Definition 8.29 is a delinearized version of Definition 8.25, the relation with the Tannakian categories coming from:

PROPOSITION 8.30. *Each partition $\pi \in P(k, l)$ produces a linear map*

$$T_\pi : (\mathbb{C}^N)^{\otimes k} \rightarrow (\mathbb{C}^N)^{\otimes l}$$

given by the following formula, where e_1, \dots, e_N is the standard basis of \mathbb{C}^N ,

$$T_\pi(e_{i_1} \otimes \dots \otimes e_{i_k}) = \sum_{j_1 \dots j_l} \delta_\pi \begin{pmatrix} i_1 & \dots & i_k \\ j_1 & \dots & j_l \end{pmatrix} e_{j_1} \otimes \dots \otimes e_{j_l}$$

and with the Kronecker type symbols $\delta_\pi \in \{0, 1\}$ depending on whether the indices fit or not. The assignment $\pi \rightarrow T_\pi$ is categorical, in the sense that we have

$$\begin{aligned} T_\pi \otimes T_\sigma &= T_{[\pi\sigma]} \\ T_\pi T_\sigma &= N^{c(\pi, \sigma)} T_{\begin{bmatrix} \sigma \\ \pi \end{bmatrix}} \\ T_\pi^* &= T_{\pi^*} \end{aligned}$$

where $c(\pi, \sigma)$ are certain integers, coming from the erased components in the middle.

PROOF. This is something very elementary, as follows:

(1) The concatenation axiom follows from the following computation:

$$\begin{aligned}
& (T_\pi \otimes T_\sigma)(e_{i_1} \otimes \dots \otimes e_{i_p} \otimes e_{k_1} \otimes \dots \otimes e_{k_r}) \\
&= \sum_{j_1 \dots j_q} \sum_{l_1 \dots l_s} \delta_\pi \begin{pmatrix} i_1 & \dots & i_p \\ j_1 & \dots & j_q \end{pmatrix} \delta_\sigma \begin{pmatrix} k_1 & \dots & k_r \\ l_1 & \dots & l_s \end{pmatrix} e_{j_1} \otimes \dots \otimes e_{j_q} \otimes e_{l_1} \otimes \dots \otimes e_{l_s} \\
&= \sum_{j_1 \dots j_q} \sum_{l_1 \dots l_s} \delta_{[\pi\sigma]} \begin{pmatrix} i_1 & \dots & i_p & k_1 & \dots & k_r \\ j_1 & \dots & j_q & l_1 & \dots & l_s \end{pmatrix} e_{j_1} \otimes \dots \otimes e_{j_q} \otimes e_{l_1} \otimes \dots \otimes e_{l_s} \\
&= T_{[\pi\sigma]}(e_{i_1} \otimes \dots \otimes e_{i_p} \otimes e_{k_1} \otimes \dots \otimes e_{k_r})
\end{aligned}$$

(2) The composition axiom follows from the following computation:

$$\begin{aligned}
& T_\pi T_\sigma(e_{i_1} \otimes \dots \otimes e_{i_p}) \\
&= \sum_{j_1 \dots j_q} \delta_\sigma \begin{pmatrix} i_1 & \dots & i_p \\ j_1 & \dots & j_q \end{pmatrix} \sum_{k_1 \dots k_r} \delta_\pi \begin{pmatrix} j_1 & \dots & j_q \\ k_1 & \dots & k_r \end{pmatrix} e_{k_1} \otimes \dots \otimes e_{k_r} \\
&= \sum_{k_1 \dots k_r} N^{c(\pi, \sigma)} \delta_{[\pi]} \begin{pmatrix} i_1 & \dots & i_p \\ k_1 & \dots & k_r \end{pmatrix} e_{k_1} \otimes \dots \otimes e_{k_r} \\
&= N^{c(\pi, \sigma)} T_{[\pi]}(e_{i_1} \otimes \dots \otimes e_{i_p})
\end{aligned}$$

(3) Finally, the involution axiom follows from the following computation:

$$\begin{aligned}
& T_\pi^*(e_{j_1} \otimes \dots \otimes e_{j_q}) \\
&= \sum_{i_1 \dots i_p} \langle T_\pi^*(e_{j_1} \otimes \dots \otimes e_{j_q}), e_{i_1} \otimes \dots \otimes e_{i_p} \rangle e_{i_1} \otimes \dots \otimes e_{i_p} \\
&= \sum_{i_1 \dots i_p} \delta_\pi \begin{pmatrix} i_1 & \dots & i_p \\ j_1 & \dots & j_q \end{pmatrix} e_{i_1} \otimes \dots \otimes e_{i_p} \\
&= T_{\pi^*}(e_{j_1} \otimes \dots \otimes e_{j_q})
\end{aligned}$$

Summarizing, our correspondence is indeed categorical, as claimed. \square

In relation with the quantum groups, we have the following result, from [15]:

THEOREM 8.31. *Each category of partitions $D = (D(k, l))$ produces a family of compact quantum groups $G = (G_N)$, one for each $N \in \mathbb{N}$, via the following formula:*

$$\text{Hom}(u^{\otimes k}, u^{\otimes l}) = \text{span} \left(T_\pi \Big|_{\pi \in D(k, l)} \right)$$

To be more precise, the spaces on the right form a Tannakian category, and so produce a certain closed subgroup $G_N \subset U_N^+$, via the Tannakian duality correspondence.

PROOF. This follows indeed from Woronowicz's Tannakian duality, in its "soft" form from [56], as explained in Theorem 8.27 above. Indeed, let us set:

$$C(k, l) = \text{span} \left(T_\pi \Big| \pi \in D(k, l) \right)$$

By using the axioms in Definition 8.29, and the categorical properties of the operation $\pi \rightarrow T_\pi$, from Proposition 8.30 above, we deduce that $C = (C(k, l))$ is a Tannakian category. Thus the Tannakian duality applies, and gives the result. \square

All the above might seem a bit complicated, but we will see examples in a moment. Philosophically speaking, the quantum groups appearing as in Theorem 8.31 are the simplest, from the perspective of Tannakian duality, so let us formulate:

DEFINITION 8.32. *A closed subgroup $G \subset U_N^+$ is called easy when we have*

$$\text{Hom}(u^{\otimes k}, u^{\otimes l}) = \text{span} \left(T_\pi \Big| \pi \in D(k, l) \right)$$

for any colored integers k, l , for a certain category of partitions $D \subset P$.

In other words, a compact quantum group is called easy when its Tannakian category appears in the simplest possible way: from partitions. The terminology is quite natural, because Tannakian duality is basically our only serious tool. See [15], [22], [94].

Observe that the category D is not unique, for instance because at $N = 1$ all the categories of partitions produce the same easy quantum group, namely $G = \{1\}$. We will be back to this issue on several occasions, with various results about it.

Before getting into examples, let us formulate, as a first result based on easiness, the following remarkable particularization of Theorem 8.28 above:

THEOREM 8.33. *For an easy quantum group $G \subset U_N^+$, coming from a category of partitions $D = (D(k, l))$, we have the Weingarten integration formula*

$$\int_G u_{i_1 j_1}^{e_1} \dots u_{i_k j_k}^{e_k} = \sum_{\pi, \sigma \in D(k)} \delta_\pi(i) \delta_\sigma(j) W_{kN}(\pi, \sigma)$$

for any $k = e_1 \dots e_k$ and any i, j , where $D(k) = D(\emptyset, k)$, δ are usual Kronecker symbols, and $W_{kN} = G_{kN}^{-1}$, with $G_{kN}(\pi, \sigma) = N^{|\pi \vee \sigma|}$, where $|\cdot|$ is the number of blocks.

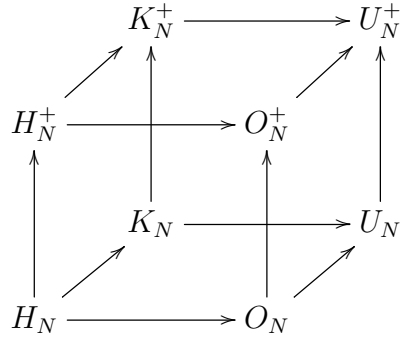
PROOF. With notations from Theorem 8.28, the Kronecker symbols are given by:

$$\begin{aligned} \delta_{\xi_\pi}(i) &= \langle \xi_\pi, e_{i_1} \otimes \dots \otimes e_{i_k} \rangle \\ &= \delta_\pi(i_1, \dots, i_k) \end{aligned}$$

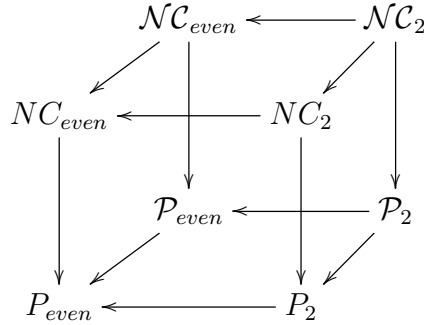
The Gram matrix being as well the correct one, we obtain the result. \square

The above result is very interesting, because everything there is computational, and can be run on a basic laptop. Getting now into examples, we have:

THEOREM 8.34. *The basic quantum unitary and reflection groups, namely*



are all easy, the corresponding categories of partitions being as follows,



exactly as for the main limiting laws in classical and free probability.

PROOF. This is something crucial, the idea being as follows:

(1) The quantum group U_N^+ is defined via the following relations:

$$u^* = u^{-1}$$

$$u^t = \bar{u}^{-1}$$

Thus, the following operators must be in the associated Tannakian category C :

$$T_\pi \quad , \quad \pi = \begin{array}{c} \cap \\ \bullet \bullet \end{array}$$

$$T_\pi \quad , \quad \pi = \begin{array}{c} \cap \\ \bullet \circ \end{array}$$

Thus the associated Tannakian category is $C = span(T_\pi | \pi \in D)$, with:

$$D = \langle \begin{array}{c} \cap \\ \bullet \bullet \end{array} , \begin{array}{c} \cap \\ \bullet \circ \end{array} \rangle = \mathcal{NC}_2$$

Thus, we are led to the conclusion in the statement.

(2) The quantum group $O_N^+ \subset U_N^+$ is defined by imposing the following relations:

$$u_{ij} = \bar{u}_{ij}$$

Thus, the following operators must be in the associated Tannakian category C :

$$T_\pi \quad , \quad \pi = \begin{array}{c} \circ \\ | \\ \bullet \end{array}$$

$$T_\pi \quad , \quad \pi = \begin{array}{c} \bullet \\ | \\ \circ \end{array}$$

Thus the associated Tannakian category is $C = \text{span}(T_\pi | \pi \in D)$, with:

$$D = \langle \mathcal{NC}_2, \begin{array}{c} \circ \\ | \\ \bullet \end{array}, \begin{array}{c} \bullet \\ | \\ \circ \end{array} \rangle = \mathcal{NC}_2$$

Thus, we are led to the conclusion in the statement.

(3) The group $U_N \subset U_N^+$ is defined via the following relations:

$$[u_{ij}, u_{kl}] = 0$$

$$[u_{ij}, \bar{u}_{kl}] = 0$$

Thus, the following operators must be in the associated Tannakian category C :

$$T_\pi \quad , \quad \pi = \begin{array}{c} \circ \\ | \\ \circ \end{array}$$

$$T_\pi \quad , \quad \pi = \begin{array}{c} \bullet \\ | \\ \bullet \end{array}$$

Thus the associated Tannakian category is $C = \text{span}(T_\pi | \pi \in D)$, with:

$$D = \langle \mathcal{NC}_2, \begin{array}{c} \circ \\ | \\ \circ \end{array}, \begin{array}{c} \bullet \\ | \\ \bullet \end{array} \rangle = \mathcal{P}_2$$

Thus, we are led to the conclusion in the statement.

(4) In order to deal now with O_N , we can simply use the following formula:

$$O_N = O_N^+ \cap U_N$$

At the categorical level, this tells us that the associated Tannakian category is given by $C = \text{span}(T_\pi | \pi \in D)$, with:

$$D = \langle \mathcal{NC}_2, \mathcal{P}_2 \rangle = \mathcal{P}_2$$

Thus, we are led to the conclusion in the statement.

(5) Summarizing, we have results for the right face of the cube, involving rotations and quantum rotations. The results for the left face of the cube, involving reflections and quantum reflections, are similar, and for full details here, we refer to [7]. \square

8d. Truncated characters

Getting back now to integration questions and to the Weingarten formula from Theorem 8.33, we are not ready yet for applications, because we still have to understand which assumptions on $N \in \mathbb{N}$ make the vectors ξ_π linearly independent. For this purpose, we will study the Gram matrix of these vectors, which is as follows:

PROPOSITION 8.35. *The Gram matrix of the vectors coming from partitions,*

$$G_{kN}(\pi, \sigma) = \langle \xi_\pi, \xi_\sigma \rangle$$

is given by the following formula,

$$G_{kN}(\pi, \sigma) = N^{|\pi \vee \sigma|}$$

where $|\cdot|$ is the number of blocks.

PROOF. According to the formula of the vectors ξ_π , we have:

$$\begin{aligned} \langle \xi_\pi, \xi_\sigma \rangle &= \sum_{i_1 \dots i_k} \delta_\pi(i_1, \dots, i_k) \delta_\sigma(i_1, \dots, i_k) \\ &= \sum_{i_1 \dots i_k} \delta_{\pi \vee \sigma}(i_1, \dots, i_k) \\ &= N^{|\pi \vee \sigma|} \end{aligned}$$

Thus, we have obtained the formula in the statement. □

We will need several standard facts about the partitions. We have:

DEFINITION 8.36. *The Möbius function of any lattice, and so of P , is given by*

$$\mu(\pi, \sigma) = \begin{cases} 1 & \text{if } \pi = \sigma \\ -\sum_{\pi \leq \tau < \sigma} \mu(\pi, \tau) & \text{if } \pi < \sigma \\ 0 & \text{if } \pi \not\leq \sigma \end{cases}$$

with the construction being performed by recurrence.

The main interest in the Möbius function comes from the Möbius inversion formula, which states that the following happens:

$$f(\sigma) = \sum_{\pi \leq \sigma} g(\pi) \implies g(\sigma) = \sum_{\pi \leq \sigma} \mu(\pi, \sigma) f(\pi)$$

In linear algebra terms, the statement and proof of this formula are as follows:

THEOREM 8.37. *The inverse of the adjacency matrix of P , given by*

$$A_{\pi\sigma} = \begin{cases} 1 & \text{if } \pi \leq \sigma \\ 0 & \text{if } \pi \not\leq \sigma \end{cases}$$

is the Möbius matrix of P , given by $M_{\pi\sigma} = \mu(\pi, \sigma)$.

PROOF. This is well-known, coming for instance from the fact that A is upper triangular. Indeed, when inverting, we are led into the recurrence from Definition 8.36. \square

As an illustration, for $P(2) = \{|\cdot, \square\}$, the formula $M = A^{-1}$ appears as follows:

$$\begin{pmatrix} 1 & -1 \\ 0 & 1 \end{pmatrix} = \begin{pmatrix} 1 & 1 \\ 0 & 1 \end{pmatrix}^{-1}$$

Also, for $P(3) = \{||\cdot, \square, \square, |\cdot, \square, \square\}$ the formula $M = A^{-1}$ reads:

$$\begin{pmatrix} 1 & -1 & -1 & -1 & 2 \\ 0 & 1 & 0 & 0 & -1 \\ 0 & 0 & 1 & 0 & -1 \\ 0 & 0 & 0 & 1 & -1 \\ 0 & 0 & 0 & 0 & 1 \end{pmatrix} = \begin{pmatrix} 1 & 1 & 1 & 1 & 1 \\ 0 & 1 & 0 & 0 & 1 \\ 0 & 0 & 1 & 0 & 1 \\ 0 & 0 & 0 & 1 & 1 \\ 0 & 0 & 0 & 0 & 1 \end{pmatrix}^{-1}$$

Now back to our Gram matrix considerations, we have the following result:

PROPOSITION 8.38. *The Gram matrix is given by $G_{kN} = AL$, where*

$$L(\pi, \sigma) = \begin{cases} N(N-1)\dots(N-|\pi|+1) & \text{if } \sigma \leq \pi \\ 0 & \text{otherwise} \end{cases}$$

and where $A = M^{-1}$ is the adjacency matrix of $P(k)$.

PROOF. We have indeed the following computation:

$$\begin{aligned} N^{|\pi \vee \sigma|} &= \#\{i_1, \dots, i_k \in \{1, \dots, N\} \mid \ker i \geq \pi \vee \sigma\} \\ &= \sum_{\tau \geq \pi \vee \sigma} \#\{i_1, \dots, i_k \in \{1, \dots, N\} \mid \ker i = \tau\} \\ &= \sum_{\tau \geq \pi \vee \sigma} N(N-1)\dots(N-|\tau|+1) \end{aligned}$$

According to Proposition 8.35 and to the definition of A, L , this formula reads:

$$\begin{aligned} (G_{kN})_{\pi\sigma} &= \sum_{\tau \geq \pi} L_{\tau\sigma} \\ &= \sum_{\tau} A_{\pi\tau} L_{\tau\sigma} \\ &= (AL)_{\pi\sigma} \end{aligned}$$

Thus, we obtain in this way the formula in the statement. \square

With the above result in hand, we can now investigate the linear independence properties of the vectors ξ_π . To be more precise, we have the following result:

THEOREM 8.39. *The determinant of the Gram matrix G_{kN} is given by:*

$$\det(G_{kN}) = \prod_{\pi \in P(k)} \frac{N!}{(N - |\pi|)!}$$

In particular, the vectors $\{\xi_\pi | \pi \in P(k)\}$ are linearly independent for $N \geq k$.

PROOF. According to the formula in Proposition 8.38 above, we have:

$$\det(G_{kN}) = \det(A) \det(L)$$

Now if we order $P(k)$ with respect to the number of blocks, then lexicographically, A is upper triangular, and L is lower triangular. Thus $\det(A)$ can be computed by making the product on the diagonal, and we obtain 1. As for $\det(L)$, this can be computed as well by making the product on the diagonal, and we obtain the number in the statement. \square

Now back to the laws of characters, we can formulate:

THEOREM 8.40. *For an easy group $G = (G_N)$, coming from a category of partitions $D = (D(k, l))$, the asymptotic moments of the main character are given by*

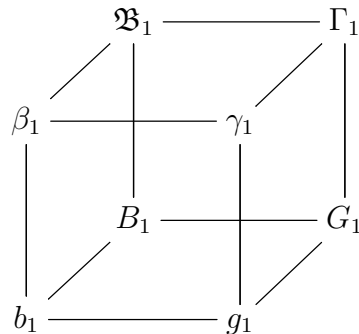
$$\lim_{N \rightarrow \infty} \int_{G_N} \chi^k = \#D(k)$$

where $D(k) = D(\emptyset, k)$, with the limiting sequence on the left consisting of certain integers, and being stationary at least starting from the k -th term.

PROOF. This follows indeed from the general formula for easy groups, by using the linear independence result from Theorem 8.39 above. \square

In practice now, for the basic rotation and reflection groups, we obtain:

THEOREM 8.41. *The character laws for the basic rotation and reflection groups are*



in the $N \rightarrow \infty$ limit.

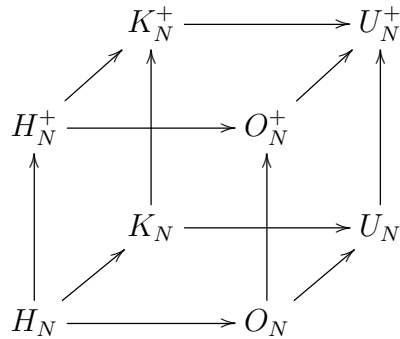
PROOF. This follows indeed from Theorem 8.34 and Theorem 8.40, by using the known moment formulae for the laws in the statement, from Theorem 8.23, at $t = 1$. \square

An interesting feature of the above result is that in the free case, the convergence can be shown to be actually stationary starting from $N = 4$. The “fix” comes by looking at truncated characters, which are constructed as follows:

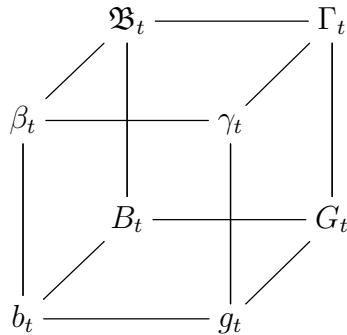
$$\chi_t = \sum_{i=1}^{[tN]} u_{ii}$$

With this convention, we have the following final result on the subject, with the convergence being non-stationary at $t < 1$, in both the classical and free cases:

THEOREM 8.42. *For the basic quantum rotation and reflection groups, namely*



the corresponding asymptotic truncated character laws are as follows,



with these latter laws appearing as well from the basic probabilistic limiting theorems.

PROOF. We already know that the result holds at $t = 1$, and the proof at arbitrary $t > 0$ is once again based on easiness, but this time using the Weingarten formula for the computation of the moments. For the full story here, we refer to the original papers. \square

Finally, let us mention that there are as well similar results for quotient spaces of type G/H . As before, we refer here to the literature on the subject.

8e. Exercises

There has been a lot of exciting theory in this chapter, often with some details missing, and our exercises will be about this. In relation with free probability, we have:

EXERCISE 8.43. *Look up all the details for Voiculescu's R -transform theorem, and write a brief account of what you learned.*

Alternatively, we can not look up anything at all, and try to recover the result from the brief indications given in the above, when talking about it. Up to you.

Regarding now groups and quantum groups, the must-do exercise here is:

EXERCISE 8.44. *Consider the symmetric group S_N , regarded as symmetry group of the N coordinate axes of \mathbb{R}^N , and so as group of orthogonal matrices:*

$$S_N \subset O_N$$

Compute the main character for this group, then the law of this main character, and work out the $N \rightarrow \infty$ asymptotics. Then, study as well the truncated characters.

As a comment here, since the permutation matrices have 0 – 1 entries, the law of the main character is supported by \mathbb{N} . Thus, with a bit of luck, the asymptotic spectral measure can only be the most basic measure in discrete probability.

We have a similar exercise regarding the quantum group S_N^+ , as follows:

EXERCISE 8.45. *Prove that S_N^+ is easy, coming from the category of all noncrossing partitions NC , and compute the asymptotic law of the main character.*

As a bonus exercise, try to do as well the computation for the truncated characters.

Finally, in relation with the Weingarten formula, we have:

EXERCISE 8.46. *Work out the explicit formula of the Weingarten matrix*

$$W_{kN} = G_{kN}^{-1} \quad : \quad G_{kN}(\pi, \sigma) = N^{|\pi \vee \sigma|}$$

for the basic free quantum groups, at small values of $N \in \mathbb{N}$.

In practice, before even starting, you will have to find a good order relation on the partitions, in order to write down the Gram and Weingarten matrices. As for the computations themselves, these are all very instructive, do as many as you can.

Part III

Theory of factors

*And the story tellers say
That the score brave souls inside
For many a lonely day sailed across the milky seas
Never looked back, never feared, never cried*

CHAPTER 9

Functional analysis

9a. Kaplansky density

We have seen so far the definition of the von Neumann algebras $A \subset B(H)$, along with some basic theory and examples. In the remainder of this book we go for the real thing, namely advanced theory, and what can be done with such algebras.

This is actually no easy task, things being quite heavy here. The normal lineup of presentation, roughly based on the historical discovery of the results, but with some retrospective modifications too, due to the fact that some technical fundamental results were discovered, or at least fully proved, with a bit of delay, is as follows:

(1) First comes a heavy result by von Neumann, called “reduction theory” theorem, stating that when writing the center of a von Neumann algebra as $Z(A) = L^\infty(X)$, the whole algebra will decompose as an integral as follows, with the fibers A_x being von Neumann algebras with trivial center, $Z(A_x) = \mathbb{C}$, also called “factors”:

$$A = \int_X A_x dx$$

We have already seen an instance of this result in chapter 5 above, when talking about finite dimensional algebras, which decompose as direct sums of matrix algebras:

$$A = \bigoplus_x M_{n_x}(\mathbb{C})$$

More generally, it is possible to axiomatize a certain class of “type I algebras”, and then show that these algebras appear as direct integrals of matrix algebras:

$$A = \int_X M_{n_x}(\mathbb{C}) dx$$

Beyond type I, however, things become fairly complicated, requiring an excellent mastering of advanced functional analysis techniques. The only ray of light comes from a possible trace $tr : A \rightarrow \mathbb{C}$, whose existence simplifies a bit things, and which actually adds to the final result, in the form of the following supplementary formula:

$$tr = \int_X tr_x dx$$

However, even this latter reduction theory result, known as reduction theory theorem for the “finite von Neumann algebras”, remains something quite complicated.

(2) Assuming the reduction theory understood, the next natural question is the study of the factors. This was first done by Murray and von Neumann, in a series of papers written actually before von Neumann’s reduction theory paper, notably with some beautiful and surprising results regarding the “II₁ factors”, which are by definition the factors which are infinite dimensional, and which have a trace $tr : A \rightarrow \mathbb{C}$.

(3) Among the II₁ factors, of particular interest is the Murray-von Neumann hyperfinite factor R , which is the “smallest” such factor, obtained as an inductive limit of matrix algebras, via a theorem stating that the weak closure of such a limit does not depend on the exact matrix algebras involved, nor on the particular inclusions between them:

$$R = \overline{\bigcup_{n_i} M_{n_i}(\mathbb{C})}^w$$

(4) Regarding the factors which are left, these are either of type I, meaning $B(H)$, or of type II_∞, meaning $B(H)$ tensor II₁, or of type III. The structure and classification of the type III factors has been an open problem for a long time, since Murray and von Neumann themselves, until Connes came with a solution in the 70s. The solution is complicated, but the final answer is quite simple, stating more or less that the type III factors appear in general from type II factors, via crossed product constructions.

(5) In regards now with hyperfiniteness, difficult questions here as well. In type II, Connes completed in the 70s the work of Murray and von Neumann, with a powerful functional analytic characterization of R , adding to the previously known results. In type III the general philosophy is that the hyperfinite factors should basically appear from R , via crossed product constructions, and this is indeed the case, with a lot of work done by Connes in the 70s, and with the whole thing completed by Haagerup in the 80s.

(6) While purely mathematical, all the above work, and especially the work of Murray-von Neumann in type II, then of Connes in type III, and also some fundamental work of Jones in type II, not in direct relation with the classification, eventually gave an indication too on what von Neumann algebras should be good for. To be more precise, the applications of the von Neumann algebras tend to come from factors, of non-trivial type, II or III, and more specifically from the hyperfinite factors of type II or III.

So, this was for the story, and now good luck with learning all this.

Jokes left aside now, the main problem comes from von Neumann’s reduction theory, (1) above. On one hand, this is definitely something mandatory for talking about factors, (2-3-4-5-6) above. On the other hand, reduction theory (1) is something quite heavy, and

once you get to factors, (2-3-4-5-6), as a conclusion to it, the theory is not of that much practical use, and you might regret not having learned other useful things instead.

In view of these difficulties, which are actually commonplace for everyone willing to learn von Neumann algebra theory, or trying to write a book on the subject, here will be our plan for the remainder of this chapter, part, and second half of this book:

- We will discuss (2) in detail, and (3) in some detail too.
- We will also discuss (1) briefly, and (4-5) even more briefly.
- And in chapters 13-16 we will go for (6), in the simplest case, type II.

In order to get started now, let us go back to the general theory of the von Neumann algebras, developed by using basic operator theory, in chapters 5-6 above. As a first objective, we would like to have a better understanding of the precise difference between the arbitrary norm closed $*$ -algebras $A \subset B(H)$, and the weakly closed such algebras, which are the von Neumann algebras, from a functional analytic viewpoint.

Let us start with a standard result, as follows:

PROPOSITION 9.1. *Given an operator algebra $A \subset B(H)$, a linear form $f : A \rightarrow \mathbb{C}$ is weakly continuous precisely when it is of the form*

$$f(T) = \sum_{i=1}^n \langle Tx_i, y_i \rangle$$

for a certain number $n \in \mathbb{N}$, and certain vectors $x_i, y_i \in H$.

PROOF. This is something standard, using the same tools at those already used in chapters 5-6 above, namely basic functional analysis, and amplification tricks:

(1) In one sense, consider families of vectors $x_i, y_i \in H$. The following linear form is then clearly weakly continuous:

$$f(T) = \sum_{i=1}^n \langle Tx_i, y_i \rangle$$

(2) Conversely now, assume that $f : A \rightarrow \mathbb{C}$ is weakly continuous. By continuity we can find vectors $x_1, \dots, x_n \in H$ and a number $\varepsilon > 0$ such that:

$$\sum_{i=1}^n \|Tx_i\|^2 \leq \varepsilon \implies |f(T)| \leq 1$$

It follows from this that we have the following estimate:

$$|f(T)| \leq \sqrt{\frac{\sum_{i=1}^n \|Tx_i\|^2}{\varepsilon}}$$

Consider now the direct sum $H^{\oplus n}$, and inside it, the following vector:

$$x = (x_1, \dots, x_n) \in H^{\oplus n}$$

Consider also the following linear space, written in tensor product notation:

$$K = \overline{(A \otimes 1)x} \subset H^{\oplus n}$$

We can define then a linear form $f' : K \rightarrow \mathbb{C}$ by the following formula, and continuity:

$$f'(Tx_1, \dots, Tx_n) = f(T)$$

We conclude that there exists a vector $y \in K$ such that:

$$f'((T \otimes 1)y) = \langle (T \otimes 1)x, y \rangle$$

But in terms of the original linear form $f : A \rightarrow \mathbb{C}$, this means that we have:

$$f(T) = \sum_{i=1}^n \langle Tx_i, y_i \rangle$$

Thus, we are led to the conclusion in the statement. \square

We have the following result, called Kaplansky density theorem, which is something very useful, and is of independent interest as well:

THEOREM 9.2. *Given an operator algebra $A \subset B(H)$, the following happen:*

- (1) *The unit ball of A is strongly dense in the unit ball of A'' .*
- (2) *The same happens for the self-adjoint parts of the above unit balls.*

PROOF. Here the first assertion is standard, and the second assertion is something more tricky, making use of functional calculus with the following function:

$$f(t) = \frac{2t}{1+t^2}$$

Indeed, by using this function, and then a standard 2×2 matrix trick, we can eventually deduce the first assertion from the second one, and we are done. To be more precise:

(1) Consider the self-adjoint part $A_{sa} \subset A$. By taking real parts of operators, and using the fact that $T \rightarrow T^*$ is weakly continuous, we have then:

$$\overline{A_{sa}}^w \subset (\overline{A}^w)_{sa}$$

Now since the set A_{sa} is convex, and all weak operator topologies coincide on the convex sets, we conclude that we have in fact equality:

$$\overline{A_{sa}}^w = (\overline{A}^w)_{sa}$$

(2) With this result in hand, let us prove now the second assertion of the theorem. Consider an element $T \in \overline{A}^w$, satisfying $T = T^*$ and $\|T\| \leq 1$. Consider as well the following function, from $[-1, 1]$ to itself:

$$f(t) = \frac{2t}{1+t^2}$$

By functional calculus we can find an element $S \in (\overline{A}^w)_{sa}$ such that:

$$f(S) = T$$

In other words, we can find an element $S \in (\overline{A}^w)_{sa}$ such that:

$$T = \frac{2S}{1+S^2}$$

Now given vectors $x_1, \dots, x_n \in H$ and a number $\varepsilon > 0$, let us pick $R \in A_{sa}$, subject to the following two inequalities:

$$\begin{aligned} \|RTx_i - STx_i\| &\leq \varepsilon \\ \left\| \frac{R}{1+S^2}x_i - \frac{S}{1+S^2}x_i \right\| &\leq \varepsilon \end{aligned}$$

Finally, consider the following element, which has norm ≤ 1 :

$$L = \frac{2R}{1+R^2}$$

We have then the following computation, using the above formulae:

$$\begin{aligned} L - T &= \frac{2R}{1+R^2} - \frac{2S}{1+S^2} \\ &= 2 \left(\frac{1}{1+R^2} (R(1+S^2) - (1+S^2)R) \frac{1}{1+S^2} \right) \\ &= 2 \left(\frac{1}{1+R^2} (R-S) \frac{1}{1+S^2} + \frac{R}{1+R^2} (S-R) \frac{S}{1+S^2} \right) \\ &= \frac{2}{1+R^2} (R-S) \frac{1}{1+S^2} + \frac{L}{2} (S-R)T \end{aligned}$$

Thus, we have the following estimate, for any $i \in \{1, \dots, n\}$:

$$\|(L - T)x_i\| \leq \varepsilon$$

But this gives the density assertion, (2) in the statement.

(3) Let us prove now the first assertion of the theorem. Given an arbitrary element $T \in \overline{A}^w$, satisfying $\|T\| \leq 1$, let us look at the following element:

$$T' = \begin{pmatrix} 0 & T \\ T^* & 0 \end{pmatrix} \in M_2(\overline{A}^w)$$

This element is then self-adjoint, and we can use what we proved in the above, and we are led to the conclusion in the statement. \square

We can go back now to our original question, and we have:

THEOREM 9.3. *An operator algebra $A \subset B(H)$ is a von Neumann algebra precisely when its unit ball is weakly compact.*

PROOF. This is now something clear, coming from the density results established in Theorem 9.2 above. To be more precise:

(1) In one sense, assuming that $A \subset B(H)$ is a von Neumann algebra, this algebra is weakly closed. But since the unit ball of $B(H)$ is weakly compact, we are led to the conclusion that the unit ball of A is weakly compact too.

(2) Conversely, assume that an operator algebra $A \subset B(H)$ is such that its unit ball is weakly compact. In particular, the unit ball of A is weakly closed. Now if T satisfying $\|T\| \leq 1$ belongs to the weak closure of A , by Kaplansky density we conclude that we have $T \in A$. Thus our algebra A must be a von Neumann algebra, as claimed. \square

There are several other abstract characterizations of the von Neumann algebras, inside the class of C^* -algebras, and we will be back to this, on several occasions, and notably at the end of the present chapter, with such a characterization involving the predual.

9b. Projections, factors

Among the von Neumann algebras $A \subset B(H)$, of particular interest are the “free” ones, having trivial center, $Z(A) = \mathbb{C}$. These algebras are called factors:

DEFINITION 9.4. *A factor is a von Neumann algebra $A \subset B(H)$ whose center*

$$Z(A) = A \cap A'$$

which is a commutative von Neumann algebra, reduces to the scalars, $Z(A) = \mathbb{C}$.

Here the fact that the center is indeed a von Neumann algebra follows from the bicommutant theorem, which shows that the commutant of any $*$ -algebra is a von Neumann algebra. Thus, the intersection $Z(A) = A \cap A'$ is a von Neumann algebra as well.

In what follows we will be mainly interested in these factors, with our motivation coming from the fact that the condition $Z(A) = \mathbb{C}$ defining them is somehow opposite to the condition $Z(A) = A$ defining the commutative von Neumann algebras. Thus, the factors are the von Neumann algebras which are “free”, with this meaning being as far as possible from the commutative ones. Equivalently, in terms of the quantum space writing $A = L^\infty(X)$, the spaces X coming from factors are as “free” as one can hope for.

In order to investigate the factors, the idea, coming from the analysis of the finite dimensional algebras from chapter 5 above, will be that of looking at the projections.

We will explain now this material, following the classical work of Murray and von Neumann. Let us start with some generalities. In analogy with what happens in finite dimensions, we have the following notions, over an arbitrary Hilbert space H :

DEFINITION 9.5. *Associated to any two projections $P, Q \in B(H)$ are:*

- (1) *The projection $P \wedge Q$, projecting on the common range.*
- (2) *The projection $P \vee Q$, projecting on the span of the ranges.*

Abstractly speaking, these two operations can be thought of as being inf and sup type operations, and all the known algebraic formulae for inf and sup hold in this setting. For the moment we will not need all this, and we will be back to it later.

Let us record however the following basic formula, which is useful:

PROPOSITION 9.6. *We have the following formula,*

$$P + Q = P \wedge Q + P \vee Q$$

valid for any two projections $P, Q \in B(H)$.

PROOF. This is clear from definitions, because when computing $P + Q$ we obtain the projection $P \vee Q$ on the span on the ranges, modulo the fact that the vectors in the common range are obtained twice, which amounts in saying that we must add $P \wedge Q$. \square

With the above notions in hand, we have the following result:

THEOREM 9.7. *Consider two projections $P, Q \in B(H)$.*

- (1) *In finite dimensions, over $H = \mathbb{C}^N$, we have, in norm:*

$$(PQ)^n \rightarrow P \wedge Q$$

- (2) *In infinite dimensions, we have the following convergence, for any $x \in H$,*

$$(PQ)^n x \rightarrow (P \wedge Q)x$$

but the operators $(PQ)^n$ do not necessarily converge in norm.

PROOF. We have several assertions here, the proof being as follows:

(1) Assume that we are in the case $P, Q \in M_N(\mathbb{C})$. By subtracting $P \wedge Q$ from both P, Q , we can assume $P \wedge Q = 0$, and we must prove that we have:

$$P \wedge Q = 0 \implies (PQ)^n \rightarrow 0$$

Our claim is that we have $\|PQ\| < 1$. Indeed, we know that we have:

$$\|PQ\| \leq \|P\| \cdot \|Q\| = 1$$

Assuming now by contradiction that we have $\|PQ\| = 1$, since we are in finite dimensions, we must have, for a certain norm one vector, $\|x\| = 1$:

$$\|PQx\| = 1$$

Thus, we must have equalities in the following estimate:

$$\|PQx\| \leq \|Qx\| \leq \|x\|$$

But the second equality tells us that we must have $x \in \text{Im}(Q)$, and with this in hand, the first equality tells us that we must have $x \in \text{Im}(P)$. But this contradicts $P \wedge Q = 0$, so we have proved our claim, and the convergence $(PQ)^n \rightarrow 0$ follows.

(2) In infinite dimensions now, as before by subtracting $P \wedge Q$ from both P, Q , we can assume $P \wedge Q = 0$, and we must prove that we have, for any $x \in H$:

$$P \wedge Q = 0 \implies (PQ)^n x \rightarrow 0$$

For this purpose, consider the following operator:

$$R = PQP$$

This operator is positive, because we have $R = (PQ)(PQ)^*$, and we have:

$$\|R\| \leq \|P\| \cdot \|Q\| \cdot \|P\| = 1$$

Our claim, which will finish the proof, is that for any $x \in H$ we have:

$$R^n x \rightarrow 0$$

In order to prove this claim, let us diagonalize R , by using the spectral theorem for self-adjoint operators, from chapter 3 above. If all the eigenvalues are < 1 then we are done. If not, this means that we can find a nonzero vector $x \in H$ such that:

$$\|Rx\| = \|x\|$$

But this condition means that we must have equalities in the following estimate:

$$\|PQP Px\| \leq \|QP Px\| \leq \|P Px\| \leq \|x\|$$

The point now is that this is impossible, due to our assumption $P \wedge Q = 0$. Indeed, the last equality tells us that we must have $x \in \text{Im}(P)$, and with this in hand, the middle equality tells us that we must have $x \in \text{Im}(Q)$. But this contradicts $P \wedge Q = 0$, so we have proved our claim, and the convergence $(PQ)^n x \rightarrow 0$ follows.

(3) Finally, for a counterexample to $(PQ)^n \rightarrow 0$, in infinite dimensions, we can take $H = l^2(\mathbb{N})$, and then find projections P, Q such that $(PQ)^n e_k \rightarrow 0$ for any k , but with the convergence arbitrarily slowing down with $k \rightarrow \infty$. Thus, $(PQ)^n \not\rightarrow 0$. \square

As a consequence of the above, in connection with von Neumann algebras, we have:

THEOREM 9.8. *Given two projections $P, Q \in B(H)$, the projections*

$$P \wedge Q \in B(H)$$

$$P \vee Q \in B(H)$$

belong to the von Neumann algebra generated by P, Q .

PROOF. This is something that follows from the above, as follows:

(1) In what regards $P \wedge Q$, this is something that follows from Theorem 9.7, with the comment that there are some other proofs as well for this.

(2) As for $P \vee Q$, here the result follows from the result for $P \wedge Q$, discussed above, and from the formula $P + Q = P \wedge Q + P \vee Q$, from Proposition 9.6. \square

The idea now will be that of studying the von Neumann algebras $A \subset B(H)$ by using their projections, $p \in A$. Let us start with the following result:

THEOREM 9.9. *Any von Neumann algebra*

$$A \subset B(H)$$

is generated by its projections.

PROOF. This is something that we know, from chapter 5, which comes from the measurable functional calculus, which can cut any normal operator into projections. \square

There are many other things that can be said about projections, in the general setting. In what follows we will just discuss the most important and useful such results. A first such result, providing us with some geometric intuition on projections, is as follows:

THEOREM 9.10. *Given a von Neumann algebra $A \subset B(H)$, and a projection $p \in A$, we have the following equalities, between von Neumann algebras on pH :*

- (1) $pAp = (A'p)'$.
- (2) $(pAp)' = A'p$.

PROOF. This is not exactly obvious, but can be proved as follows:

(1) As a first observation, the algebras pAp and $A'p$ commute on pH . Thus, we must prove that we have the following implication:

$$x \in (A'p)' \implies x \in pAp$$

For this purpose, consider the element $y = xp$. Then for any $z \in A'$ we have:

$$\begin{aligned} zy &= zxp \\ &= zpxp \\ &= xpzp \\ &= xpz \\ &= yz \end{aligned}$$

Thus we obtain $y \in A$, and so we have, as desired:

$$x = pyp \in pAp$$

(2) As before, one of the inclusions being clear, we must prove that we have:

$$x \in (pAp)' \implies x \in A'p$$

By using the standard fact that any bounded operator appears as a linear combination of 4 unitaries, that we know from chapter 4 above, it is enough to prove this for a unitary element, $x = u$. So, assume that we have a unitary as follows:

$$u \in (pAp)'$$

In order to prove our claim, consider the following vector space:

$$K = \overline{ApH}$$

This space being invariant under both A, A' , the projection $q = Proj(K)$ onto it belongs to the center of our von Neumann algebra:

$$q \in Z(A)$$

Our claim is that we can extend the above unitary $u \in (pAp)'$ to the space $K = \overline{ApH}$ via the following formula, valid for any elements $x_i \in A$, and any vectors $\xi_i \in pH$:

$$v \left(\sum_i x_i \xi_i \right) = \sum_i x_i u \xi_i$$

Indeed, we have the following computation:

$$\begin{aligned} \left\| v \left(\sum_i x_i \xi_i \right) \right\|^2 &= \sum_{ij} \langle x_i u \xi_i, x_j u \xi_j \rangle \\ &= \sum_{ij} \langle x_j^* x_i u \xi_i, u \xi_j \rangle \\ &= \sum_{ij} \langle p x_j^* x_i p u \xi_i, u \xi_j \rangle \\ &= \sum_{ij} \langle u p x_j^* x_i p \xi_i, u \xi_j \rangle \\ &= \sum_{ij} \langle p x_j^* x_i p \xi_i, \xi_j \rangle \\ &= \sum_{ij} \langle x_j^* x_i \xi_i, \xi_j \rangle \\ &= \sum_{ij} \langle x_i \xi_i, x_j \xi_j \rangle \\ &= \left\| \sum_i x_i \xi_i \right\|^2 \end{aligned}$$

Thus v is well-defined by the above formula, and is an isometry of K . Now observe that this element v commutes with A on the space ApH , hence on K . Thus $vq \in A'$, and so $u = vqp$, which proves that we have $u \in A'p$, as desired. \square

As a second result now, once again in the general setting, we have:

PROPOSITION 9.11. *Given a von Neumann algebra $A \subset B(H)$, the formula*

$$p \simeq q \iff \exists u, \begin{cases} uu^* = p \\ u^*u = q \end{cases}$$

defines an equivalence relation for the projections $p \in A$.

PROOF. This is something elementary, which follows from definitions, with the transitivity coming by composing the corresponding partial isometries. \square

As a third and final result, once again in the general setting, which once again provides us with some intuition, but this time of somewhat abstract type, we have:

THEOREM 9.12. *Given a von Neumann algebra $A \subset B(H)$, we have a partial order on the projections $p \in A$, constructed as follows, with u being a partial isometry,*

$$p \preceq q \iff \exists u, \begin{cases} uu^* = p \\ u^*u \leq q \end{cases}$$

which is related to the equivalence relation \simeq constructed above by:

$$p \simeq q \iff p \preceq q, q \preceq p$$

Thus, \preceq is a partial order on the equivalence classes of projections $p \in A$.

PROOF. We have several assertions here, the idea being as follows:

(1) The fact that we have indeed a partial order is clear, with the transitivity coming, as before, by composing the corresponding partial isometries.

(2) Regarding now the relation with \simeq , via the equivalence in the statement, the implication \implies is clear. Thus, we are left with proving \impliedby , which reads:

$$p \preceq q, q \preceq p \implies p \simeq q$$

Our assumption is that we have partial isometries u, v such that:

$$uu^* = p \quad , \quad u^*u \leq q$$

$$v^*v \leq p \quad , \quad vv^* = q$$

We can construct two sequences of decreasing projections, as follows:

$$p_0 = p \quad , \quad p_{n+1} = v^*q_nv$$

$$q_0 = q \quad , \quad q_{n+1} = u^*p_nu$$

Consider now the limits of these two sequences of projections:

$$p_\infty = \bigwedge_i p_i \quad , \quad q_\infty = \bigwedge_i q_i$$

In terms of all these projections that we constructed, we have the following decomposition formulae for the original projections p, q :

$$p = (p - p_1) + (p_1 - p_2) + \dots + p_\infty$$

$$q = (q - q_1) + (q_1 - q_2) + \dots + q_\infty$$

Now observe that the summands are equivalent, with this being clear from the definition of p_n, q_n at the finite indices $n < \infty$, and with $p_\infty \simeq q_\infty$ coming from:

$$v^* q_\infty v = p_\infty \quad , \quad q_\infty v v^* q_\infty = q_\infty$$

Thus we obtain that we have $p \simeq q$, as desired, by summing.

(3) Finally, the fact that the order \preceq factorizes indeed to the equivalence classes under \simeq follows from the equivalence established in (2). \square

Summarizing, in view of Theorem 9.9, and of Theorem 9.12, we can think of a von Neumann algebra $A \subset B(H)$ as being a kind of object belonging to “mathematical logic”, consisting of equivalence classes of projections $p \in A$, ordered via the relation \preceq , and producing A itself via transport by partial isometries, and then linear combinations, and weak limits. This was von Neumann’s original vision, still largely used nowadays.

In what concerns us, we will rather stick to our $A = L^\infty(X)$ viewpoint, with X being a quantum measured space, and the most often being a “quantum manifold”. This is more of a “smooth” philosophy, and in order to keep it intact, and powerful, we will have to take sometimes distances with the von Neumann philosophy, especially in what concerns the terminology. In short, we will be definitely users of the von Neumann projection technology, which is extremely powerful, and is quite often the only available tool, but keeping in mind however that we are dealing with smooth objects X , and choosing the terminology and notations accordingly, inspired from smooth geometry.

Getting back now to work, our next purpose will be that of understanding what happens to the above, in the particular case of the factors. We will need:

PROPOSITION 9.13. *Given two projections $p, q \neq 0$ in a factor A , we have*

$$puq \neq 0$$

for a certain unitary $u \in A$.

PROOF. Assume by contradiction $puq = 0$, for any unitary $u \in A$. This gives:

$$u^*puq = 0$$

By using this for all the unitaries $u \in A$, we obtain:

$$\left(\bigvee_{u \in U_A} u^*pu \right) q = 0$$

On the other hand, from $p \neq 0$ we obtain, by factoriality of A :

$$\bigvee_{u \in U_A} u^*pu = 1$$

Thus, our previous formula is in contradiction with $q \neq 0$, as desired. \square

Getteing back now to the order constructed in Theorem 9.12, and to the whole projection philosophy, in the case of factors things become very simple, as follows:

THEOREM 9.14. *Given two projections $p, q \in A$ in a factor, we have*

$$p \preceq q \quad \text{or} \quad q \preceq p$$

and so \preceq is a total order on the equivalence classes of projections $p \in A$.

PROOF. This basically follows from Proposition 9.13, and from the Zorn lemma, by using standard functional analysis arguments. To be more precise:

(1) Consider indeed the following set of partial isometries:

$$S = \left\{ u \mid uu^* \leq p, u^*u \leq q \right\}$$

We can order this set S by saying that $u \leq v$ when $u^*u \leq v^*v$, and when $u = v$ on the initial domain u^*uH of u . With this convention made, the Zorn lemma applies, and provides us with a maximal element $u \in S$.

(2) In the case where this maximal element $u \in S$ satisfies $uu^* = p$ or $u^*u = q$, we are led to one of the conditions $p \preceq q$ or $q \preceq p$ in the statement, and we are done.

(3) So, assume that we are in the case left, $uu^* \neq p$ and $u^*u \neq q$. By Proposition 9.13 we obtain a unitary $v \neq 0$ satisfying the following conditions:

$$vv^* \leq p - uu^*$$

$$v^*v \leq q - u^*u$$

But these conditions show that the element $u + v \in S$ is strictly bigger than $u \in S$, which is a contradiction, and we are done. \square

We will be back to all this in chapter 10 below, in the tracial setting, where a number of simplifications appear, and where it is possible to go way beyond the above, with a number of quite unexpected results, due to Murray and von Neumann.

9c. States, isomorphism

One question that we met on several occasions, and that we would like to clarify now, is the relation between abstract isomorphism and spatial isomorphism. To be more precise, we would like to understand when two von Neumann algebras $A \subset B(H)$ and $B \subset B(K)$ are isomorphic, in an algebraic and topological sense, but without reference to H, K . Once this understood, we will be able to talk about the von Neumann algebras A as being abstract objects, a bit as were the C^* -algebras, discussed in chapter 7.

In order to discuss this, let us start with a technical result, as follows:

PROPOSITION 9.15. *Given a von Neumann algebra $A \subset B(H)$, and a positive linear form $f : A \rightarrow \mathbb{C}$, the following are equivalent:*

(1) *f is normal, in the sense that we have the formula*

$$f\left(\sup_i x_i\right) = \sup_i f(x_i)$$

for any increasing sequence of positive elements $x_i \in A$.

(2) *f is completely additive, in the sense that we have*

$$f\left(\bigvee_i p_i\right) = \sum_i f(p_i)$$

for any family of pairwise orthogonal projections $\{p_i\}$.

(3) *f is weakly continuous.*

PROOF. This is something very standard, as follows:

(1) \implies (2) Given a family of pairwise orthogonal projections $\{p_i\}$, we can consider the following increasing sequence of positive elements:

$$x_n = \sum_{i=1}^n p_i$$

By using now the formula in (1) for these elements we obtain, as desired:

$$\begin{aligned} f\left(\bigvee_i p_i\right) &= f\left(\sup_n x_n\right) \\ &= \sup_n f(x_n) \\ &= \sup_n \sum_{i=1}^n f(p_i) \\ &= \sum_i f(p_i) \end{aligned}$$

(2) \implies (3) This is something more technical, that we will prove in several steps. Let us fix a projection $q \in A$, and consider a vector $\xi \in \text{Im}(q)$ such that:

$$\langle q\xi, \xi \rangle > 1$$

Our claim is that there exists a projection $p \leq q$ such that, for any $x \in A$:

$$f(xpx) \leq \langle xpx\xi, \xi \rangle$$

Indeed, let us pick, by using the Zorn lemma, a maximal family of pairwise orthogonal projections $\{p_i\} \subset A$ such that, for any i , we have:

$$f(p_i) \geq \langle p_i\xi, \xi \rangle$$

By using our complete additivity assumption, we have then:

$$\begin{aligned} f\left(\bigvee_i p_i\right) &= \sum_i f(p_i) \\ &\geq \sum_i \langle p_i\xi, \xi \rangle \\ &= \left\langle \left(\bigvee_i p_i\right)\xi, \xi \right\rangle \end{aligned}$$

Now consider the following projection, which is nonzero:

$$p = q - \bigvee_i p_i$$

By maximality of the family $\{p_i\}$, for any nonzero projection $r \leq p$, we have:

$$f(r) < \langle r\xi, \xi \rangle$$

We therefore obtain the following estimate, valid for any $x \in A_+$, as desired:

$$f(xpx) \leq \langle xpx\xi, \xi \rangle$$

Now by Cauchy-Schwarz we obtain that for any $x \in A$, $\|x\| \leq 1$, we have:

$$\begin{aligned} |f(xp)|^2 &\leq f(px^*xp)f(1) \\ &\leq \langle px^*xp\xi, \xi \rangle \\ &= \|\xi\|^2 \end{aligned}$$

Thus the following linear form is strongly continuous on the unit ball of A :

$$x \rightarrow f(px)$$

In order to finish now, once again by using the Zorn lemma, let us pick a maximal family of pairwise orthogonal projections $\{p_i\} \subset A$ such that $x \rightarrow f(p_i x)$ is strongly continuous on the unit ball of A , for any i . By maximality we have then:

$$\begin{aligned} \sum_i f(p_i) &= f\left(\sum_i p_i\right) \\ &= f(1) \\ &= 1 \end{aligned}$$

Now given $\varepsilon > 0$, let us choose a finite subset of our index set, $F \subset I$, such that for all the finite subsets $F \subset J \subset I$, we have an inequality as follows:

$$1 - f\left(\sum_{j \in J} p_j\right) \leq \varepsilon$$

By Cauchy-Schwarz we have then, for any $x \in A$, $\|x\| = 1$, the following estimate:

$$\begin{aligned} \left|f\left(x\left(1 - \sum_{j \in J} p_j\right)\right)\right|^2 &\leq f\left(1 - \sum_{j \in J} p_j\right) f(xx^*) \\ &\leq \varepsilon \end{aligned}$$

We conclude from this that we have the following estimate:

$$\left\|f - f\left(\cdot\left(1 - \sum_{j \in J} p_j\right)\right)\right\| \leq \sqrt{\varepsilon}$$

Thus we obtain $f \in A_*$, as desired.

(3) \implies (1) This is something trivial, coming from definitions. \square

We will need as well the following result:

PROPOSITION 9.16. *Given a von Neumann algebra $A \subset B(H)$, a positive linear form $f : A \rightarrow \mathbb{C}$ is completely additive, in the sense that*

$$f\left(\bigvee_i p_i\right) = \sum_i f(p_i)$$

for any pairwise orthogonal projections $\{p_i\}$, precisely when it is of the form

$$f(T) = \langle Tx, x \rangle$$

when suitably extended to the representation of A on the Hilbert space $H \otimes l^2(\mathbb{N})$.

PROOF. This is something quite standard, which can be proved by using our usual tools, namely basic functional analysis, and amplification tricks. \square

The linear forms $f : A \rightarrow \mathbb{C}$ appearing as above, namely $f(T) = \langle Tx, x \rangle$ with $x \in H$, taken with the normalization $\|x\| = 1$, are called vector states on A .

We can go back now to our original question, and we have:

THEOREM 9.17. *Given two von Neumann algebras $A \subset B(H)$ and $B \subset B(K)$, acting on possibly different Hilbert spaces H, K , any algebraic isomorphism*

$$\Phi : A \simeq B$$

is spatial up to amplification, in the sense that we have a formula as follows,

$$\Phi(T) \otimes 1 = U(T \otimes 1)U^*$$

for a certain Hilbert space L , and a certain unitary $U : H \otimes L \rightarrow K \otimes L$.

PROOF. This is something standard, coming from Proposition 9.16, as follows:

(1) As a first observation, assuming that a positive unital linear form $f : A \rightarrow \mathbb{C}$ is a vector state, given by a certain vector $x \in H$, then by Proposition 9.16 the linear form $f\Phi^{-1}$ is also a vector state, say given by a vector $y \in K$.

(2) We conclude from this that we have a unitary as follows, intertwining the corresponding actions of the von Neumann algebras A and B :

$$U_x : \overline{Ax} \rightarrow \overline{By}$$

Now by making the above vector $x \in H$ vary, and performing a direct sum, we obtain with $L = l^2(\mathbb{N})$ an isometry as in the statement, namely:

$$U : H \otimes L \rightarrow K \otimes L$$

Our construction shows that U intertwines indeed the actions of the von Neumann algebras A and B , and what is left to do is to study the unitarity of U .

(3) We will prove now that, up to a suitable replacement, the above operator U can be taken to be unitary, still intertwining the actions of the von Neumann algebras A and B . For this purpose, consider the action of von Neumann algebra A on the direct sum Hilbert space $(H \otimes L) \oplus (K \otimes L)$ given by the following matrices:

$$x' = \begin{pmatrix} x \otimes 1 & 0 \\ 0 & \Phi(x) \otimes 1 \end{pmatrix}$$

Since U intertwines the actions of the von Neumann algebras A and B , in terms of 2×2 matrices, we are led to the following conclusion:

$$\begin{pmatrix} 0 & 0 \\ U & 0 \end{pmatrix} \in A'$$

Thus, the following happens inside the von Neumann algebra A' :

$$\begin{pmatrix} 1 & 0 \\ 0 & 0 \end{pmatrix} \preceq \begin{pmatrix} 0 & 0 \\ 0 & 1 \end{pmatrix}$$

On the other hand, the same reasoning applied to the isomorphism Φ^{-1} shows that we have as well, once again inside the von Neumann algebra A' :

$$\begin{pmatrix} 0 & 0 \\ 0 & 1 \end{pmatrix} \preceq \begin{pmatrix} 1 & 0 \\ 0 & 0 \end{pmatrix}$$

(4) We are now in position to finish. By combining the above two conclusions, we obtain an equivalence of projections inside A' , as follows:

$$\begin{pmatrix} 1 & 0 \\ 0 & 0 \end{pmatrix} \simeq \begin{pmatrix} 0 & 0 \\ 0 & 1 \end{pmatrix}$$

Now pick a partial isometry implementing this equivalence. This partial isometry must be of the following form, with U' being now a unitary:

$$V = \begin{pmatrix} 0 & 0 \\ U' & 0 \end{pmatrix}$$

Thus, we have now a certain unitary $U' : H \otimes L \rightarrow K \otimes L$, which intertwines the actions of A and B , which is the unitary we were looking for. \square

The above result is something quite fundamental, allowing us to talk about von Neumann algebras A as abstract objects, without reference to the exact Hilbert space H where the elements $a \in A$ live as operators $a \in B(H)$, and with this being of course possible modulo some functional analysis knowledge. We will heavily use this point of view in chapter 10 below, and then in chapters 13-16 below, when talking about II_1 factors.

9d. Predual theory

Going ahead now with more abstract functional analysis, that we will be using in what follows, on several occasions, let us formulate:

DEFINITION 9.18. *Given a von Neumann algebra $A \subset B(H)$, we set*

$$A_* = \left\{ f : A \rightarrow \mathbb{C}, \text{ weakly continuous} \right\}$$

regarded as a linear subspace, $A_ \subset A^*$, of the usual dual, given by:*

$$A^* = \left\{ f : A \rightarrow \mathbb{C}, \text{ norm continuous} \right\}$$

Our first goal will be that of proving that we have the following duality formula, between the linear space A_* constructed above, and the algebra A itself:

$$A = (A_*)^*$$

In order to do so, let us first discuss the case of the full operator algebra $A = B(H)$ itself. This is actually the key case, with the extension to the arbitrary von Neumann algebras $A \subset B(H)$ being something coming afterwards, quite straightforward.

We will need some standard operator theory, developed in chapter 4 above. First, we have the following result, regarding the trace class operators, established there:

THEOREM 9.19. *The space of trace class operators, which appears as an intermediate space between the finite rank operators and the compact operators,*

$$B_0(H) \subset B_1(H) \subset B_\infty(H)$$

*is a two-sided *-ideal of $B_\infty(H)$. The following is a Banach space norm on $B_1(H)$,*

$$\|T\|_1 = \text{Tr}|T|$$

satisfying $\|T\| \leq \|T\|_1$, and for $T \in B_1(H)$ and $S \in B(H)$ we have:

$$\|ST\|_1 \leq \|S\| \cdot \|T\|_1$$

Also, the subspace $B_0(H)$ is dense inside $B_1(H)$, with respect to this norm.

PROOF. This is something standard, explained in chapter 4 above. □

We will need as well the following result, regarding this time the Hilbert-Schmidt operators, which is also from chapter 4 above:

THEOREM 9.20. *The space of Hilbert-Schmidt operators, which appears as an intermediate space between the trace class operators and the compact operators,*

$$B_0(H) \subset B_1(H) \subset B_2(H) \subset B_\infty(H)$$

*is a two-sided *-ideal of $B_\infty(H)$. In terms of the singular values (λ_n) , the Hilbert-Schmidt operators are characterized by the following formula:*

$$\sum_n \lambda_n^2 < \infty$$

Also, the following formula, taking as input two Hilbert-Schmidt operators,

$$\langle S, T \rangle = \text{Tr}(ST^*)$$

defines a scalar product of $B_2(H)$, making it a Hilbert space.

PROOF. As before, this is something standard, explained in chapter 4 above. □

With these ingredients in hand, let us go ahead with the study of the space $B(H)_*$. We will need the following technical result, regarding the Hilbert-Schmidt operators:

PROPOSITION 9.21. *We have the following formula,*

$$\text{Tr}(ST) = \text{Tr}(TS)$$

valied for any Hilbert-Schmidt operators $S, T \in B_2(H)$.

PROOF. We can prove this in two steps, as follows:

(1) Assume first that $|S|$ is trace class. Consider the polar decomposition $S = U|S|$, and choose an orthonormal basis $\{x_i\}$ for the image of U , suitably extended to an orthonormal basis of H . We have then the following computation, as desired:

$$\begin{aligned} \operatorname{Tr}(ST) &= \sum_i \langle U|S|Tx_i, x_i \rangle \\ &= \sum_i \langle |S|TUU^*x_i, U^*x_i \rangle \\ &= \operatorname{Tr}(|S|TU) \\ &= \operatorname{Tr}(TU|S|) \\ &= \operatorname{Tr}(TS) \end{aligned}$$

(2) Assume now that we are in the general case, where S is only assumed to be Hilbert-Schmidt. For any finite rank operator S' we have then:

$$\begin{aligned} |\operatorname{Tr}(ST) - \operatorname{Tr}(TS)| &= |\operatorname{Tr}((S - S')T) - \operatorname{Tr}(T(S - S'))| \\ &\leq 2\|S - S'\|_2 \cdot \|T\|_2 \end{aligned}$$

Thus by choosing S' with $\|S - S'\|_2 \rightarrow 0$, we obtain the result. \square

With the above technical result in hand, and getting back now to von Neumann algebras, and to our predual questions, we have the following result:

THEOREM 9.22. *The linear space $B(H)_* \subset B(H)^*$ consisting of the linear forms $f : B(H) \rightarrow \mathbb{C}$ which are weakly continuous is given by*

$$B(H)_* = \left\{ T \rightarrow \operatorname{Tr}(ST) \mid S \in B_1(H) \right\}$$

and we have the following duality formula

$$B(H) = (B(H)_*)^*$$

as a duality in the usual Banach space sense.

PROOF. There are several things to be proved, the idea being as follows:

(1) First of all, any linear form of type $T \rightarrow \operatorname{Tr}(ST)$, with S being trace class, is weakly continuous. Thus, if we denote by $B(H)_\circ$ the subspace of $B(H)$ in the statement, consisting of such linear forms, we have an inclusion as follows:

$$B(H)_\circ \subset B(H)_*$$

(2) In order to prove the reverse inclusion, consider an arbitrary weakly continuous linear form $f \in B(H)_*$. We can then find vectors (x_i) and (y_i) such that:

$$f(T) = \sum_i \langle Tx_i, y_i \rangle$$

Let us consider now the following operators, going by definition from the Hilbert space $l^2(\mathbb{N})$ to our Hilbert space H , and which are both Hilbert-Schmidt:

$$Q : e_i \rightarrow x_i \quad , \quad R : e_i \rightarrow y_i$$

In terms of these operators, our linear form can be written as follows:

$$f(T) = \text{Tr}(R^*TQ)$$

On the other hand, by using Proposition 9.21 we obtain:

$$\text{Tr}(R^*TQ) = \text{Tr}(TQR^*)$$

Thus, with $S = QR^*$, which is trace class, we have the following formula:

$$f(T) = \text{Tr}(TS)$$

Thus, we have proved that we have an inclusion as follows:

$$B(H)_* \subset B(H)_\circ$$

(3) Summing up, from (1) and (2) we obtain that we have an equality as follows, which proves the first assertion in the statement:

$$B(H)_* = B(H)_\circ$$

(4) It remains to prove that $B(H)$ is indeed the dual of $B(H)_*$. For this purpose, we use the above identification, which ultimately identifies $B(H)_*$ with the space of trace class operators $B_1(H)$. So, assume that we have a linear form, as follows:

$$f : B_1(H) \rightarrow \mathbb{C}$$

It is then routine to show that f must come from evaluation on a certain operator $T \in B(H)$, and this leads to the conclusion that $B(H)$ is indeed the dual of $B(H)_*$. \square

More generally now, for the arbitrary von Neumann algebras $A \subset B(H)$, we have:

THEOREM 9.23. *Given a von Neumann algebra $A \subset B(H)$, if we set*

$$A_* = \left\{ f : A \rightarrow \mathbb{C}, \text{ weakly continuous} \right\}$$

regarded as a linear subspace, $A_ \subset A^*$, of the usual dual, given by:*

$$A^* = \left\{ f : A \rightarrow \mathbb{C}, \text{ norm continuous} \right\}$$

then we have the duality formula $A = (A_)^*$, in the usual Banach space sense.*

PROOF. This can be proved in several steps, as follows:

(1) First of all, we know from the above that the result holds for the von Neumann algebra $A = B(H)$ itself, in the sense that we have:

$$B(H) = (B(H)_*)^*$$

(2) The point now is that for any von Neumann subalgebra $A \subset B(H)$, or more generally for any weakly closed linear subspace $A \subset B(H)$, we have an equality as follows, coming as a consequence of the Hahn-Banach theorem:

$$A = A^{\perp\perp}$$

(3) Thus, modulo some standard algebra, and some standard identifications for quotient spaces and their duals, we are led to the conclusion in the statement. \square

In fact, we have the following result, due to Sakai:

THEOREM 9.24. *The von Neumann algebras are exactly the C^* -algebras which have a predual, in the above sense.*

PROOF. This is a variation of the above, which caps the above series of results, and closes any further discussions, and for full details here, we refer to Sakai's book [75]. \square

There are many other things that can be said, of purely abstract nature, on the von Neumann algebras. We will be back to this, from time to time, in what follows.

9e. Exercises

Things have been quite tricky in this chapter, with a number of detours, and by avoiding some difficulties, and as unique exercise, which is quite difficult, we have:

EXERCISE 9.25. *Look up and learn von Neumann's reduction theory, stating that given a von Neumann algebra $A \subset B(H)$, if we write its center as*

$$Z(A) = L^\infty(X)$$

then we have a decomposition as follows, with the fibers A_x being factors,

$$A = \int_X A_x dx$$

and then write down a brief account of what you learned.

This is something very fundamental and instructive, because it provides us with a whole new point of view on the factors, and in particular justifies the name "factors". We will be actually back to this later on in this book, but only under the assumption that the algebra has a trace, $tr : A \rightarrow \mathbb{C}$, which simplifies a number of things.

CHAPTER 10

Finite factors

10a. Type II factors

In this chapter we go for the real thing, namely the study of the type II factors, following the work of Murray and von Neumann [60], [61], [62], [87], [88], [89], [90], which is the basis for everything more advanced, in relation with operator algebras.

There are several possible ways of introducing the factors, and also of dividing them afterwards into several classes, for further study. All this can be less or more technical, depending on taste. In what concerns us, we will use a rather intuitive approach.

To be more precise, the general idea, which is quite natural, not based on anything advanced, is that among the von Neumann algebras $A \subset B(H)$, of particular interest are the “free” ones, having trivial center, $Z(A) = \mathbb{C}$. These algebras are called factors:

DEFINITION 10.1. *A factor is a von Neumann algebra $A \subset B(H)$ whose center*

$$Z(A) = A \cap A'$$

which is a commutative von Neumann algebra, reduces to the scalars, $Z(A) = \mathbb{C}$.

Here the fact that the center is indeed a von Neumann algebra follows from the bicommutant theorem, which shows that the commutant of any $*$ -algebra is a von Neumann algebra. Thus, the intersection $Z(A) = A \cap A'$ is a von Neumann algebra as well.

In what follows we will be mainly interested in these factors, with our motivation coming from the fact that the condition $Z(A) = \mathbb{C}$ defining them is somehow opposite to the condition $Z(A) = A$ defining the commutative von Neumann algebras. Thus, the factors are the von Neumann algebras which are “free”, with this meaning being as far as possible from the commutative ones. Equivalently, in terms of the quantum space writing $A = L^\infty(X)$, the spaces X coming from factors are as “free” as one can hope for.

So, this will be our philosophy, looking for purely noncommutative mathematics, by studying the factors. Before going further, however, we should mention that there are some deeper reasons as well for the consideration of the factors, which among others fully justify the term “factor”, coming from the reduction theorem of von Neumann.

This theorem, that we cannot prove at this stage, with our current technology, states that given any von Neumann algebra $A \subset B(H)$, if we write its center as $Z(A) = L^\infty(X)$, then we have a decomposition as follows, with the fibers A_x being factors:

$$A = \int_X A_x dx$$

Moreover, in the case where A has a trace, $tr : A \rightarrow \mathbb{C}$, this trace decomposes as follows, with each $tr_x : A_x \rightarrow \mathbb{C}$ being the restriction of tr to the factor A_x :

$$tr = \int_X tr_x dx$$

Thus, we can see in particular why factors are called factors, because they appear as factors when decomposing the arbitrary von Neumann algebras. More on this later.

Moving ahead now, in order to do probability on our factors we will need a trace as well. Leaving aside the somewhat trivial case $A = M_N(\mathbb{C})$, we are led in this way to:

DEFINITION 10.2. *A II_1 factor is a von Neumann algebra $A \subset B(H)$ which is infinite dimensional, has trivial center, and has a trace $tr : A \rightarrow \mathbb{C}$.*

As a first observation, according to the above-mentioned reduction theorem, such factors are exactly those appearing in the decomposition of the von Neumann algebras $A \subset B(H)$ which have traces, $tr : A \rightarrow \mathbb{C}$, provided that we add some extra axioms which avoid trivial summands, of type $A_x = M_N(\mathbb{C})$. Moreover, by results of Connes, adding to those of von Neumann, and which are non-trivial as well, the non-tracial case basically reduces to the tracial case, via crossed product type operations, and the conclusion is that “the II_1 factors are the building blocks of the von Neumann algebra theory”.

Summarizing, some heavy things going on here. In what follows we will be mainly interested in simple and concrete mathematics, and we will take Definition 10.2 as it is, as a simple and intuitive definition for the “good class” of von Neumann algebras.

As already mentioned, in Definition 10.2 the presence of the trace is the key condition, with this being a safe choice, in view of the fact that traces bring intuition, and technical simplifications, as explained before. Regarding the non-tracial case, we will investigate it later. Finally, the infinite dimensionality assumption is there for distinguishing the II_1 factors from the matrix algebras $M_N(\mathbb{C})$, which are called factors of type I_N .

The II denomination comes from the fact that the theory that we are currently attempting to build, extending the type I theory, can only be of type II . As for the theory that we decided to leave for later, covering as well the factors without traces, this will be of type III . There is as well a reason for the subscript 1, but more on this later.

As a first observation, in practice, and forgetting about reduction theory, which raises the possibility of decomposing any type II von Neumann algebra into factors, in order to obtain explicit examples of II_1 factors, it is not even clear that such beasts exist. Fortunately the group von Neumann algebras are there, and we have the following result, which provides us with some examples of II_1 factors, to start with:

THEOREM 10.3. *The center of a group von Neumann algebra $L(\Gamma)$ is*

$$Z(L(\Gamma)) = \left\{ \sum_g \lambda_g g \mid \lambda_{gh} = \lambda_{hg} \right\}''$$

and if $\Gamma \neq \{1\}$ has infinite conjugacy classes, in the sense that

$$|\{ghg^{-1} \mid g \in G\}| = \infty \quad , \quad \forall h \neq 1$$

with this being called ICC property, the algebra $L(\Gamma)$ is a II_1 factor.

PROOF. There are two assertions here, the idea being as follows:

(1) Consider a linear combination of group elements, which is in the weak closure of $\mathbb{C}[\Gamma]$, and so defines an element of the group von Neumann algebra $L(\Gamma)$:

$$a = \sum_g \lambda_g g$$

By linearity, this element $a \in L(\Gamma)$ belongs to the center of $L(\Gamma)$ precisely when it commutes with all the group elements $h \in \Gamma$, and this gives:

$$\begin{aligned} a \in Z(A) &\iff ah = ha \\ &\iff \sum_g \lambda_g gh = \sum_g \lambda_g hg \\ &\iff \sum_k \lambda_{kh^{-1}k} = \sum_k \lambda_{h^{-1}k}k \\ &\iff \lambda_{kh^{-1}} = \lambda_{h^{-1}k} \end{aligned}$$

Thus, we obtain the formula for $Z(L(\Gamma))$ in the statement.

(2) We have to examine the 3 conditions defining the II_1 factors. We already know, from chapter 7 above, that the group algebra $L(G)$ has a trace, given by:

$$\text{tr}(g) = \delta_{g,1}$$

Regarding now the center, the condition $\lambda_{gh} = \lambda_{hg}$ that we found is equivalent to the fact that $g \rightarrow \lambda_g$ is constant on the conjugacy classes, and we obtain:

$$Z(L(\Gamma)) = \mathbb{C} \iff \Gamma = \text{ICC}$$

Finally, assuming that this ICC condition is satisfied, with $\Gamma \neq \{1\}$, then our group Γ is infinite, and so the algebra $L(\Gamma)$ is infinite dimensional, as desired. \square

In order to look now for more examples of II_1 factors, an idea would be that of attempting to decompose into factors the group von Neumann algebras $L(\Gamma)$, but this is something difficult, and in fact we won't really exit the group world in this way.

Difficult as well is to investigate the factoriality of the von Neumann algebras of discrete quantum groups $L(\Gamma)$, because the basic computations from the proof of Theorem 10.3 won't extend to this setting, where the group elements $g \in \Gamma$ become corepresentations $g \in M_N(L(\Gamma))$. Despite years of efforts, it is presently not clear at all what the "quantum ICC" condition should mean, and the problem comes from this.

In short, we have to stop here the construction of the examples, and Theorem 10.3 will be what we have, at least for the moment. With this being actually not a big issue, the group factors $L(\Gamma)$ being known to be quite close to the generic II_1 factors.

Getting away now from the above difficulties, let us go back to the abstract II_1 factors, as axiomatized in Definition 10.2 above. In order to investigate them, the idea will be that from chapter 9, namely looking at the projections, and their equivalence classes.

In the case of the II_1 factors, as a first interesting remark, the presence of the trace trivializes the proof of the main result that we know about projections, as follows:

THEOREM 10.4. *Given two projections $p, q \in A$ in a II_1 factor, we have*

$$p \preceq q \quad \text{or} \quad q \preceq p$$

and so \preceq is a total order on the equivalence classes of projections $p \in A$.

PROOF. This is something that we know from chapter 9 above, which actually holds for any factor, with the only non-trivial part being the following implication:

$$p \preceq q, q \preceq p \implies p \simeq q$$

But this is clear in the present II_1 factor setting, by using the trace. \square

The above theorem and proof, which is remarkable, was first in a series of mysteries, in what concerns the special case of the II_1 factors. More such mysteries to follow.

In order to study the trace of the II_1 factors, we will need:

PROPOSITION 10.5. *Given a weakly closed left ideal $I \subset A$ in a von Neumann algebra, there exists a unique projection $p \in A$ such that:*

$$I = Ap$$

Moreover, if $I \subset A$ is assumed to be a two-sided ideal, then $p \in Z(A)$.

PROOF. We have several things to be proved, the idea being as follows:

(1) Given an ideal $I \subset A$ as in the statement, consider the following intersection:

$$I \cap I^* \subset A$$

This is a weakly closed non-unital $*$ -subalgebra of A , so if we denote by $p \in A$ its largest projection, or unit, then we have the following inclusion:

$$Ap \subset I$$

(2) Conversely now, let us pick $x \in I$. By polar decomposition we can write $x = u|x|$, and we have the following sequence of implications:

$$\begin{aligned} x \in I &\implies |x| = u^*x \in I \\ &\implies |x| \in I \cap I^* \\ &\implies |x|p = |x| \\ &\implies x = u|x| = u|x|p \in Ap \end{aligned}$$

Thus, we have proved the reverse inclusion $I \subset Ap$.

(3) The uniqueness assertion is clear from the comparison theorem for projections.

(4) Regarding now the last assertion, assume that $I \subset A$ is a two-sided weakly closed ideal. Then for any unitary $u \in A$ we have:

$$\begin{aligned} I = uIu^* &\implies uIu^* = Ap \\ &\implies I = Aupu^* \end{aligned}$$

Thus by uniqueness we obtain $upu^* = p$, and so $p \in Z(A)$, as desired. \square

As a first main result now regarding the II_1 factors, we have:

THEOREM 10.6. *Given a II_1 factor A , any weakly continuous positive trace*

$$tr : A \rightarrow \mathbb{C}$$

is automatically faithful.

PROOF. Consider the null space of the trace, which is by definition:

$$I = \left\{ x \in A \mid tr(x^*x) = 0 \right\}$$

We have the following inequality, which shows that I is a left ideal:

$$x^*a^*ax \leq \|a\|^2 x^*x$$

Now by using the trace condition $tr(ab) = tr(ba)$, we conclude that I is a two-sided ideal. Also, the Cauchy-Schwarz inequality gives:

$$tr(x^*x) = 0 \iff tr(xy) = 0, \forall y \in A$$

We conclude from this that I is an intersection of kernels of weakly closed functionals, which are weakly closed, and so it is weakly closed. Thus the last assertion in Proposition 10.5 applies, and produces a projection $p \in Z(A)$ such that:

$$I = Ap$$

Now since A was assumed to be a factor, we have $Z(A) = \mathbb{C}$. Thus $p = 0$, and so the null ideal of the trace is $I = \{0\}$, and so our trace tr is faithful, as desired. \square

Our goal now will be that of proving that the trace on a II_1 factor is unique, and takes on projections any value in $[0, 1]$. Let us start with a technical result, as follows:

PROPOSITION 10.7. *Given a II_1 factor A , the traces of the projections*

$$tr(p) \in [0, 1]$$

can take arbitrarily small values.

PROOF. Consider the set formed by all values of the trace on the projections:

$$S = \left\{ tr(p) \mid p^2 = p = p^* \in A \right\}$$

We want to prove that the following number equals 0:

$$c = \inf(S - \{0\})$$

In order to do so, assume by contradiction $c > 0$, pick $\varepsilon > 0$ small, and pick a projection $p \in A$ such that:

$$tr(p) < c + \varepsilon$$

Since we are in a II_1 factor, this projection $p \in A$ cannot be minimal, and so we can find another projection $q \in A$ satisfying $q < p$. Now observe that we have:

$$\begin{aligned} tr(p - q) &= tr(p) - tr(q) \\ &\leq tr(p) - c \\ &\leq \varepsilon \end{aligned}$$

Thus with $\varepsilon < c$ we obtain a contradiction, and so $c = 0$, as desired. \square

In order to prove now our main result, we will need:

PROPOSITION 10.8. *Given a II_1 factor A on a Hilbert space H and a projection $p \in A$, the von Neumann algebra pAp is a II_1 factor on the Hilbert space pH .*

PROOF. We have to prove that the von Neumann algebra pAp has a trace, and is infinite dimensional, and these two properties can be proved as follows:

(1) In what regards the trace, we know that the trace $tr : A \rightarrow \mathbb{C}$ restricts to a trace $tr : pAp \rightarrow \mathbb{C}$, which must be nonzero, as desired.

(2) In what regards the infinite dimensionality, this follows from the fact that a minimal projection in pAp would be minimal in A , which is impossible. \square

We can now formulate a first main result regarding the II_1 factors, as follows:

THEOREM 10.9. *Given a II_1 factor A , the traces of the projections*

$$\text{tr}(p) \in [0, 1]$$

can take any values in $[0, 1]$.

PROOF. Given a number $c \in [0, 1]$, consider the following set:

$$S = \left\{ p^2 = p = p^* \in A \mid \text{tr}(p) \leq c \right\}$$

This set satisfies the assumptions of the Zorn lemma, and so by this lemma we can find a maximal element $p \in S$. Assume by contradiction that:

$$\text{tr}(p) < c$$

Now by using Proposition 10.7 and Proposition 10.8 above, we can slightly enlarge the trace of p , and we obtain a contradiction, as desired. \square

As a second main result now regarding the II_1 factors, we have:

THEOREM 10.10. *The trace of a II_1 factor*

$$\text{tr} : A \rightarrow \mathbb{C}$$

is unique.

PROOF. This can be proved in many ways, a standard one being that of proving that any two traces agree on the projections, as a consequence of the above results:

(1) Assume indeed that we have a second trace $\text{tr}' : A \rightarrow \mathbb{C}$. Since A is generated by its projections, it is enough to show that we have $\text{tr} = \text{tr}'$ on projections.

(2) As a first observation, since traces on matrix algebras are unique, we obtain that we have $\text{tr} = \text{tr}'$ on the projections $p \in A$ having rational trace, $\text{tr}(p) \in \mathbb{Q}$.

(3) So, let us pick $p \in A$ having non-rational trace, $\text{tr}(p) \notin \mathbb{Q}$, and prove that we have $\text{tr}(p) = \text{tr}'(p)$. The idea will be that of using the result for the projections having rational traces, applied to an infinite direct sum of projections, converging to p .

(4) To be more precise, assume that we have constructed our sequence $p_i \rightarrow p$ up to order $n \in \mathbb{N}$, and let us try to construct p_{n+1} . The idea is to use the following algebra:

$$A_n = (p - p_n)A(p - p_n)$$

(5) Indeed this algebra is a II_1 factor, and we can choose inside it a projection p_{n+1} satisfying $p_n \leq p_{n+1} \leq p$, such that $\text{tr} = \text{tr}'$ on it, and such that:

$$\text{tr}(p - p_{n+1}) \leq \frac{1}{2} \cdot \text{tr}(p - p_n)$$

(6) According to our choices for these projections p_n , we have:

$$p = \bigvee_{n=1}^{\infty} p_n$$

Thus when evaluating tr, tr' on p we obtain the same result, as desired. \square

This was for the basic theory of the II_1 factors. There are many other things that can be said, regarding the general theory of the arbitrary II_1 factors, ranging from rather soft to very technical, and we will be back to this on several occasions, in what follows.

In what regards the examples, we have so far the group von Neumann algebras $L(\Gamma)$, with Γ being an ICC group. In certain cases, it is possible to say more about all the above, and in particular about the projections, for instance with quite explicit procedures for constructing projections $p \in L(\Gamma)$ having an arbitrary prescribed trace $x \in [0, 1]$.

As before, we refer here to the literature on the subject, and notably to the books by Dixmier, and to the related research papers. We will be back to this later, in chapter 11 below, when discussing more in detail the group von Neumann algebras $L(\Gamma)$, and their quantum group generalizations, from the reduction theory perspective.

10b. Standard form

We have seen that the II_1 factors are very interesting objects, naturally upgrading the matrix algebras $M_N(\mathbb{C})$, which are type I factors. From this perspective, a II_1 factor $A \subset B(H)$ is not really in need of the ambient Hilbert space H , and the question of “representing” it appears. We will discuss this question, in two steps:

- (1) A first question is that of understanding the possible embeddings $A \subset B(H)$, with H being a Hilbert space. The main result here will be the construction of a numeric invariant $\dim_A H$, called coupling constant.
- (2) A second question is that of understanding the possible embeddings $A \subset B$, with B being another II_1 factor. By using the coupling constant for both A, B we will construct a numeric invariant $[B : A]$, called index.

We will discuss here (1), and leave (2) for later, towards the end of this book. In order to get started, let us formulate the following definition:

DEFINITION 10.11. *Given a von Neumann algebra A with a trace $tr : A \rightarrow \mathbb{C}$, the emdedding*

$$A \subset B(L^2(A))$$

obtained by GNS construction is called standard form of A .

Here we use the GNS construction, explained in chapter 7 above. As the name indicates, the standard representation is something “standard”, to be compared with any other representation $A \subset B(H)$, in order to understand this latter representation.

As already seen in chapter 7, the GNS construction has a number of unique features, that can be exploited. In the present setting, the main result is as follows:

THEOREM 10.12. *In the context of the standard representation we have*

$$A' = JAJ$$

with $J : L^2(A) \rightarrow L^2(A)$ being the antilinear map given by $T \rightarrow T^*$.

PROOF. Observe first that any $T \in A$ can be regarded as a vector $T \in L^2(A)$, to which we can associate, in an antilinear way, the vector $T^* \in L^2(A)$. Thus we have indeed an antilinear map J as in the statement. In terms of the standard cyclic and separating vector Ω for the GNS representation, the formula of this formula J is:

$$J(x\Omega) = x^*\Omega$$

(1) Our first claim is that we have the following formula:

$$\langle J\xi, J\eta \rangle = \langle \xi, \eta \rangle$$

Indeed, with $\xi = x\Omega$ and $\eta = y\Omega$, we have the following computation:

$$\begin{aligned} \langle J\xi, J\eta \rangle &= \langle yx^*\Omega, \Omega \rangle \\ &= \text{tr}(yx^*) \\ &= \langle \xi, \eta \rangle \end{aligned}$$

(2) Our second claim is that we have the following formula:

$$JxJ(y\Omega) = yx^*\Omega$$

Indeed, this follows from the following computation:

$$\begin{aligned} JxJ(y\Omega) &= J(xy^*\Omega) \\ &= yx^*\Omega \end{aligned}$$

(3) Our claim now is that we have an inclusion as follows:

$$JAJ \subset A'$$

Indeed, this follows from the formula obtained in (2) above.

(4) In order to prove the reverse inclusion, our claim is that for $x \in A'$ we have:

$$Jx\Omega = x^*\Omega$$

Indeed, this follows from the following computation, valid for any $y \in A$:

$$\begin{aligned} \langle Jx\Omega, y\Omega \rangle &= \langle Jy\Omega, x\Omega \rangle \\ &= \langle y^*\Omega, x\Omega \rangle \\ &= \langle \Omega, xy\Omega \rangle \\ &= \langle x^*\Omega, y\Omega \rangle \end{aligned}$$

(5) Our claim now is that the following formula defines a trace on A' :

$$Tr(x) = \langle x\Omega, \Omega \rangle$$

Indeed, for any two elements $x, y \in A'$ we have:

$$\begin{aligned} \langle xy\Omega, \Omega \rangle &= \langle y\Omega, x^*\Omega \rangle \\ &= \langle y\Omega, Jx\Omega \rangle \\ &= \langle x\Omega, Jy\Omega \rangle \\ &= \langle x\Omega, y^*\Omega \rangle \\ &= \langle yx\Omega, \Omega \rangle \end{aligned}$$

(6) We can now finish the proof. Indeed, by using the trace constructed in (5), we can apply our results obtained so far to A' , and we obtain $JA'J \subset A$, as desired. \square

As a basic illustration for the above result, we have:

THEOREM 10.13. *The commutant of a von Neumann group algebra $L(\Gamma)$, which is obtained by definition by using the left regular representation, is the von Neumann group algebra $R(\Gamma)$, obtained by using the right regular representation.*

PROOF. This is indeed clear from Theorem 10.12 above. Observe that, as a consequence, the center of a group von Neumann algebra appears as follows:

$$Z(L(\Gamma)) = L(\Gamma) \cap R(\Gamma)$$

We will be back to this in chapter 11 below, when doing reduction theory. \square

As another application of the standard representation, let us go back to the uniqueness of the trace, that we know from Theorem 10.10 above. There are as well several alternative proofs for this fact, which are all instructive. As a first such statement and proof, which is something quite beautiful, and also technically very useful, we have:

THEOREM 10.14. *Given a II_1 factor A , and an element $a \in A$, we have the following Dixmier averaging property:*

$$\overline{\text{span} \left\{ uau^* \mid u \in U_A \right\}}^w \cap \mathbb{C}1 \neq \emptyset$$

In particular, the II_1 factor trace $tr : A \rightarrow \mathbb{C}$ is unique.

PROOF. We use the basic theory of the regular representation $A \subset L^2(A)$, with respect to the given trace $tr : A \rightarrow \mathbb{C}$, explained above. The proof goes as follows:

(1) Given an element $a \in A$, consider the space in the statement, obtained as the weak closure of the space spanned by the spinned versions of a , namely:

$$K_a = \overline{\text{span} \left\{ uau^* \mid u \in U_A \right\}}^w$$

This linear space $K_a \subset A$ is by definition weakly closed, and it follows that the subset $K_a\Omega \subset L^2(A)$, where $\Omega \in L^2(A)$ is the canonical trace vector, is a weakly closed convex subset. In particular, we see that $K_a\Omega \subset L^2(A)$ is a norm closed convex subset.

(2) In view of this, we can consider the unique element $b \in K_a$ having the property that $b\Omega$ has a minimal norm. We have then the following formula, for any unitary $u \in U_A$, where $J : L^2(A) \rightarrow L^2(A)$ is the standard antilinear map, given by $T \rightarrow T^*$:

$$\|uJuJb\Omega\| = \|b\Omega\|$$

By uniqueness of b , it follows that for any unitary $u \in U_A$, we have:

$$uJuJb\Omega = b\Omega$$

But this shows that for any unitary $u \in U_A$, we have:

$$ubu^* = b$$

We conclude that we have $b \in \mathbb{C}1$, and this proves the first assertion.

(3) Regarding now the second assertion, consider an arbitrary trace $tr : A \rightarrow \mathbb{C}$. By using $tr(uau^*) = tr(a)$, we conclude that this trace is constant on the following set:

$$K_a = \overline{\text{span} \left\{ uau^* \mid u \in U_A \right\}}^w$$

Now by using the first assertion, we conclude that we have the following formula:

$$\overline{\text{span} \left\{ uau^* \mid u \in U_A \right\}}^w \cap \mathbb{C}1 = \{tr(a)1\}$$

Summarizing, we have obtained a purely algebraic formula for our trace $tr : A \rightarrow \mathbb{C}$, and it follows that this trace is indeed unique, as claimed. \square

In relation with the above, let us mention that there is as well a third proof for the uniqueness of the trace, based on nothing or almost, meaning the definition of II_1 factors, and some abstract functional analysis. We refer here to the paper of Yeadon [100].

We will be back to more generalities regarding the II_1 factors in a moment, with this time some general theory regarding their representations, inside other II factors, $A \subset B$, as outlined in the above, after taking a short break, and looking for examples.

10c. Basic examples

Before developing more general theory for the II_1 factors, let us discuss the examples. We have so far only one class of examples, namely the group von Neumann algebras $L(\Gamma)$, which are II_1 factors precisely when the discrete groups Γ have the ICC property.

This suggests doing several things, in order to have more examples, as follows:

(1) A first idea is that of looking at the von Neumann algebras of discrete quantum groups, $A = L(\Gamma)$. Indeed, we can write $A = L^\infty(G)$, with $G = \widehat{\Gamma}$ being the compact quantum group dual to Γ , and so we are into familiar territory, namely that of the Woronowicz algebras, or rather of their weak closures, developed in chapter 7. However, despite years of efforts, no one really knows what “quantum ICC” should mean.

(2) Along the same lines, and a bit more modestly, a natural conjecture would be that if a compact quantum group $G \subset U_N^+$ appears as a “liberation” of a classical group $G_{class} \subset U_N$, in the technical liberation sense explained in chapters 7-8 above, then the corresponding von Neumann algebra $A = L^\infty(G)$ should be a II_1 factor. However, while there are many known such results, no one knows how to do this in general.

(3) As a variation of the above freeness conjecture, which is more general, and takes us away from the group algebras, a conjecture would be that if a homogeneous space $X = G/H$, or more general manifold X , appears as a “liberation” of a homogeneous space $X_{class} = G_{class}/H_{class}$, or of a more general manifold X_{class} , then the corresponding von Neumann algebra $A = L^\infty(X)$ should be a II_1 factor. This is difficult as well.

(4) Along the same lines, but having this time von Neumann’s reduction theory results in mind, we have the question of understanding how the various algebras considered above, namely $L(\Gamma)$ with Γ being a discrete group, or $L(\Gamma) = L^\infty(G)$ with $\Gamma = \widehat{G}$ being a discrete quantum group, or $L^\infty(X)$, with $X = G/H$ being a quantum homogeneous space, or a more general quantum manifold, decompose as sums of II_1 factors. Difficult, again.

Summarizing, we have many interesting questions here, which are all related to each other, and which are all difficult, and with the Holy Grail being the reduction theory for the algebras of type $A = L^\infty(X)$, with X being a quantum manifold. We will be back to some of these questions in chapter 11 below, when talking about reduction theory.

Fortunately, the above questions (1-2-3-4), all difficult, are not the only possible ones, and we have as well a series of alternative questions (5-6-7-8), also inspired by the group von Neumann algebras $L(\Gamma)$, and which are supposedly easier, as follows:

(5) A group von Neumann algebra $L(\Gamma)$ can be thought of as coming from the trivial action $\Gamma \curvearrowright \{\cdot\}$, and the question is that of investigating von Neumann algebras associated to more general actions, $\Gamma \curvearrowright X$, by using various crossed product techniques.

(6) There are many natural examples of compact groups G acting on von Neumann algebras P , and the question is that of understanding under which exact assumptions on the action $G \curvearrowright P$, the corresponding fixed point algebra P^G is a factor.

(7) There are as well many examples of discrete groups Γ acting on von Neumann algebras R , and the question is that of understanding under which exact assumptions on the action $G \curvearrowright R$, the corresponding crossed product algebra $R \rtimes \Gamma$ is a factor.

(8) Finally, the above questions are related to each other, and even more general questions come by looking at actions of compact quantum groups G , or discrete quantum groups Γ , on various quantum spaces X , or von Neumann algebras P or R .

All this is good news, so work for us to be done here. In what follows we will discuss a bit the questions (1-2-3-4) above, with some general conjectures and comments. Then we will discuss, a bit more in detail, (5-6-7-8), following some standard work of Popa, Wassermann and others. We will relegate the most difficult questions for later.

In order to get started, in connection with (1-2-3-4), let us first talk about free quantum groups, as a continuation of the material from chapters 7-8 above. The various combinatorial considerations there lead to the following notion, from [15]:

DEFINITION 10.15. *Given an easy group $G \subset U_N$, coming from a category of partitions*

$$D = (D(k, l))$$

its free version is the easy quantum group $G^+ \subset U_N^+$ coming from the following category:

$$D^- = D \cap NC$$

In this case, we also say that the algebra $L^\infty(G^+)$ is the free version of $L^\infty(G)$.

With this definition in hand, the conjecture would be that the von Neumann algebras of type $L^\infty(G^+)$ are factors. Although verifications of this conjecture abound, it is not clear how to attack the question, in general. See [20], [21], [66], [78], [81], [82].

As a first observation, the easiness condition on G , as explained in chapter 7 above, implies that this group appears as an intermediate object, as follows:

$$S_N \subset G \subset U_N$$

Regarding now its liberation G^+ , the categorial condition $D^- = D \cap NC$ which defines it tells us that this must appear as an intermediate object, as follows:

$$S_N^+ \subset G^+ \subset U_N^+$$

With this observation in hand, which is something quite trivial, it is tempting to simply forget about easiness, and formulate the following definition:

DEFINITION 10.16. *A quantum group $G \subset U_N^+$ is called free when it appears as follows:*

$$S_N^+ \subset G^+ \subset U_N^+$$

In this case, we also say that the algebra $L^\infty(G)$ is free.

With this new definition in hand, the conjecture would be that the von Neumann algebras $L^\infty(G)$ which are free in the above sense are factors. But again, the subject remains quite technical, and it is not clear how to attack the question, in general.

Yet another approach comes from the notion of free complexification. This is a quite tricky operation, which makes sense for quantum groups only, as follows:

DEFINITION 10.17. *The free complexification of a Woronowicz algebra (A, u) with $u \in M_N(A)$ is the Woronowicz algebra (\tilde{A}, \tilde{u}) obtained by setting*

$$\tilde{A} = \langle zu_{ij} \rangle \subset C(\mathbb{T}) * A$$

$$\tilde{u} = zu \in M_N(\tilde{A})$$

where $z \in C(\mathbb{T})$ is the standard generator, given by $x \rightarrow x$ for any $x \in \mathbb{T}$.

The point indeed with this notion is that, in the context of the liberation operation constructed in Definition 10.15 above, we usually have embeddings as follows:

$$G \subset \tilde{G} \subset G^+$$

Thus, we are led into a conjecture about free complexifications, as follows:

CONJECTURE 10.18. *Given a closed subgroup $G \subset U_N^+$, the von Neumann algebra*

$$A = L^\infty(\tilde{G})$$

of L^∞ functions on its free complexification $\tilde{G} \subset U_N^+$ is a factor.

Observe that this is not exactly the same thing as the freeness conjectures mentioned above, but in view of all the above, it is probably something quite close to them.

Importantly, the above conjecture is something far more general and flexible than everything coming from easiness. For instance the conjecture makes sense for any group dual $G = \hat{\Gamma}$, and we even have a positive result here, as follows:

THEOREM 10.19. *The factoriality conjecture for free complexifications holds for any group dual, $G = \hat{\Gamma}$, with $\Gamma = \langle g_1, \dots, g_N \rangle$ being a discrete group.*

PROOF. Let us go back to the construction in Definition 10.17. By using discrete quantum group notations for the algebras involved, namely $A = C^*(\Gamma)$ and $\tilde{A} = C^*(\tilde{\Gamma})$, and also by replacing the algebra $C(\mathbb{T})$ with the algebra $C^*(\mathbb{Z})$, which is isomorphic to it, we conclude that the free complexification operation appears as follows:

$$C^*(\tilde{\Gamma}) = \langle zu_{ij} \rangle \subset C^*(\mathbb{Z}) * C^*(\Gamma)$$

$$\tilde{u} = zu \in M_N(C^*(\tilde{\Gamma}))$$

Now in the usual group dual case, where $\Gamma = \langle g_1, \dots, g_N \rangle$ is a usual discrete group, this shows that $\tilde{\Gamma}$ is a usual discrete group as well, appearing as follows:

$$\tilde{\Gamma} = \langle zg_i \rangle \subset \mathbb{Z} * \Gamma$$

But such discrete groups are easily seen to have the ICC property, and so the corresponding von Neumann algebras $L(\tilde{\Gamma})$ are factors, as claimed. \square

In general, it is quite unclear how to approach Conjecture 10.18, but everything tends to point to the free probability techniques from chapter 8. Indeed, by using standard results about the Haar functionals of free products, from [91], we see that, in the context of Definition 10.17, we have an embedding of von Neumann algebras, as follows:

$$L^\infty(\tilde{G}) \subset L^\infty(\mathbb{T}) * L^\infty(G)$$

Alternatively, in terms of the associated discrete quantum groups, as in the proof of Theorem 10.19, we have an embedding of von Neumann algebras, as follows:

$$L(\tilde{\Gamma}) \subset L(\mathbb{Z}) * L(\Gamma)$$

It is possible to further build along these lines, with a purely free probabilistic formulation of the factoriality question. However, in practice, all this remains quite complicated. Illustrating here is the simplest case, that of the group $G = U_N$, with the free complexification and factoriality questions being noted since [2], but solved there only at $N = 2$, and with the general case, $N \in \mathbb{N}$, being solved only much later, in [82].

Finally, as a last topic regarding the factoriality of the free versions and free complexifications, let us discuss now the general quantum manifold case. Generally speaking, the conjecture here would be that if a quantum manifold X is free, in some suitable algebraic sense, then its associated von Neumann algebra $L^\infty(X)$ should be a factor.

However, things are quite tricky here, and even formulating a precise conjecture in this sense turns to be a non-trivial task. Indeed, in order to talk about $L^\infty(X)$ we must be able to integrate over X , and so our quantum manifold X must be Riemannian, in some suitable sense. But the most known and straightforward axiomatization of the quantum Riemannian manifolds, due to Connes [24], does unfortunately not apply to the free case, precisely, and so we are left with some difficult axiomatization questions here.

Moving ahead from these difficulties, one idea, more modest, would be that of talking about quotient spaces $X = G/H$ only, with $H \subset G$ being compact quantum groups, because such spaces can be shown to have Haar measures, so at least our conjecture would make sense. However, there are some difficulties here too, because the free spheres discussed in chapter 7, which normally should be our main examples here, do not exactly appear as such quotient spaces, due to a number of algebraic and analytic issues.

The solution to these difficulties, or at least the best solution known so far, comes from the notion of affine homogeneous space, which is as follows:

DEFINITION 10.20. *An affine homogeneous space over a closed subgroup $G \subset U_N^+$ is a closed subset $X \subset S_{\mathbb{C},+}^{N-1}$, together with an index set $I \subset \{1, \dots, N\}$, such that*

$$\alpha(x_i) = \frac{1}{\sqrt{|I|}} \sum_{j \in I} u_{ji}$$

$$\Phi(x_i) = \sum_j x_j \otimes u_{ji}$$

define morphisms of C^ -algebras, satisfying the following condition:*

$$\left(id \otimes \int_G \right) \Phi = \int_G \alpha(\cdot) 1$$

This definition is something quite tricky. As a first observation, due to the above explicit formulae for the maps α, Φ , the following conditions are satisfied:

$$(\Phi \otimes id)\Phi = (id \otimes \Delta)\Phi$$

$$(\alpha \otimes id)\Phi = \Delta\alpha$$

At the level of the examples, both the quotient spaces $O_N^+ \rightarrow S_{\mathbb{R},+}^{N-1}$ and $U_N^+ \rightarrow S_{\mathbb{C},+}^{N-1}$, that we know from chapter 7, are affine in the above sense, with $I = \{1\}$. There are many other examples, and things that can be said about the affine homogeneous spaces, and getting back now to our von Neumann algebra questions, we can formulate:

CONJECTURE 10.21. *Given an affine homogeneous space $G \rightarrow X$, the algebra*

$$A = L^\infty(\tilde{X})$$

of L^∞ functions on its free complexification $\tilde{G} \rightarrow \tilde{X}$ is a factor.

This conjecture is something quite general, and having it formulated is certainly a good thing. However, in what regards a potential proof, things are difficult here.

This was for our basic discussion of factoriality questions, in the free geometry setting. There are some further things that can be said, on one hand in relation with reduction theory, and on the other hand, in relation with various group actions, fixed point algebras, and crosses products. We will be back to this on several occasions, in what follows.

10d. Coupling constant

Let us go back now to the general theory of the II_1 factors, with the aim of talking about representations of such II_1 factors, inside the category of the II_1 factors, $A \subset B$. For this purpose we will need a key notion, called coupling constant.

In order to discuss the construction of the coupling constant, we will need some further results on the type II factors, complementing those that we already have. The point indeed is that the class of II factors, to be axiomatized later, and with this being not something urgent, comprises, besides the II_1 factors discussed above, the II_∞ factors as well:

DEFINITION 10.22. *A II_∞ factor is a von Neumann algebra of the form*

$$B = A \otimes B(H)$$

with A being a II_1 factor, and with H being an infinite dimensional Hilbert space.

There are many things that can be said about the II_∞ factors, and as a first such result, which is a technically useful characterization of such factors, we have:

PROPOSITION 10.23. *For an infinite factor B , the following are equivalent:*

- (1) *There exists a projection $p \in B$ such that pBp is a II_1 factor.*
- (2) *B is a II_∞ factor.*

PROOF. This is something elementary, as follows:

(1) \implies (2) Assume indeed that $p \in B$ is a projection such that pBp is a II_1 factor. We choose a maximal family of pairwise orthogonal projections $\{p_i\} \subset B$ satisfying $p_i \simeq p$, for any i , and we consider the following projection, which satisfies $q \preceq p$:

$$q = 1 - \sum_i p_i$$

Since the indexing set for our set of projections $\{p_i\}$ must be infinite, we can use a strict embedding of this index set into itself, as to write a formula as follows:

$$\begin{aligned} 1 &= q + \sum_i p_i \\ &\preceq p_0 + \sum_{i \neq 0} p_i \\ &\preceq 1 \end{aligned}$$

Thus we have $\sum_i p_i \simeq 1$, and we may further suppose that we have in fact:

$$\sum_i p_i = 1$$

Thus the family $\{p_i\}$ can be used in order to construct a copy $B(H) \subset B$, with $H = l^2(\mathbb{N})$, and we must have $B = A \otimes B(H)$, with A being a II_1 factor, as desired.

(2) \implies (1) This is clear, because when assuming $B = A \otimes B(H)$, as in Definition 10.22, we can take our projection $p \in B$ to be of the form $p = 1 \otimes q$, with $q \in B(H)$ being a rank 1 projection, and we have then $pBp = A$, which is a II_1 factor, as desired. \square

Getting back now to the original interpretation of the II_∞ factors, from Definition 10.22 above, the tensor product writing there $B = A \otimes B(H)$ suggests tensoring the trace of the II_1 factor A with the usual operator trace of $B(H)$. We are led in this way to:

DEFINITION 10.24. *Given a II_∞ factor B , written as $B = A \otimes B(H)$, with A being a II_1 factor and with H being an infinite dimensional Hilbert space, we define a map*

$$\begin{aligned} \text{tr} : B_+ &\rightarrow [0, \infty] \\ \text{tr}((x_{ij})) &= \sum_i \text{tr}(x_{ii}) \end{aligned}$$

where we have chosen a basis of H , as to have $H \simeq l^2(\mathbb{N})$, and so $B(H) \subset M_\infty(\mathbb{C})$.

As an important observation, to start with, unlike in the II_1 factor case, that of the factor A , or in the I_∞ factor case, that of the factor $B(H)$, it is not possible to suitably normalize the trace constructed above. On the positive side now, this trace that we constructed has all sorts of good properties, which can be summarized as follows:

PROPOSITION 10.25. *The II_∞ factor trace $\text{tr} : B_+ \rightarrow [0, \infty]$ constructed above has the following properties:*

- (1) $\text{tr}(x + y) = \text{tr}(x) + \text{tr}(y)$, and $\text{tr}(\lambda x) = \lambda \text{tr}(x)$ for $\lambda \geq 0$.
- (2) If $x_i \nearrow x$ then $\text{tr}(x_i) \rightarrow \text{tr}(x)$.
- (3) $\text{tr}(xx^*) = \text{tr}(x^*x)$.
- (4) $\text{tr}(uxu^*) = \text{tr}(x)$ for any $u \in U_B$.

PROOF. All this is obvious, with (1) being clear from definitions, (2) and (3) being elementary as well, and finally with (4) coming from (3), via $uxu^* = u\sqrt{x} \cdot \sqrt{x}u^*$. \square

As a main result now regarding the II_∞ factor trace, we have:

THEOREM 10.26. *The II_∞ factor trace $\text{tr} : B_+ \rightarrow [0, \infty]$ constructed above, when restricted to the projections*

$$\text{tr} : P(B) \rightarrow [0, \infty]$$

induces an isomorphism between the totally ordered set of equivalence classes of projections in B and the interval $[0, \infty]$.

PROOF. We have several things to be checked here, as follows:

- (1) Our first claim is that a projection $p \in B$ is finite precisely when $\text{tr}(p) < \infty$.

Indeed, in one sense, assume that we have $\text{tr}(p) < \infty$. If our projection p was to be infinite, we would have a subprojection $q \leq p$ having the same trace as p , and so $r = p - q$ would be a projection of trace 0, which is impossible. Thus p is indeed finite.

In the other sense now, assuming $tr(p) = \infty$, we have to prove that p is infinite. For this purpose, let us pick a projection $q \leq p$ having finite trace. Then $r = p - q$ satisfies $tr(r) = \infty$, and so we can iterate the procedure, and we end up with an infinite sequence of pairwise orthogonal projections, which are all smaller than p . But this shows that p dominates an infinite projection, and so that p itself is infinite, as desired.

(2) Our second claim is that if $p, q \in B$ are projections, with p finite, then:

$$p \preceq q \iff tr(p) = tr(q)$$

But this is something which follows exactly as in the II_1 factor case, discussed above.

(3) Our third and final claim, which will finish the proof, is that any infinite projection is equivalent to the identity. For this purpose, assume that $p \in B$ is infinite. By definition, this means that we can find a unitary $u \in B$ such that:

$$uu^* = p$$

$$u^*u \leq p \quad , \quad uu^* \neq p$$

But these conditions show that $(u^n)^i u^n$ is a strictly decreasing sequence of equivalent projections, and by using this sequence we conclude that we have $1 \preceq p$, as desired. \square

Moving ahead now, in order to further investigate the II_∞ factors, we will need:

THEOREM 10.27. *Given a II_1 factor $A \subset B(H)$, there exists an isometry*

$$u : H \rightarrow L^2(A) \otimes l^2(\mathbb{N})$$

such that $ux = (x \otimes 1)u$, for any $x \in A$.

PROOF. We use a standard idea, that we used many times before, namely an amplification trick. Given a II_1 factor $A \subset B(H)$, consider the following Hilbert space:

$$K = H \oplus L^2(A) \otimes l^2(\mathbb{N})$$

Consider, as operators over this space K , the following projections:

$$p = id \oplus 0$$

$$q = 0 \oplus id$$

Both these projections p, q belong then to A' , which is a type II_∞ factor. Now since $q \in A'$ is infinite, by Theorem 10.26 we can find a partial isometry $u \in A'$ such that:

$$u^*u = p$$

$$uu^* \leq q$$

Now let us represent this partial isometry $u \in B(K)$ as a 2×2 matrix, as follows:

$$u = \begin{pmatrix} a & b \\ c & d \end{pmatrix}$$

The above conditions $u^*u = p$ and $uu^* \leq q$ reformulate then as follows:

$$b^*b + d^*d = 0$$

$$aa^* + bb^* = 0$$

We conclude that our partial isometry $u \in B(K)$ has the following special form:

$$u = \begin{pmatrix} 0 & 0 \\ c & 0 \end{pmatrix}$$

But the operator $c : H \rightarrow l^2(A) \otimes l^2(\mathbb{N})$ that we found in this way must be an isometry, and from $u \in A'$ we obtain $ux = (x \otimes 1)u$, for any $x \in A$, as desired. \square

As a basic consequence of the above result, which is something good to know, and that we will use many times in what follows, we have:

THEOREM 10.28. *The commutant of a II_1 factor is a II_1 factor, or a II_∞ factor.*

PROOF. This follows indeed from the explicit interpretation of the operator algebra embedding $A \subset B(H)$ of our II_1 factor A , found in Theorem 10.27 above. \square

Summarizing, we have an extension of the general theory of the II_1 factors, developed before, to the general case of the type II factors, which comprises by definition the II_1 factors and the II_∞ factors. All this is of course technically very useful.

With this discussion made, we are now in position of constructing the coupling constant, which will eventually close the discussion regarding the representations of the II_1 factors, and will lead as well to a whole number of new perspectives on the II_1 factors.

The idea here will be that given a representation of a II_1 factor $A \subset B(H)$, we can try to understand how far is this representation from the standard form, where $H = L^2(A)$, from “above” or from “below”. In order to discuss this, let us start with:

PROPOSITION 10.29. *Given a II_1 factor $A \subset B(H)$, with its embedding into $B(H)$ being represented as above, in terms of an isometry*

$$u : H \rightarrow L^2(A) \otimes l^2(\mathbb{N})$$

$$ux = (x \otimes 1)u$$

the following quantity does not depend on the choice of this isometry u :

$$C = \text{tr}(uu^*)$$

Moreover, for the standard form, where $H = L^2(A)$, this constant takes the value 1.

PROOF. Assume indeed that we have an isometry u as in the statement, and that we have as well a second such isometry, of the same type, namely:

$$v : H \rightarrow L^2(A) \otimes l^2(\mathbb{N})$$

$$vx = (x \otimes 1)v$$

We have then $uu^* = uv^*vu^*$, and by using this, we obtain:

$$\begin{aligned} C_u &= \operatorname{tr}(uu^*) \\ &= \operatorname{tr}(uv^*vu^*) \\ &= \operatorname{tr}(vu^*uv^*) \\ &= \operatorname{tr}(vv^*) \\ &= C_v \end{aligned}$$

Thus, we are led to the conclusion in the statement. As for the last assertion, regarding the standard form, this is clear from definitions, because here we can take $u = 1$. \square

As a conclusion to all this, given a II_1 factor $A \subset B(H)$, we know from Theorem 10.27 above that H must appear as an “inflated” version of $L^2(A)$. The corresponding inflation constant is a certain number, that we can call coupling constant, as follows:

DEFINITION 10.30. *Given a representation of a II_1 factor $A \subset B(H)$, we can talk about the corresponding coupling constant, as being the number*

$$\dim_A H \in (0, \infty]$$

constructed as follows, with $u : H \rightarrow L^2(A) \otimes l^2(\mathbb{N})$ isometry satisfying $ux = (x \otimes 1)u$:

$$\dim_A H = \operatorname{tr}(uu^*)$$

For the standard form, where $H = L^2(A)$, this coupling constant takes the value 1.

This definition might seem a bit complicated, but things here are quite non-trivial, and there is no way of doing something substantially simpler. Alternatively, we can define the coupling constant via the following formula, after proving first that the number on the right is indeed independent of the choice on a nonzero vector $x \in H$:

$$\dim_A H = \frac{\operatorname{tr}_A(P_{A'x})}{\operatorname{tr}_{A'}(P_{Ax})}$$

This latter formula was in fact the original definition of the coupling constant, by Murray and von Neumann. However, technically speaking, things are slightly easier when using the approach in Definition 10.30 above. We will be back to this key formula of Murray and von Neumann, with full explanations, and a proof, in a moment.

The coupling constant is a key notion, having numerous remarkable properties, which are all useful, and that we will explore now. First, we have the following result:

PROPOSITION 10.31. *The coupling constant $\dim_A H \in (0, \infty]$ associated to a II_1 factor representation $A \subset B(H)$ has the following properties:*

- (1) *For the standard form, $H = L^2(A)$, we have $\dim_A H = 1$.*
- (2) *For the usual representation on $H = L^2(A) \otimes l^2(\mathbb{N})$, we have $\dim_A H = \infty$.*
- (3) *We have $\dim_A H < \infty$ precisely when A' is a II_1 factor.*
- (4) *We have additivity, $\dim_A(\oplus_i H_i) = \sum_i \dim_A H_i$.*
- (5) *We have $\dim_A(L^2(A)p) = \text{tr}(p)$, for any projection $p \in A$.*
- (6) *The coupling constant can take any value in $(0, \infty]$.*

PROOF. All these assertions are elementary, the idea being as follows:

- (1) This is something that we already know, coming from definitions.
- (2) This is something that comes from definitions too.
- (3) This comes from the general properties of the II_∞ factors, and their traces.
- (4) Again, this is clear from the definition of the coupling constant.
- (5) This follows by using $u(x) = x \otimes \xi$, with $\xi \in l^2(\mathbb{N})$ being of norm 1.
- (6) This follows by starting with (5), and then making direct sums, as in (4). \square

At a more advanced level now, in relation with projections and compressions, and getting towards the above-mentioned Murray-von Neumann approach, we have:

PROPOSITION 10.32. *We have the compression formula*

$$\dim_{pAp}(pH) = \frac{\dim_A H}{\text{tr}_A(p)}$$

valid for any projection $p \in A$.

PROOF. We can prove this result in two steps, as follows:

(1) In the case where $H = L^2(A)q$, with $q \in A$ being a projection satisfying $q \leq p$, we can use the following unitary, intertwining the left and right actions of pAp :

$$\begin{aligned} L^2(pAp) &\rightarrow pL^2(A)p \\ pxp\Omega &\rightarrow p(x\Omega)p \end{aligned}$$

Indeed, we obtain that the following algebras are unitarily equivalent:

$$\begin{aligned} pAp &\subset B(pL^2(A)q) \\ pAp &\subset B(L^2(pAp)q) \end{aligned}$$

Thus, by using the formula (5) in Proposition 10.31 we obtain, as desired:

$$\begin{aligned} \dim_{pAp}(pH) &= \operatorname{tr}_{pAp}(q) \\ &= \frac{\operatorname{tr}_A(q)}{\operatorname{tr}_A(p)} \\ &= \frac{\dim_A H}{\operatorname{tr}_A(p)} \end{aligned}$$

(2) In the general case now, where H is arbitrary, the result follows from what we proved above, and from the additivity property from Proposition 10.31 (4). \square

With all these properties established, we can now recover, as a theorem, the original definition of the coupling constant, due to Murray and von Neumann, as follows:

THEOREM 10.33. *Given a II_1 factor $A \subset B(H)$, with the commutant $A' \subset B(H)$ assumed to be finite, the corresponding coupling constant is finite, given by*

$$\dim_A H = \frac{\operatorname{tr}_A(P_{A'x})}{\operatorname{tr}_{A'}(P_{Ax})}$$

with the number on the right being independent of the choice on a nonzero vector $x \in H$. In the case where A' is infinite, the corresponding coupling constant is infinite.

PROOF. There are several things to be proved here, the idea being as follows:

(1) We know from Proposition 10.31 (3) that we have $\dim_A H < \infty$ precisely when the commutant $A' \subset B(H)$ is finite. Thus, we may assume that we are in this case.

(2) Assuming so, we have the following formula, valid for any projection $p \in A'$, which follows from the basic properties of the coupling constant, established above:

$$\dim_{Ap}(pH) = \operatorname{tr}_{A'}(p) \dim_A H$$

(3) Now with this formula in hand, the formula in the statement follows as well, once again by doing a number of standard amplification and compression manipulations. \square

As an illustration for all this, given an inclusion of ICC groups $\Lambda \subset \Gamma$, whose group algebras are both II_1 factors, we have the following formula:

$$\dim_{L(\Lambda)} L^2(\Gamma) = [\Gamma : \Lambda]$$

There are many other examples of explicit computations of the coupling constant, all leading into interesting mathematics. We will be back to this.

Given a II_1 factor A , let us discuss now the representations of type $A \subset B$, with B being another II_1 factor. This is a quite natural notion too, and perhaps even more natural than the representations $A \subset B(H)$ studied above, because we have previously decided that the II_1 factors B , and not the full operator algebras $B(H)$, are the correct infinite dimensional generalization of the usual matrix algebras $M_N(\mathbb{C})$.

This was for the philosophy, and one can of course agree or not with this. Or at least agree or not at the present point of the presentation, because once we will get into the structure of the subfactors $A \subset B$, which is something amazing, there is no way back.

Given an inclusion of II_1 factors $A \subset B$, a first question is that of defining its index, measuring how big is B compared to A . The first thought here goes into defining the index of $A \subset B$ as being a purely algebraic quantity, as follows:

$$N = \dim_A B$$

However, this is non-trivial, due to the fact that we are in the “continuous dimension” setting, and so our algebraic intuition, where indices are always integers, will not help us much. We will be back to this question later, with a technical solution to it.

In order to solve our index problem, a much better approach is by using the ambient operator algebra $B(H)$, or rather the ambient Hilbert space H , as follows:

THEOREM 10.34. *Given an inclusion of II_1 factors $A \subset B$, the number*

$$N = \frac{\dim_A H}{\dim_B H}$$

is independent of the ambient Hilbert space H , and is called index.

PROOF. The fact that the index of the subfactor $A \subset B$, as defined by the above formula, is indeed independent of the ambient Hilbert space H , comes from the various basic properties of the coupling constant, established in the above. \square

There are many examples of subfactors coming from groups, and every time we obtain the intuitive index. More suprisingly now, Jones proved in [43] that the index, when small, is in fact “quantized”, subject to the following unexpected restriction:

$$N \in \left\{ 4 \cos^2 \left(\frac{\pi}{n} \right) \mid n \geq 3 \right\} \cup [4, \infty]$$

This is in fact part of a series of non-trivial results about the subfactors, involving as well the Temperley-Lieb algebra [80], and many more. We will be back to this.

10e. Exercises

In relation with the Murray-von Neumann theory, we have:

EXERCISE 10.35. *Fully clarify the basic properties of the II_∞ factors, and the related construction of the coupling constant.*

This is something that we already discussed in the above, but with a few details missing, and the problem now is that of clarifying all this. You can either go through the discussion which was made above, and come up with the missing details, or do something alternative, based on the various historical comments given above.

CHAPTER 11

Reduction theory

11a. Preliminaries

Our purpose in this chapter is to discuss some key decomposition methods for the general von Neumann algebras, altogether called “reduction theory”. The whole subject is something quite technical, and unlike what we did in this book so far, we will not present complete proofs for everything, and leave some things not fully explained.

The reduction theory was developed by von Neumann in [89], long after developing the basics of the von Neumann algebra theory. The whole subject is fairly technical, but the main result, in its most general form, is quite easy to formulate, as follows:

THEOREM 11.1 (Reduction theory). *Given a von Neumann algebra $A \subset B(H)$, if we write its center $Z(A) \subset A$, which is a commutative von Neumann algebra, as*

$$Z(A) = L^\infty(X)$$

with X being a measured space, then the whole algebra decomposes as

$$A = \int_X A_x dx$$

with the fibers A_x being von Neumann algebras with trivial center, $Z(A_x) = \mathbb{C}$, or factors.

As a first comment, we have already seen an instance of the above decomposition result in chapter 5 above, and with full details, when talking about finite dimensional algebras, which decompose as direct sums of matrix algebras, as follows:

$$A = \bigoplus_x M_{n_x}(\mathbb{C})$$

More generally, it is possible to axiomatize a certain class of “type I algebras”, and then show that these algebras appear as direct integrals of matrix algebras:

$$A = \int_X M_{n_x}(\mathbb{C}) dx$$

Observe in particular that in the case where the decomposition is isotypic, $n_x = N$ for some $N \in \mathbb{N}$, we obtain the random matrix algebras studied in chapter 6:

$$A = M_N(L^\infty(X))$$

Beyond type I, however, things become quite complicated. Next in the hierarchy is the general “finite case”, where the algebra is assumed to have a trace:

$$\text{tr} : A \rightarrow \mathbb{C}$$

Here the existence of the trace simplifies a bit things, although these still remain fairly complicated, and actually adds to the final result, in the form of the supplementary formula, regarding its decomposition, the precise statement being as follows:

THEOREM 11.2 (Reduction theory, finite case). *Given a von Neumann algebra $A \subset B(H)$ coming with a trace $\text{tr} : A \rightarrow \mathbb{C}$, if we write its center $Z(A) \subset A$ as*

$$Z(A) = L^\infty(X)$$

with X being a measured space, then the whole algebra decomposes as

$$A = \int_X A_x dx$$

and the trace decomposes as well, as

$$\text{tr} = \int_X \text{tr}_x dx$$

with the fibers A_x being factors which are “finite”, in the sense that they have traces.

As for the most general case, where the algebra $A \subset B(H)$ is taken arbitrary, corresponding to Theorem 11.1 above, this is something even more complicated, with the only possible tools coming from advanced operator theory, and functional analysis.

Obviously, having all this material in the middle of a presumably serious book is not a good thing. The problem comes from the very nature of reduction theory, which is at the same time something fundamental, and so unavoidable in the context of a basic von Neumann algebra book, but which is technically complicated as well, and so not really fully explainable in the context of the same basic von Neumann algebra book.

This situation has been known to generations of mathematicians, starting with von Neumann himself, who finished and published his reduction theory paper [89] long after developing the basics of the operator algebra theory. The various books written afterwards, including Blackadar [19], Connes [26], Dixmier [29], Jones [50], Kadison-Ringrose [53], Sakai [75], Strătilă-Zsido [76] and Takesaki [77] did not arrange things, being either evasive, or way too technical, not to say unreadable, on this subject.

So, this is the situation, and the present book won't be an exception to the rule, the present chapter being something that the author is not really proud of. But bad things happen, and that's life, and take it easy. Finally, let me mention that I learned myself reduction theory from Dixmier's book [29], long ago, as a brave young PhD student, so in case you're unhappy with the material from the present chapter, or with some other

von Neumann algebra books and texts that you might find, on this topic, you can just do this: carefully read [29], and of course with all exercises done, by yourself.

Getting back now to the present book and chapter, our plan in what follows will be that of discussing all this more in detail, notably with a study of the examples. We will follow the standard hierarchy of the von Neumann algebras, which as follows:

(1) First we have the commutative von Neumann algebras. Here we already know that these are of the form $A = L^\infty(X)$, so the reduction theorem is proved for them, with the fibers being trivial, $A_x = \mathbb{C}$. However, there are many other things that can be said about such algebras, and we will explore here the subject, with a number of results.

(2) Next, we have the general type I algebras, which are direct integrals of matrix algebras $M_{n_x}(\mathbb{C})$, with the case $n_x = 1$ corresponding to commutativity, the case $n_x \in \mathbb{N}$ corresponding to the “type I finite case”, and with the general case being $n_x \in \mathbb{N} \cup \{\infty\}$. At the level of main examples, these come from finite groups and quantum groups.

(3) Then we have the type II algebras, where we can have both type I and type II factors in the decomposition. Of particular interest is the “finite” case, where the algebra is simply assumed to come with a trace, $tr : A \rightarrow \mathbb{C}$, and where the reduction theory result is Theorem 11.2 above, with the factors being matrix algebras $M_N(\mathbb{C})$, or II_1 factors.

(4) Finally, we have the general type III case, with no assumption on the algebra $A \subset B(H)$, corresponding to Theorem 11.1 above. Here the factors in the decomposition can be of type I, or of type II, or neither of type I or II, which are called by definition of type III. The interesting questions here regard the structure of the type III factors.

In order to get started, let us look at the commutative von Neumann algebras. Here we have the following result, that we basically know from chapter 5 above:

THEOREM 11.3. *The commutative von Neumann algebras are the algebras of type*

$$A = L^\infty(X)$$

with X being a measured space. Thus, we formally have for them the formula

$$A = \int_X A_x dx$$

with the fibers A_x being trivial in this case, $A_x = \mathbb{C}$, for any $x \in X$.

PROOF. We have several assertions to be proved, the idea being as follows:

(1) In one sense, we must prove that given a measured space X , we can realize the commutative algebra $A = L^\infty(X)$ as a von Neumann algebra, on a certain Hilbert space H . But this is something that can be done via multiplicity operators, as follows:

$$L^\infty(X) \subset B(L^2(X))$$

(2) In the other sense, given a commutative von Neumann algebra $A \subset B(H)$, we must construct a certain measured space X , and an identification $A = L^\infty(X)$. But this can be done by writing our von Neumann algebra as follows:

$$A = \langle T_i \rangle$$

Indeed, no matter what particular family of generators $\{T_i\}$ we choose for our algebra A , these generators T_i will be commuting normal operators. Thus the spectral theorem for such families of operators, from chapter 3 above, applies and gives the result.

(3) In fact, by using the theory of projections from chapters 9-10 above, we can write our commutative von Neumann algebra $A \subset B(H)$ in singly generated form:

$$A = \langle T \rangle$$

But this simplifies the situation, because the spectral theorem for normal operators, from chapter 3 above, applies to our generator T , and gives the result.

(4) Finally, the last assertion, regarding the validity of the reduction theory result in this case, is something rather trivial, and of course without much practical interest. \square

Moving forward, the above result is not the end of the story with the commutative von Neumann algebras, because we still have to understand how a given such algebra $A = L^\infty(X)$, or rather the weak topology isomorphism class of such an algebra, can be represented as an operator algebra, over the various Hilbert spaces H :

$$L^\infty(X) \subset B(H)$$

But this can be again solved by writing our algebra as $A = \langle T \rangle$, and then applying the spectral theorem for normal operators, with the conclusion that the commutative von Neumann algebras are, up to spatial isomorphism, the algebras of the following form, with X being a measured space, and with all this being up to a multiplicity:

$$L^\infty(X) \subset B(L^2(X))$$

With these results in hand, we are now in position of better understanding the idea behind von Neumann's reduction theory. Indeed, given an arbitrary von Neumann algebra $A \subset B(H)$, the idea is to consider its center, and write it as follows:

$$Z(A) = L^\infty(X) \subset B(H)$$

The point is then that everything will decompose over the measured space X , and in particular, the whole algebra A itself will decompose as a direct integral of fibers:

$$A = \int_X A_x dx$$

As already mentioned, we will only partly explain this in what follows, and by insisting on examples. Also, we will do this slowly, following the type I, II, III hierarchy.

Before getting into this, however, let us explore some further perspectives opened by the above results, which are quite sharp, regarding the commutative algebras. Given a von Neumann algebra $A \subset B(H)$, looking at the center $Z(A) = A \cap A'$ is not the only possible way of getting to commutative subalgebras, and we have as well:

DEFINITION 11.4. *Given a von Neumann algebra $A \subset B(H)$, a commutative subalgebra $B \subset A$ which is maximal, in the sense that there is no bigger commutative algebra*

$$B \subset B' \subset A$$

is called maximal commutative subalgebra (MCSA).

We should mention that it is quite common in the literature to call the commutative subalgebras “abelian”, and so the maximal commutative ones, MASA. However, the term “abelian” is a bit unfortunate, with respect to our philosophy and conventions in this book, and we will rather use instead the above terminology.

It is possible to say many interesting things about the MCSA, and skipping some details here, if we want to further build on this notion, we are led to:

DEFINITION 11.5. *Given a von Neumann algebra A coming with a trace $tr : A \rightarrow \mathbb{C}$, assume that we have a pair of maximal commutative subalgebras*

$$B, C \subset A$$

satisfying the following orthogonality condition, with respect to the trace:

$$(B \ominus \mathbb{C}1) \perp (C \ominus \mathbb{C}1)$$

We say then that B, C are orthogonal maximal commutative subalgebras.

Here the scalar product is by definition $\langle b, c \rangle = tr(bc^*)$, and by taking into account the multiples of the identity, the orthogonality condition reformulates as follows:

$$tr(bc) = tr(b)tr(c)$$

As a “toy example”, we can try and see what happens for the simplest factor that we know, namely the matrix algebra $M_N(\mathbb{C})$, endowed with its usual matrix trace. And in this context, we have the following surprising result of Popa [69]:

THEOREM 11.6. *Up to a conjugation by a unitary, the pairs of orthogonal maximal commutative subalgebras in the simplest factor, namely $M_N(\mathbb{C})$, are as follows,*

$$A = \Delta$$

$$B = H\Delta H^*$$

with $\Delta \subset M_N(\mathbb{C})$ being the diagonal matrices, and with $H \in M_N(\mathbb{C})$ being Hadamard, in the sense that $|H_{ij}| = 1$ for any i, j , and the rows of H are pairwise orthogonal.

PROOF. Any maximal commutative subalgebra in $M_N(\mathbb{C})$ being conjugated to Δ , we can assume, up to conjugation by a unitary, that we have, with $U \in U_N$:

$$A = \Delta$$

$$B = U\Delta U^*$$

Now observe that given two diagonal matrices $D, E \in \Delta$, we have:

$$\begin{aligned} \text{tr}(D \cdot UEU^*) &= \frac{1}{N} \sum_i (DEU^*U)_i \\ &= \frac{1}{N} \sum_{ij} D_{ii} U_{ij} E_{jj} \bar{U}_{ij} \\ &= \frac{1}{N} \sum_{ij} D_{ii} E_{jj} |U_{ij}|^2 \end{aligned}$$

Thus, the orthogonality condition $A \perp B$ reformulates as follows:

$$\frac{1}{N} \sum_{ij} D_{ii} E_{jj} |U_{ij}|^2 = \frac{1}{N^2} \sum_{ij} D_{ii} E_{jj}$$

Thus the rescaled matrix $H = \sqrt{N}U$ must satisfy the following condition:

$$|H_{ij}| = 1$$

Thus, we are led to the conclusion in the statement. \square

The Hadamard matrices appearing in Theorem 11.6 above are well-known objects, appearing in several branches of combinatorics, and quantum physics. The basic example of such a matrix is the Fourier matrix F_N , constructed as follows:

PROPOSITION 11.7. *Given $N \in \mathbb{N}$, the following matrix, with $w = e^{2\pi i/N}$,*

$$F_N = \begin{pmatrix} 1 & 1 & 1 & \dots & 1 \\ 1 & w & w^2 & \dots & w^{N-1} \\ 1 & w^2 & w^4 & \dots & w^{2(N-1)} \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 1 & w^{N-1} & w^{2(N-1)} & \dots & w^{(N-1)^2} \end{pmatrix}$$

is a complex Hadamard matrix, called Fourier matrix.

PROOF. By using the standard fact that the averages of complex numbers correspond to barycenters, we conclude that the scalar products between the rows of F_N are:

$$\begin{aligned} \langle R_a, R_b \rangle &= \sum_j w^{aj} w^{-bj} \\ &= \sum_j w^{(a-b)j} \\ &= N \delta_{ab} \end{aligned}$$

Thus F_N is indeed a complex Hadamard matrix, as claimed. \square

There is a relation here with the basic Fourier transform theory discussed in chapter 7 above, because F_N is precisely the matrix of the Fourier transform over \mathbb{Z}_N :

$$F_N : L(\mathbb{Z}_N) \simeq L^\infty(\mathbb{Z}_N)$$

It is possible to further build on this, by associating to any finite abelian group, written as $G = \mathbb{Z}_{N_1} \times \dots \times \mathbb{Z}_{N_k}$, its Fourier matrix F_G , which is Hadamard, as follows:

$$F_G = F_{N_1} \otimes \dots \otimes F_{N_k}$$

To be more precise, we have the following result, generalizing Proposition 11.7:

THEOREM 11.8. *Given a finite abelian group G , with dual group $\widehat{G} = \{\chi : G \rightarrow \mathbb{T}\}$, consider the Fourier coupling $\mathcal{F}_G : G \times \widehat{G} \rightarrow \mathbb{T}$, given by $(i, \chi) \rightarrow \chi(i)$.*

- (1) *Via the standard isomorphism $G \simeq \widehat{\widehat{G}}$, this Fourier coupling can be regarded as a square matrix, $F_G \in M_G(\mathbb{T})$, which is a complex Hadamard matrix.*
- (2) *In the case of the cyclic group $G = \mathbb{Z}_N$ we obtain in this way, via the standard identification $\mathbb{Z}_N = \{1, \dots, N\}$, the Fourier matrix F_N .*
- (3) *In general, when using a decomposition $G = \mathbb{Z}_{N_1} \times \dots \times \mathbb{Z}_{N_k}$, the corresponding Fourier matrix is given by $F_G = F_{N_1} \otimes \dots \otimes F_{N_k}$.*

PROOF. This follows indeed from some basic facts from group theory:

(1) With the identification $G \simeq \widehat{\widehat{G}}$ made our matrix is given by $(F_G)_{i\chi} = \chi(i)$, and the scalar products between the rows are computed as follows:

$$\begin{aligned} \langle R_i, R_j \rangle &= \sum_\chi \chi(i) \overline{\chi(j)} \\ &= \sum_\chi \chi(i - j) \\ &= |G| \cdot \delta_{ij} \end{aligned}$$

Thus, we obtain indeed a complex Hadamard matrix.

(2) This follows from the well-known and elementary fact that, via the identifications $\mathbb{Z}_N = \widehat{\mathbb{Z}_N} = \{1, \dots, N\}$, the Fourier coupling here is as follows, with $w = e^{2\pi i/N}$:

$$(i, j) \rightarrow w^{ij}$$

(3) We use here the following well-known formula, for the duals of products:

$$\widehat{H \times K} = \widehat{H} \times \widehat{K}$$

At the level of the corresponding Fourier couplings, we obtain from this:

$$F_{H \times K} = F_H \otimes F_K$$

Now by decomposing G into cyclic groups, as in the statement, and by using (2) for the cyclic components, we obtain the formula in the statement. \square

Summarizing, we have some interesting connections with finite group theory, and with the associated Fourier matrices. However, there are as well many exotic examples of Hadamard matrices, nor necessarily coming from finite groups, as in Theorem 11.8, and all this is quite of interest for us, in connection with Theorem 11.6 above.

To be more precise, there are some wild, unexplored areas in the foundations of the von Neumann algebra theory, coming from the Hadamard matrices and their combinatorics, via Theorem 11.6, and all this looks very exciting. We will be back to this in chapters 13-16 below, with more results on the subject, when talking about subfactors.

11b. Type I algebras

With the above discussion made, about the commutative algebras, let us go back now to the reduction theory program, as outlined in the beginning of this chapter. Our goal will be that of writing any von Neumann algebra $A \subset B(H)$ as a direct sum of factors, by decomposing everything with respect to the center, $Z(A) = L^\infty(X)$.

Let us first talk about factors. We can now make some upgrades to our terminology and notations regarding these factors, as follows:

DEFINITION 11.9. *The von Neumann algebras having trivial center, also called factors, can be divided into several types, as follows:*

- (1) *The matrix algebra $M_N(\mathbb{C})$ is of type I_N .*
- (2) *The operator algebra $B(H)$ is of type I_∞ .*
- (3) *The II_1 factors are those which are infinite dimensional, and have a trace.*
- (4) *The tensor products $A \otimes B(H)$, with A being a II_1 factor, are of type II_∞ .*
- (5) *As for the factors left, these are called of type III.*

It is possible to be more abstract here, but in practice, this is how these factors are best remembered. We will be back to this, with some abstract results as well.

As already mentioned, on several occasions, we will present the reduction theory of von Neumann gradually, by following the general type I, II, III hierarchy for the von Neumann algebras, and its subdivisions, coming from the above classification of the factors.

Let us first discuss the type I case. Here as starting point we have the following result, which is something that we know well, from chapter 5 above:

THEOREM 11.10. *The finite dimensional von Neumann algebras $A \subset B(H)$ are exactly the direct sums of matrix algebras,*

$$A = M_{r_1}(\mathbb{C}) \oplus \dots \oplus M_{r_k}(\mathbb{C})$$

with the summands coming by decomposing the unit into central minimal projections:

$$1 = P_1 + \dots + P_k$$

Thus, the general reduction theory formula, namely

$$A = \int_X A_x dx$$

holds for them, with the measured space X , coming via the formula $Z(A) = L^\infty(X)$, being in this case a finite space, $X = \{1, \dots, k\}$, and with the fibers being matrix algebras.

PROOF. This is standard, as in the case $A \subset M_N(\mathbb{C})$. Consider the center of A , which is a finite dimensional commutative von Neumann algebra, of the following form:

$$Z(A) = \mathbb{C}^k$$

Now let P_i be the Dirac mass at $i \in \{1, \dots, k\}$. Then $P_i \in B(H)$ is an orthogonal projection, and these projections form a partition of unity, as follows:

$$1 = P_1 + \dots + P_k$$

With $A_i = P_i A P_i$, which is a non-unital $*$ -subalgebra of A , it is elementary to check that we have a non-unital $*$ -algebra sum decomposition, as follows:

$$A = A_1 \oplus \dots \oplus A_k$$

On the other hand, it follows from the minimality of each of the projections $P_i \in Z(A)$ that we have unital $*$ -algebra isomorphisms, as follows:

$$A_i \simeq M_{r_i}(\mathbb{C})$$

Thus, we are led to the conclusion in the statement. For full details on all this, we refer to the standard material from chapter 5 above. \square

It is possible to further build on the above result, in several directions, either by allowing the factors in the decomposition to be type I_∞ factors as well, $A_x \simeq B(H)$, or by allowing the center to be an infinite measured space, $|X| = \infty$, or by allowing both.

The first possible generalization, with X being kept finite, and with the fibers being allowed to be arbitrary type I factors, is not very interesting. The second possible generalization, however, is something quite interesting, and we have here:

THEOREM 11.11 (Reduction theory, type I finite case). *Given a von Neumann algebra $A \subset B(H)$ which is of discrete type, and has a trace $tr : A \rightarrow \mathbb{C}$, we can write*

$$A = \int_X A_x dx$$

with X coming via $Z(A) = L^\infty(X)$, and the trace decomposes as well, as

$$tr = \int_X tr_x dx$$

with the fibers A_x being usual matrix algebras, $A_x = M_{n_x}(\mathbb{C})$, with $n_x \in \mathbb{N}$.

As a first observation, this statement generalizes both what we know about the commutative algebras, and the finite dimensional ones. However, having these two things jointly generalized is something quite technical, that we will not explain here in detail.

The idea is of course first that of axiomatizing what “discrete” should mean in the above, say by looking at the finiteness properties of the projections $p \in A$, and then, once the statement properly formulated, to prove it by jointly generalizing what we know about the commutative algebras, and the finite dimensional ones.

Moving ahead now, with this done, let us lift now the assumption that the factors in the decomposition are of type I_N , with $N < \infty$. We are led in this way to the general reduction theory result in the type I case, whose statement is as follows:

THEOREM 11.12 (Reduction theory, type I case). *Given a von Neumann algebra $A \subset B(H)$ which is of type I, in the sense that it is of a suitable discrete type, we can write*

$$A = \int_X A_x dx$$

with X coming via $Z(A) = L^\infty(X)$, and with the fibers A_x being type I factors, meaning $A_x \simeq B(H_x)$, with each H_x being either finite dimensional, or separable.

As before with Theorem 11.11, we will not attempt to fully explain all this here. As a comment, however, this can only follow from Theorem 11.11 applied to the “finite” part of the algebra, obtained by removing the infinite part, and after proving that this infinite part is something of type $L^\infty(Y) \otimes B(H)$, with $Y \subset X$, and with H being separable.

All the above was quite abstract, and as something more concrete now, illustrating for all this, let us discuss the reduction theory for the group von Neumann algebras $L(\Gamma)$, in the finite group case, $|\Gamma| < \infty$, and their quantum group generalizations.

For this purpose, it is technically convenient to change a bit our terminology and notations, making them more in tune with the quantum group formalism from chapter 7. First, we will denote our finite group Γ , which is at the same time discrete and compact, by F , and we will think of it as being the dual of a finite quantum group $G = \widehat{F}$. Also, since in the finite group case everything is automatically norm or weakly closed, we will use the more familiar notation $C^*(F)$ for the associated von Neumann algebra $L(F)$. With these conventions, we have the following result, which is standard:

THEOREM 11.13. *Given a finite group F , the center of the associated von Neumann algebra is isomorphic to the algebra of central functions on F ,*

$$Z(C^*(F)) \simeq C(F)_{\text{central}}$$

and the reduction theory applied to this von Neumann algebra, which is a formula of type

$$C^*(F) \simeq \bigoplus_{r \in X} M_{n_r}(\mathbb{C})$$

appears by dualizing the Peter-Weyl decomposition of the usual function algebra

$$C(F) \simeq \bigoplus_{r \in \text{Irr}(F)} M_{\dim(r)}(\mathbb{C})$$

via the standard identification between representations r and their characters χ_r .

PROOF. In what concerns the first assertion, regarding the center, this is something that we already know, from chapter 10, coming from our study there of the general group algebras $L(\Gamma)$, with Γ being a discrete group. To be more precise, in the case where $\Gamma = F$ is a finite group, the computation there gives the following formula for the center:

$$Z(C^*(F)) = \left\{ \sum_g \lambda_g g \mid \lambda_{gh} = \lambda_{hg}, \forall h \in F \right\}''$$

Now since on the right we have central functions on our group, $\lambda \in C(F)_{\text{central}}$, we obtain the isomorphism in the statement, namely:

$$Z(C^*(F)) \simeq C(F)_{\text{central}}$$

Regarding now the second assertion, let us first recall that the Peter-Weyl theory applied to the finite group F gives a direct sum decomposition as follows, which is technically an isomorphism of linear spaces, which is in addition a *-coalgebra isomorphism:

$$C(F) \simeq \bigoplus_{r \in \text{Irr}(F)} M_{\dim(r)}(\mathbb{C})$$

Thus by dualizing, which is a standard functional analysis procedure, to be explained more in detail below, in a more general setting, we obtain a direct sum decomposition of the group algebra, as follows, which is this time a $*$ -algebra isomorphism:

$$C^*(F) \simeq \bigoplus_{r \in Irr(F)} M_{\dim(r)}(\mathbb{C})$$

Our claim now, which will finish the proof, is that this is exactly what comes out from von Neumann's reduction theory, applied to the von Neumann algebra $L(F) = C^*(F)$. Indeed, by using the standard identification between representations r and their characters χ_r , which are central functions on F , the center computation that we did above reads:

$$Z(C^*(F)) \simeq L^\infty(Irr(F))$$

We conclude that von Neumann's reduction theory, applied to the von Neumann algebra $L(F) = C^*(F)$, gives a $*$ -algebra isomorphism of the following type:

$$C^*(F) \simeq \bigoplus_{r \in Irr(F)} M_{n_r}(\mathbb{C})$$

But a careful examination of the fibers appearing in this decomposition, based on their very definition, shows that these are precisely the above matrix blocks coming from Peter-Weyl. That is, we have $n_r = \dim(r)$ for any $r \in Irr(F)$, and we are done. \square

Our next goal will be that of extending the above result to the finite quantum group setting. For this purpose, we will not really need the general compact quantum group formalism from chapter 7 above, and it is convenient to start with:

DEFINITION 11.14. *A finite dimensional Hopf algebra is a finite dimensional C^* -algebra, with a comultiplication, counit and antipode maps, satisfying the conditions*

$$\begin{aligned} (\Delta \otimes id)\Delta &= (id \otimes \Delta)\Delta \\ (\varepsilon \otimes id)\Delta &= (id \otimes \varepsilon)\Delta = id \\ m(S \otimes id)\Delta &= m(id \otimes S)\Delta = \varepsilon(\cdot)1 \end{aligned}$$

along with the extra condition $S^2 = id$. Given such an algebra we write

$$A = C(G) = C^*(F)$$

and call G, F finite quantum groups, dual to each other.

In this definition everything is standard, except for the last axiom, $S^2 = id$. This axiom corresponds to the fact that, in the corresponding quantum group, we have:

$$(g^{-1})^{-1} = g$$

It is possible to prove that this condition is automatic, in the present C^* -algebra setting. However, this is something non-trivial, and since all this is just a preliminary discussion, not needed later, we have opted for including $S^2 = id$ in our axioms.

We say that an algebra A as above is cocommutative if $\Sigma\Delta = \Delta$, where $\Sigma(a \otimes b) = b \otimes a$ is the flip. With this convention made, we have the following result, which summarizes the basic theory of finite quantum groups, and justifies the terminology and axioms:

THEOREM 11.15. *The following happen:*

- (1) *If G is a finite group then $C(G)$ is a commutative Hopf algebra, with*

$$\Delta(\varphi) = (g, h) \rightarrow \varphi(gh)$$

$$\varepsilon(\varphi) = \varphi(1)$$

$$S(\varphi) = g \rightarrow \varphi(g^{-1})$$

as structural maps. Any commutative Hopf algebra is of this form.

- (2) *If F is a finite group then $C^*(F)$ is a cocommutative Hopf algebra, with*

$$\Delta(g) = g \otimes g$$

$$\varepsilon(g) = 1$$

$$S(g) = g^{-1}$$

as structural maps. Any cocommutative Hopf algebra is of this form.

- (3) *If G, F are finite abelian groups, dual to each other via Pontrjagin duality, then we have an identification of Hopf algebras*

$$C(G) = C^*(F)$$

coming via a Fourier transform type operation.

PROOF. These results are all elementary, the idea being as follows:

(1) The fact that Δ, ε, S satisfy the axioms is clear from definitions, and the converse follows from the Gelfand theorem, by working out the details, regarding Δ, ε, S .

(2) Once again, the fact that Δ, ε, S satisfy the axioms is clear from definitions, with the remark that the use of the opposite multiplication $(a, b) \rightarrow a \cdot b$ is really needed here, in order for the antipode S to be an algebra morphism:

$$S(gh) = (gh)^{-1} = g^{-1} \cdot h^{-1} = S(g) \cdot S(h)$$

For the converse, we use a trick. Let A be an arbitrary finite dimensional Hopf algebra, as in Definition 11.15, and consider its comultiplication, counit, multiplication, unit and antipode maps. The transposes of these maps are then linear maps as follows:

$$\Delta^t : A^* \otimes A^* \rightarrow A^*$$

$$\varepsilon^t : \mathbb{C} \rightarrow A^*$$

$$m^t : A^* \rightarrow A^* \otimes A^*$$

$$u^t : A^* \rightarrow \mathbb{C}$$

$$S^t : A^* \rightarrow A^*$$

It is routine to check that these maps make A^* into a Hopf algebra. Now assuming that A is cocommutative, it follows that A^* is commutative, so by (1) we obtain $A^* = C(G)$ for a certain finite group G , which in turn gives $A = C^*(G)$, as desired.

(3) This follows from the discussion in the proof of (2) above. \square

There are many other things that can be said about the finite dimensional Hopf algebras, and in what follows we will be particularly interested in the notion of corepresentation. These corepresentations can be introduced as follows:

DEFINITION 11.16. *A unitary corepresentation of a finite dimensional Hopf algebra A is a unitary matrix $u \in M_r(A)$ satisfying the following conditions:*

$$\begin{aligned}\Delta(u_{ij}) &= \sum_k u_{ik} \otimes u_{kj} \\ \varepsilon(u_{ij}) &= \delta_{ij} \\ S(u_{ij}) &= u_{ji}^*\end{aligned}$$

We say that u is irreducible, and we write $u \in \text{Irr}(A)$, when it has no nontrivial intertwiners, in the sense that $Tu = uT$ with $T \in M_n(\mathbb{C})$ implies $T \in \mathbb{C}1$.

Observe the similarity with the notions introduced in chapter 7 above, for the Woronowicz algebras. In fact, by using left regular representations we can see that any finite dimensional Hopf algebra, in the sense of Definition 11.14, is a Woronowicz algebra, in the sense of chapter 7 above. Thus, we can freely use here the results established in chapter 7 above, and in particular, we can use the Peter-Weyl theory developed there.

In relation now with our von Neumann algebra questions, we have the following result, coming from this Peter-Weyl theory, which generalizes Theorem 11.13 above:

THEOREM 11.17. *Given a finite quantum group F , the center of the associated von Neumann algebra is isomorphic to the algebra of central functions on F ,*

$$Z(C^*(F)) \simeq C(F)_{\text{central}}$$

and the reduction theory applied to this von Neumann algebra, which is a formula of type

$$C^*(F) \simeq \bigoplus_{u \in X} M_{n_u}(\mathbb{C})$$

appears by dualizing the Peter-Weyl decomposition of the usual function algebra

$$C(F) \simeq \bigoplus_{u \in \text{Irr}(F)} M_{\dim(u)}(\mathbb{C})$$

via the standard identification between representations u and their characters χ_u .

PROOF. The proof here is nearly identical to the proof of Theorem 11.13. To be more precise, with the more familiar notation $A = C^*(F)$, the proof goes as follows:

(1) In what concerns the first assertion, regarding the center, we recall from Woronowicz [98] that $A_{central}$ is by definition the subalgebra of A appearing as follows:

$$A_{central} = \left\{ a \in A \mid \Delta a = a \right\}$$

But this shows, first by dualizing, and then by doing some computations similar to those that we did in chapter 10 above, when computing the centers of the usual group von Neumann algebras, that we have an isomorphism as in the statement, namely:

$$Z(A) \simeq (A^*)_{central}$$

(2) Regarding now the second assertion, we recall that the Peter-Weyl theory applied to Hopf algebra A^* gives a direct sum decomposition as follows, which is technically an isomorphism of linear spaces, which is in addition a *-coalgebra isomorphism:

$$A^* \simeq \bigoplus_{u \in Irr(A^*)} M_{\dim(u)}(\mathbb{C})$$

Thus by dualizing, we obtain a direct sum decomposition of the group algebra, as follows, which is this time a *-algebra isomorphism:

$$A \simeq \bigoplus_{u \in Irr(A^*)} M_{\dim(u)}(\mathbb{C})$$

(3) Our claim now, which will finish the proof, is that this is exactly what comes out from von Neumann's reduction theory, applied to the algebra A . Indeed, by using the standard identification between corepresentations u of A^* and their characters χ_u , which belong to the algebra $(A^*)_{central}$, the center computation that we did above reads:

$$Z(A) \simeq L^\infty(Irr(A^*))$$

We conclude that von Neumann's reduction theory, applied to the von Neumann algebra A , gives a *-algebra isomorphism of the following type:

$$A \simeq \bigoplus_{u \in Irr(A^*)} M_{n_u}(\mathbb{C})$$

But a careful examination of the fibers shows that these are precisely the matrix blocks coming from Peter-Weyl. That is, $n_u = \dim(u)$ for any $u \in Irr(A^*)$, and we are done. \square

As a final comment here, the most interesting type I algebras are probably those having an isotypic decomposition, and so which can be written as follows:

$$A = M_N(L^\infty(X))$$

But these are precisely the random matrix algebras, that we investigated in great detail in chapter 6 above, and we refer to that chapter for more about them.

11c. Type II algebras

Let us discuss now the type II case, where the truly interesting problems are. The central result here, that we already formulated in the beginning of this chapter, is:

THEOREM 11.18 (Reduction theory, finite case). *Given a von Neumann algebra $A \subset B(H)$ coming with a trace $tr : A \rightarrow \mathbb{C}$, if we write its center $Z(A) \subset A$ as*

$$Z(A) = L^\infty(X)$$

with X being a measured space, then the whole algebra decomposes as

$$A = \int_X A_x dx$$

and the trace decomposes as well, as

$$tr = \int_X tr_x dx$$

with the fibers A_x being either factors of type I_N , with $N < \infty$, or of type II_1 .

Regarding the proof, this is something quite technical, generalizing what we know, or rather what we don't, about the type I finite case, which is substantially easier. We refer here the literature, and with the comment that we will see soon examples of all this.

As before in the type I case, it is possible to add a bit of infinity in the above, and we have the following result, which is a bit more technical, but more general too:

THEOREM 11.19 (Reduction theory, type II case). *Given a von Neumann algebra $A \subset B(H)$ which is of type II, in a suitable sense, if we write its center $Z(A) \subset A$ as*

$$Z(A) = L^\infty(X)$$

with X being a measured space, then the whole algebra decomposes as

$$A = \int_X A_x dx$$

with the fibers A_x being von Neumann factors of type I or II.

As before, for the proof of this, we basically refer to the literature. Let us mention however that this can only follow from Theorem 11.18, after some suitable manipulations, a bit as Theorem 11.12 was following from Theorem 11.11, in the type I case.

As before with what happened in type I, the above reduction theory results in type II are particularly interesting in the case of the von Neumann algebras of the discrete groups, $A = L(\Gamma)$, and their generalizations. Here the decomposition into fibers, as in Theorem 11.18, is a quite delicate business, more complicated than in type I.

In order to discuss these questions, let us recall from the above that the center of an arbitrary group von Neumann algebra $A = L(\Gamma)$ consists, up to some standard identifications, of the functions which are constant on the finite conjugacy classes. This suggests the following definition, which is something well-known in group theory:

DEFINITION 11.20. *A discrete group F is said to have the FC property if all its conjugacy classes are finite. In other words, for any $g \in F$, we must have:*

$$\left| \left\{ hgh^{-1} \mid h \in F \right\} \right| < \infty$$

If this finite conjugacy property is satisfied, we also say that F is a FC group.

As basic examples of FC groups, we have the finite groups, the abelian groups, and the products of such groups. Besides being stable under taking products, the class of FC groups is stable under a number of other basic operations, such as taking subgroups, or quotients. There are also a number of abstract equivalent formulations of the FC property, in more technical group theory terms, and we refer here to the literature.

In connection with our reduction theory questions, we have:

THEOREM 11.21. *Given a group F having the FC property, the center of the associated von Neumann algebra is isomorphic to the algebra of central functions on F ,*

$$Z(L(F)) \simeq C(F)_{\text{central}}$$

and the reduction theory applied to this von Neumann algebra, which is a formula of type

$$L(F) \simeq \int_{r \in X} A_r$$

appears in relation with the representation theory of F .

PROOF. In what concerns the first assertion, regarding the center, this is something that we already know, from chapter 10, coming from our study there of the general group algebras $L(\Gamma)$, with Γ being a discrete group. To be more precise, in the case where $\Gamma = F$ has the FC property, the computation there gives the following formula for the center:

$$Z(L(F)) = \left\{ \sum_g \lambda_g g \mid \lambda_{gh} = \lambda_{hg}, \forall h \in F \right\}''$$

Now since on the right we have central functions on our group, $\lambda \in C(F)_{\text{central}}$, we obtain the isomorphism in the statement, namely:

$$Z(L(F)) \simeq C(F)_{\text{central}}$$

Regarding now the second assertion, this is something more tricky, as follows:

(1) In the finite group case, we recall from Theorem 11.13 above that, by using the standard identification between representations r and their characters χ_r , which are central functions on F , the center computation that we did above reads:

$$Z(L(F)) \simeq L^\infty(\text{Irr}(F))$$

In order to discuss now the reduction theory for $L(F)$, we recall that the Peter-Weyl theory applied to F gives a direct sum decomposition as follows, which is technically an isomorphism of linear spaces, which is in addition a $*$ -coalgebra isomorphism:

$$L^\infty(F) \simeq \bigoplus_{r \in \text{Irr}(F)} M_{\dim(r)}(\mathbb{C})$$

Thus by dualizing, we obtain a direct sum decomposition of the group von Neumann algebra as follows, which is this time a $*$ -algebra isomorphism:

$$L(F) \simeq \bigoplus_{r \in \text{Irr}(F)} M_{\dim(r)}(\mathbb{C})$$

But this is exactly what comes out from von Neumann's reduction theory, applied to the algebra $L(F)$, and so we are fully done with the finite group case.

(2) As a second key particular case, let us discuss now the case where F is abelian. In the simplest infinite group case, where our group is $F = \mathbb{Z}$, the group algebra is:

$$L(\mathbb{Z}) \simeq L^\infty(\mathbb{T})$$

More generally, for the abelian groups $F = \mathbb{Z}^N$, which are those which are finitely generated and without torsion, we obtain the algebras of functions on various tori:

$$L(\mathbb{Z}^N) \simeq L^\infty(\mathbb{T}^N)$$

In general now, assuming that F is finitely generated and abelian, here we know from Pontrjagin duality that we have an isomorphism as follows:

$$L(F) \simeq L^\infty(\widehat{F})$$

More explicitly now, let us write our finitely generated abelian group F as a product of cyclic groups, possibly taken infinite, as follows:

$$F = \mathbb{Z}^N \times \left(\prod_i \mathbb{Z}_{n_i} \right)$$

The Pontrjagin dual of F is then the following compact abelian group:

$$F = \mathbb{T}^N \times \left(\prod_i \mathbb{Z}_{n_i} \right)$$

Thus, things are very explicit here, and we are done with the abelian case too.

(3) In the general case now, where our discrete group F is only assumed to have the FC property, the reduction theory for the corresponding von Neumann algebra $L(F)$ appears somewhat as a mixture of what happens for the finite and for the abelian groups, discussed in (1) and (2) above. For more on all this, we refer to the literature. \square

Regarding the corresponding problems for the discrete quantum groups, these are not solved yet. In fact, the knowledge here stops at a very basic level, with the correct analogue of the ICC property, leading to the factoriality of $L(\Gamma)$, not being known yet.

Moving ahead from these difficulties, let us go back now to the usual group von Neumann algebras $L(\Gamma)$, and discuss what happens in general. Once again inspired by the basic computation that we have, namely that of the center of an arbitrary group algebra $L(\Gamma)$, let us formulate the following purely group-theoretical definition:

DEFINITION 11.22. *Given a discrete group Γ , its FC subgroup $F \subset \Gamma$ is the subgroup*

$$F = \left\{ g \in \Gamma \mid \left| \left\{ hgh^{-1} \mid h \in \Gamma \right\} \right| < \infty \right\}$$

consisting of the elements in the finite conjugacy classes of Γ .

Here the fact that F is indeed a subgroup is clear from definitions, with the fact that F is stable under multiplication coming from the following trivial observation:

$$h(gk)h^{-1} = hgh^{-1} \cdot hkh^{-1}$$

Observe that Γ has the FC property, in the sense of Definition 11.21, precisely when the inclusion $F \subset \Gamma$ is an equality. As before with the FC groups, there are many known things about the FC subgroups $F \subset \Gamma$, and we refer here to the group theory literature.

In connection with our reduction theory questions, we have:

THEOREM 11.23. *Given a discrete group Γ , the center of the associated von Neumann algebra is isomorphic to the algebra of central functions on its FC subgroup $F \subset \Gamma$,*

$$Z(L(\Gamma)) \simeq C(F)_{\text{central}}$$

and the reduction theory applied to this von Neumann algebra, which is a formula of type

$$L(\Gamma) \simeq \int_{r \in X} A_r$$

appears in relation with the representation theory of Γ , and of its FC subgroup $F \subset \Gamma$.

PROOF. In what concerns the first assertion, regarding the center, this is something that we know from chapter 10, coming from our study there of the general group algebras $L(\Gamma)$, with Γ being a discrete group. To be more precise, we know from there that:

$$Z(L(\Gamma)) = \left\{ \sum_g \lambda_g g \mid \lambda_{gh} = \lambda_{hg}, \forall h \in F \right\}''$$

Now since on the right we have central functions on the FC subgroup, $\lambda \in C(F)_{\text{central}}$, we obtain the isomorphism in the statement, namely:

$$Z(L(\Gamma)) \simeq C(F)_{\text{central}}$$

Regarding now the second assertion, this is something more tricky, and we refer here to the relevant group theory and operator algebra literature. \square

Summarizing, things are quite well understood for the von Neumann algebras of discrete groups $L(\Gamma)$, with everything ultimately reducing to representation theory.

The corresponding problems for discrete quantum groups are not solved yet. In fact, the knowledge here stops at a very basic level, with the correct analogue of the ICC property, guaranteeing the factoriality of $L(\Gamma)$, not being known yet.

Finally, in connection with the quantum groups, let us mention that there are some interesting von Neumann algebra questions in connection with the corresponding matrix models. To be more precise, one interesting method for the study of the closed subgroups $G \subset U_N^+$ consists in modelling the coordinates $u_{ij} \in C(G)$ by concrete variables $U_{ij} \in B$. Indeed, assuming that the model is faithful in some suitable sense, that the algebra B is something quite familiar, and that the variables U_{ij} are not too complicated, all the questions about G would correspond in this way to routine questions inside B .

Regarding the choice of the target algebra B , some very convenient algebras are the random matrix ones, $B = M_K(C(T))$, with $K \in \mathbb{N}$, and T being a compact space. These algebras generalize indeed the most familiar algebras that we know, namely the matrix ones $M_K(\mathbb{C})$, and the commutative ones $C(T)$. We are led in this way to:

DEFINITION 11.24. *A matrix model for $G \subset U_N^+$ is a morphism of C^* -algebras*

$$\pi : C(G) \rightarrow M_K(C(T))$$

where T is a compact space, and $K \geq 1$ is an integer.

There are many examples of such models, and will discuss them later on. For the moment, let us develop some general theory. The question to be solved is that of understanding the suitable faithfulness assumptions needed on π , as for the model to “remind” the quantum group. As we will see, this is something quite tricky.

The simplest situation is when π is faithful in the usual sense. This is of course something quite restrictive, because the algebra $C(G)$ must be of type I in this case. However, there are many interesting examples here, and all this is worth a detailed look. Let us introduce the following notion, which is related to faithfulness:

DEFINITION 11.25. A matrix model $\pi : C(G) \rightarrow M_K(C(T))$ is called stationary when

$$\int_G = \left(\text{tr} \otimes \int_T \right) \pi$$

where \int_T is the integration with respect to a given probability measure on T .

Here the term “stationary” comes from a functional analytic interpretation of all this, with a certain Cesàro limit being needed to be stationary, and this will be explained later. Yet another explanation comes from a certain relation with the lattice models, but this relation is rather something folklore, not axiomatized yet. We will be back to this.

As a first result now, which is something which is not exactly trivial, and whose proof requires some functional analysis, the stationarity property implies the faithfulness:

THEOREM 11.26. Assuming that $G \subset U_N^+$ has a stationary model,

$$\pi : C(G) \rightarrow M_K(C(T))$$

$$\int_G = \left(\text{tr} \otimes \int_T \right) \pi$$

it follows that G must be coamenable, and that the model is faithful.

PROOF. We have two assertions to be proved, the idea being as follows:

(1) Assume that we have a stationary model, as in the statement. By performing the GNS construction with respect to \int_G , we obtain a factorization as follows, which commutes with the respective canonical integration functionals:

$$\pi : C(G) \rightarrow C(G)_{red} \subset M_K(C(T))$$

Thus, in what regards the coamenability question, we can assume that π is faithful. With this assumption made, observe that we have embeddings as follows:

$$C^\infty(G) \subset C(G) \subset M_K(C(T))$$

The point now is that the GNS construction gives a better embedding, as follows:

$$L^\infty(G) \subset M_K(L^\infty(T))$$

Now since the von Neumann algebra on the right is of type I, so must be its subalgebra $A = L^\infty(G)$. This means that, when writing the center of this latter algebra as $Z(A) = L^\infty(X)$, the whole algebra decomposes over X , as an integral of type I factors:

$$L^\infty(G) = \int_X M_{K_x}(\mathbb{C}) dx$$

In particular, we can see from this that $C^\infty(G) \subset L^\infty(G)$ has a unique C^* -norm, and so G is coamenable. Thus we have proved our first assertion.

(2) The second assertion follows as well from the above, because our factorization of π consists of the identity, and of an inclusion. \square

We will be back to this in chapters 13-16 below, when discussing matrix models for the quantum groups associated to the Hadamard matrices, that we met in the above.

11d. Type III algebras

In this final section we briefly discuss the reduction theory in the general case, type III. There are many things to be said here, well beyond the purposes of the present book, which basically focuses on type II algebras, and we will be quite brief.

In order to get started, we must discuss the type III factors, which are new to us. According to various conventions above, these factors are defined as follows:

DEFINITION 11.27. *A type III factor is a von Neumann algebra $A \subset B(H)$ which is a factor, $Z(A) = \mathbb{C}$, and which satisfies one of the following equivalent conditions:*

- (1) *A is not of type I, or of type II.*
- (2) *A has no semi-finite trace $tr : A \rightarrow \mathbb{C}$.*
- (3) *A has no trace $tr : A \rightarrow \mathbb{C}$, and is not of type I_∞ or II_∞ .*

In this definition (1) tells us that the type III factors are those indeed to be investigated next, (2) is a useful technical characterization, and (3) is a version of this characterization, which is important in connection with our general “quantum space” philosophy.

In order to investigate such factors, the general idea will be that of looking at the crossed products of type II factors, which can be lacking traces $tr : A \rightarrow \mathbb{C}$, and so which allow us to exit the type II world. In fact, as we will see in a moment, the main result about the type III factors, due to Connes, basically states that these appear from the type II factors, via various crossed product constructions, and their generalizations.

In order to get started, however, we have the following classical result:

THEOREM 11.28. *Any locally compact group G has a left invariant Haar measure λ , and a right invariant Haar measure ρ ,*

$$d\lambda(x) = d\lambda(yx) \quad , \quad d\rho(x) = d\lambda(xy)$$

which are unique up to multiplication by scalars. These two measures are absolutely continuous with respect to each other, and the Radon-Nikodym derivative

$$m : G \rightarrow \mathbb{R} \quad , \quad m(x) = \frac{d\lambda(x)}{d\rho(x)}$$

which is well-defined up to multiplication by scalars, is called modulus of the group. The unimodular groups, for which $m = 1$, include all compact groups, and all abelian groups.

PROOF. There are many things here, with everything being very classical, and the proof, along with comments, examples and more theory, especially in what regards the unimodular groups, can be found in any good measure theory book. \square

As is has become customary in this book, whenever talking about groups we must make some comments about quantum groups too. Things are quite interesting in connection with Theorem 11.28 above, because it is possible not only to generalize this result, with a suitable notion for the locally compact quantum groups, save for the existence result for the Haar measure, not established yet, but also to “twist” things in the compact case, as to have a notion of modulus there as well. We refer here to [98] and related papers.

In relation now with operator algebras, we have the following result:

THEOREM 11.29. *The type III factors basically appear from the type II factors, via various crossed product constructions, and their generalizations.*

PROOF. This statement is obviously something quite informal, and we will certainly not attempt to explain the proof either. Here are however the main ideas, with the result itself being basically due to Connes [24], along with some historical details:

(1) First of all, Murray and von Neumann knew of course about such questions, but were quite evasive in their papers about the type III factors, with the brief comment “we don’t know”. Whether they really worked or not on these questions, we’ll never know.

(2) Inspired by Theorem 11.28 above, it is possible to develop a whole machinery for the study of the non-tracial states $\varphi : A \rightarrow \mathbb{C}$, with the main results here being the Kubo-Martin-Schwinger (KMS) condition, and the Tomita-Takesaki theory. See [77].

(3) On the other hand, looking at the type II factors and their crossed products by automorphisms, which are not necessarily of type II, leads to a lot of interesting theory as well, leading to large classes of type III factors, appearing from type II factors.

(4) The above results are all basically from the 50s and 60s, and Connes was able to put all this together, in the early 70s, via a series of quick, beautiful and surprising Comptes Rendus notes, eventually leading to his paper [24], which is a must-read. \square

In equivalent terms, and also by remaining a bit informal, we have:

THEOREM 11.30. *The von Neumann algebra factors can be classified as follows,*

$$\begin{array}{c} \text{I}_N, \text{I}_\infty \\ \text{II}_1, \text{II}_\infty \\ \text{III}_0, \text{III}_\lambda, \text{III}_1 \end{array}$$

with the type II_1 ones being the most important, basically producing the others too.

PROOF. This follows by putting altogether what we have, results of Murray and von Neumann in type I and II, and then of Connes in type III. The last assertion is of course something quite informal, because the situation is not exactly as simple as that. \square

Getting back now to our series of reduction theory results, we have:

THEOREM 11.31. *Given an arbitrary von Neumann algebra $A \subset B(H)$, write its center as follows, with X being a measured space:*

$$Z(A) = L^\infty(X)$$

The whole algebra A decomposes then over this measured space X , as a direct sum of fibers, taken in an appropriate sense,

$$A = \int_X A_x dx$$

with the fibers A_x being von Neumann factors, which can be of type I, II, III.

As before with other such results, this is something heavy, generalizing our previous knowledge in type I, and then in type II. The proof however is quite similar, basically using the same ideas. We refer here to the literature, for instance to Dixmier [29].

11e. Exercises

Things have been quite technical in this chapter, which was more of a survey than something else, and as a unique exercise on all this, we have:

EXERCISE 11.32. *Learn some more basic von Neumann algebra theory, from the papers of von Neumann and Murray-von Neumann, then Tomita-Takesaki and Connes, and write down a brief account of what you learned.*

In what follows we will avoid ourselves this type of exercise, basically by getting back to the material in chapter 10, and building on that, following Jones.

CHAPTER 12

Hyperfiniteness

12a. The factor R

In this chapter we go back to the functional analysis methods for general von Neumann algebras developed in chapter 9, and to the theory of factors, and notably of the type II factors developed in chapter 10, with the aim of further building on this. Following old and classical work of Murray-von Neumann, our main object of study here will be the central example of a II factor, namely the “smallest” one, the hyperfinite II_1 factor R .

Once this factor R introduced, and its basic theory understood, we will go on a more advanced discussion, including more theory of R , following Connes, then a discussion of the connections with the considerations from chapter 11, and finally with an all-around discussion of various von Neumann algebra topics not discussed yet here, culminating with something quite technical, namely the classification of all hyperfinite factors.

Needless to say, all this is quite advanced, and there will be not many proofs in all this. This chapter will be a bit as a previous one, more of the survey. Also, let us mention that in the final part of the book, chapters 13-16 below, we will go back to a more normal pace, with a standard introduction to the Jones theory of the inclusions of II_1 factors, with full details. The notion of hyperfiniteness and the factor R will of course pop up there, every now and then, but usually at the end of each chapter, and most of the time using actually only its basic theory, and not most of the advanced material below.

In order to get started now, let us formulate the following definition:

DEFINITION 12.1. *A von Neumann algebra $A \subset B(H)$ is called hyperfinite when it appears as the weak closure of an increasing limit of finite dimensional algebras:*

$$A = \overline{\bigcup_i A_i}^w$$

When A is a II_1 factor, we call it hyperfinite II_1 factor, and we denote it by R .

As a first observation, there are many hyperfinite von Neumann algebras, for instance because any finite dimensional von Neumann algebra $A = \oplus_i M_{n_i}(\mathbb{C})$ is such an algebra, as one can see simply by taking $A_i = A$ for any i , in the above definition.

Also, given a measured space X , by using a dense sequence of points inside it, we can write $X = \bigcup_i X_i$ with $X_i \subset X$ being an increasing sequence of finite subspaces, and at the level of the corresponding algebras of functions this gives a decomposition as follows, which shows that the algebra $A = L^\infty(X)$ is hyperfinite, in the above sense:

$$L^\infty(X) = \overline{\bigcup_i L^\infty(X_i)}^w$$

The interesting point, however, is that when trying to construct II_1 factors which are hyperfinite, all the possible constructions lead in fact to the same factor, denoted R . This is an old theorem of Murray and von Neumann, that we will explain now.

In order to get started, we will need a number of technical ingredients. Generally speaking, our main tool will be that of using the expectation $E_i : A \rightarrow A_i$ from a hyperfinite von Neumann algebra A onto its finite dimensional subalgebras $A_i \subset A$, so talking about such conditional expectations will be our first task. Let us start with:

PROPOSITION 12.2. *Given an inclusion of finite von Neumann algebras $A \subset B$, there is a unique linear map*

$$E : B \rightarrow A$$

which is positive, unital, trace-preserving and satisfies the following condition:

$$E(b_1 a b_2) = b_1 E(a) b_2$$

This map is called conditional expectation from B onto A .

PROOF. We make use of the standard representation of the finite von Neumann algebra B , with respect to its trace $\text{tr} : B \rightarrow \mathbb{C}$, as constructed in chapter 9 above:

$$B \subset L^2(B)$$

If we denote by Ω the standard cyclic and separating vector of $L^2(B)$, we have an identification of vector spaces, as follows:

$$A\Omega = L^2(A)$$

Consider now the following orthogonal projection:

$$e : L^2(B) \rightarrow L^2(A)$$

It follows from definitions that we have an inclusion as follows:

$$e(B\Omega) \subset A\Omega$$

Thus e induces by restriction a certain linear map, as follows:

$$E : B \rightarrow A$$

This linear map E and the orthogonal projection e are then related by:

$$exe = E(x)e$$

But this shows that the linear map E satisfies the various conditions in the statement, namely positivity, unitality, trace preservation and bimodule property. As for the uniqueness assertion, this follows by using the same argument, applied backwards, the idea being that a map E as in the statement must come from the projection e . \square

Following Jones, we will be often interested in what follows in the orthogonal projection $e : L^2(B) \rightarrow L^2(A)$ producing the expectation $E : B \rightarrow A$, rather than in E itself:

DEFINITION 12.3. *Associated to any inclusion of finite von Neumann algebras $A \subset B$, as above, is the orthogonal projection*

$$e : L^2(B) \rightarrow L^2(A)$$

producing the conditional expectation $E : B \rightarrow A$ via the following formula:

$$exe = E(x)e$$

This projection is called Jones projection for the inclusion $A \subset B$.

We will heavily use Jones projections in chapters 13-16 below, in the context where both the algebras A, B are II_1 factors, when systematically studying the inclusions of such II_1 factors $A \subset B$, called subfactors. In connection with our present hyperfiniteness questions, the idea, already mentioned above, will be that of using the conditional expectation $E_i : A \rightarrow A_i$ from a hyperfinite von Neumann algebra A onto its finite dimensional subalgebras $A_i \subset A$, as well as its Jones projection versions $e_i : L^2(A) \rightarrow L^2(A_i)$.

Let us start with a technical approximation result, as follows:

PROPOSITION 12.4. *Assume that a von Neumann algebra $A \subset B(H)$ appears as an increasing limit of von Neumann subalgebras*

$$A = \overline{\bigcup_i A_i}^w$$

and denote by $E_i : A \rightarrow A_i$ the corresponding conditional expectations.

- (1) *We have $\|E_i(x) - x\| \rightarrow 0$, for any $x \in A$.*
- (2) *If $x_i \in A_i$ is a bounded sequence, satisfying $x_i = E_i(x_{i+1})$ for any i , then this sequence has a norm limit $x \in A$, satisfying $x_i = E_i(x)$ for any i .*

PROOF. Both the assertions are elementary, as follows:

(1) In terms of the Jones projections $e_i : L^2(A) \rightarrow L^2(A_i)$ associated to the expectations $E_i : A \rightarrow A_i$, the fact that the algebra A appears as the increasing union of its subalgebras A_i translates into the fact that the e_i are increasing, and converging to 1:

$$e_i \nearrow 1$$

But this gives $\|E_i(x) - x\| \rightarrow 0$, for any $x \in A$, as desired.

(2) Let $\{x_i\} \subset A$ be a sequence as in the statement. Since this sequence was assumed to be bounded, we can pick a weak limit $x \in A$ for it, and we have then, for any i :

$$E_i(x) = x_i$$

Now by (1) we obtain from this $\|x - x_n\| \rightarrow 0$, which gives the result. \square

We have now all the needed ingredients for formulating a first key result, in connection with the hyperfinite II_1 factors, due to Murray and von Neumann, as follows:

PROPOSITION 12.5. *Given an increasing union on matrix algebras, the following construction produces a hyperfinite II_1 factor*

$$R = \overline{\bigcup_{n_i} M_{n_i}(\mathbb{C})}^w$$

called Murray-von Neumann hyperfinite factor.

PROOF. This basically follows from the above, in two steps, as follows:

(1) The von Neumann algebra R constructed in the statement is hyperfinite by definition, with the remark here that the trace on it $tr : R \rightarrow \mathbb{C}$ comes as the increasing union of the traces on the matrix components $tr : M_{n_i}(\mathbb{C}) \rightarrow \mathbb{C}$, and with all the details here being elementary to check, by using the usual standard form technology.

(2) Thus, it remains to prove that R is a factor. For this purpose, pick an element belonging to its center, $x \in Z(R)$, and consider its expectation on $A_i = M_{n_i}(\mathbb{C})$:

$$x_i = E_i(x)$$

We have then $x_i \in Z(A_i)$, and since the matrix algebra $A_i = M_{n_i}(\mathbb{C})$ is a factor, we deduce from this that this expected value $x_i \in A_i$ is given by:

$$x_i = tr(x_i)1 = tr(x)1$$

On the other hand, Proposition 12.4 above applies, and shows that we have:

$$\|x_i - x\| = \|E_i(x) - x\| \rightarrow 0$$

Thus our element is a scalar, $x = tr(x)1$, and so R is a factor, as desired. \square

Next, we have the following substantial improvement of the above result, also due to Murray and von Neumann, which will be our final saying on the subject:

THEOREM 12.6. *Given an increasing union on matrix algebras, the following construction produces a II_1 factor, called Murray-von Neumann hyperfinite factor,*

$$R = \overline{\bigcup_{n_i} M_{n_i}(\mathbb{C})}^w$$

and the isomorphism class of this factor R does not depend on the exact sizes of the matrix algebras involved, nor on the particular inclusions between them.

PROOF. Here the first assertion, stating that R is indeed a II_1 factor, is something that we already know, from Proposition 12.5 above. As for the second assertion, regarding the uniqueness, this is something which comes from the magic of the weak topology, which by some kind of miracle makes everyone equal, in the end, the idea being as follows:

(1) Given a II_1 factor A , a von Neumann subalgebra $B \subset A$, and a subset $S \subset A$, let us write $S \subset_\varepsilon B$ when the following condition is satisfied, with $\|x\|_2 = \sqrt{\text{tr}(x^*x)}$:

$$\forall x \in S, \exists y \in B, \|x - y\|_2 \leq \varepsilon$$

With this convention made, given a II_1 factor A , the fact that this factor is hyperfinite in the sense of Definition 12.1 tells us that for any finite subset $S \subset A$, and any $\varepsilon > 0$, we can find a finite dimensional von Neumann subalgebra $B \subset A$ such that:

$$S \subset_\varepsilon B$$

(2) With this observation made, assume that we are given a hyperfinite II_1 factor A . Let us pick a dense sequence $\{x_k\} \subset A$, and let us set:

$$S_k = \{x_1, \dots, x_k\}$$

By choosing $\varepsilon = 1/k$ in the above, we can find, for any $k \in \mathbb{N}$, a finite dimensional von Neumann subalgebra $B_k \subset A$ such that the following condition is satisfied:

$$S_k \subset_{1/k} B_k$$

(3) Our first claim is that, by suitably choosing our subalgebra $B_k \subset A$, we can always assume that this is a matrix algebra, of the following special type:

$$B_k = M_{2^{n_k}}(\mathbb{C})$$

But this is something which is quite routine, which can be proved by starting with a finite dimensional subalgebra $B_k \subset A$ as above, and then perturbing its set of minimal projections $\{e_i\}$ into a set of projections $\{e'_i\}$ which are close in norm, and have as traces multiples of 2^n , with $n \gg 0$. Indeed, the algebra $B'_k \subset A$ having these new projections $\{e'_i\}$ as minimal projections will be then arbitrarily close to the algebra B_k , and so will still contain the subset S_k in the above approximate sense, and due to our trace condition, will be contained in a subalgebra of type $B''_k \simeq M_{2^{n_k}}(\mathbb{C})$, as desired.

(4) Our next claim, whose proof is similar, by using standard perturbation arguments for the corresponding sets of minimal projections, is that in the above the sequence of subalgebras $\{B_k\}$ can be chosen increasing. Thus, up to a rescaling of everything, we can assume that our sequence of subalgebras $\{B_k\}$ is as follows:

$$B_k = M_{2^k}(\mathbb{C})$$

(5) But this finishes the proof. Indeed, according to the above, we have managed to write our arbitrary hyperfinite II_1 factor A as a weak limit of the following type:

$$A = \overline{\bigcup_k M_{2^k}(\mathbb{C})}^w$$

Thus we have uniqueness indeed, and our result is proved. □

The above result is something quite fundamental, and adds to a series of similar results, or rather philosophical conclusions, which are quite surprising, as follows:

(1) We have seen early on in this book that, up to isomorphism, there is only one Hilbert space to be studied, namely the infinite dimensional separable Hilbert space, which can be taken to be, according to knowledge and taste, either $H = L^2(\mathbb{R})$, or $H = l^2(\mathbb{N})$.

(2) Regarding now the study of the operator algebras $A \subset B(H)$ over this unique Hilbert space, another somewhat surprising conclusion, from chapter 6 above, is that we won't miss much by assuming that $A = M_N(L^\infty(X))$ is a random matrix algebra.

(3) And now, guess what, what we just found is that when trying to get beyond random matrices, and what can be done with them, we are led to yet another unique von Neumann algebra, namely the above Murray-von Neumann hyperfinite II_1 factor R .

(4) And for things to be complete, we will see later that when getting beyond type II_1 , things won't change, because the other types of hyperfinite factors, not necessarily of type II_1 , can be all shown to ultimately come from R , via various constructions.

All this is certainly quite interesting, philosophically speaking. All in all, always the same conclusion, no need to go far to get to interesting algebras and questions: these interesting algebras and questions are just there, the most obvious ones.

Finally, as more advanced comment in relation with this, one recurrent question in von Neumann algebras, twisting the minds of many mathematicians, is "what is the von Neumann algebra to be considered, and what exactly is to be done with it".

There are many possible answers here, depending on knowledge and taste, and if you have a cat, best is to ask the cat. Cat will probably answer that "for vegans like you, the algebra to be considered is R , and the thing to do with it is to classify its subfactors".

This is actually an extremely difficult question already, and we will comment more on this in chapters 13-16 below, when systematically doing subfactors. As for the von Neumann algebra questions at the purely carnivorous level, felines only know.

12b. Amenability

The hyperfinite II_1 factor R , which is quite a fascinating object, was heavily investigated by Murray-von Neumann, and then by Connes. There are many things that can be said about it, which all interesting, but which are usually quite technical as well.

As a central result, in what regards advanced hyperfiniteness theory, we have:

THEOREM 12.7. *For a finite von Neumann algebra A , the following are equivalent:*

- (1) *A is hyperfinite in the usual sense, namely it appears as the weak closure of an increasing limit of finite dimensional algebras:*

$$A = \overline{\bigcup_i A_i}^w$$

- (2) *A amenable, in the sense that the standard inclusion $A \subset B(H)$, with $H = L^2(A)$, admits a conditional expectation $E : B(H) \rightarrow A$.*

PROOF. This result, due to Connes, is something heavy, the idea being as follows:

(1) \implies (2) Assuming that the algebra A is hyperfinite, let us write it as the weak closure of an increasing limit of finite dimensional subalgebras:

$$A = \overline{\bigcup_i A_i}^w$$

Consider the inclusion $A \subset B(H)$, with $H = L^2(A)$. In order to construct an expectation $E : B(H) \rightarrow A$, let us pick an ultrafilter ω on \mathbb{N} . Given $T \in B(H)$, we can define the following quantity, with μ_i being the Haar measure on the unitary group $U(A_i)$:

$$\psi(T) = \lim_{i \rightarrow \omega} \int_{U(A_i)} UTU^* d\mu_i(U)$$

With this construction made, by using now the standard involution $J : H \rightarrow H$, given by the formula $T \rightarrow T^*$, we can further define a map as follows:

$$E : B(H) \rightarrow A$$

$$E(T) = J\psi(T)J$$

But this is the expectation that we are looking for, with its left and right invariance properties coming from the left and right invariance of each Haar measure μ_i .

(2) \implies (1) This is something fairly heavy, using lots of advanced functional analysis, and for full details here, we refer to Connes' original paper [25]. \square

The above result is of course something quite abstract, and understanding its concrete implications will be our next task. As a first observation, in the case of the II_1 factors, that we are mostly interested in, in this book, the above result reads:

THEOREM 12.8. *For a II_1 factor R , the following are equivalent:*

- (1) *R amenable, in the sense that we have an expectation, as follows:*

$$E : B(L^2(R)) \rightarrow R$$

- (2) *R is the Murray-von Neumann hyperfinite II_1 factor.*

PROOF. This follows indeed from Theorem 12.7, when coupled with the Murray-von Neumann uniqueness result for the hyperfinite II_1 factor, from Theorem 12.6. \square

Getting back now to the general case, that of the finite von Neumann algebras, from Theorem 12.7, a first question is about how all this applies to the group von Neumann algebras, and more generally to the quantum group von Neumann algebras $L(\Gamma)$.

In order to discuss this, we will need the following result, which is standard:

THEOREM 12.9. *For a discrete group Γ , the following two conditions are equivalent, and if they are satisfied, we say that Γ is amenable:*

- (1) *Γ admits an invariant mean $m : l^\infty(\Gamma) \rightarrow \mathbb{C}$.*
 (2) *The projection map $C^*(\Gamma) \rightarrow C_{red}^*(\Gamma)$ is an isomorphism.*

Moreover, the class of amenable groups contains all the finite groups, all the abelian groups, and is stable under taking subgroups, quotients and products.

PROOF. This is something very standard, the idea being as follows:

(1) The equivalence (1) \iff (2) is standard, with the amenability conditions (1,2) being in fact part of a much longer list of amenability conditions, including well-known criteria of Følner, Kesten and others. We will be back to this, with details, in a moment, directly in a more general setting, that of the discrete quantum groups.

(2) As for the last assertion, regarding the finite groups, the abelian groups, and then the stability under taking subgroups, quotients and products, this is something elementary, which follows by using either of the above definitions of the amenability. \square

Getting back now to operator algebras, we can complement Theorem 12.7 with:

THEOREM 12.10. *For a group von Neumann algebra $A = L(\Gamma)$, the following conditions are equivalent:*

- (1) *A is hyperfinite.*
 (2) *A amenable.*
 (3) *Γ is amenable.*

PROOF. The group von Neumann algebras $A = L(\Gamma)$ being by definition finite, Theorem 12.7 above applies, and gives the equivalence (1) \iff (2). Thus, it remains to prove that we have (2) \iff (3), and we can prove this as follows:

(2) \implies (3) This is something clear, because if we assume that $A = L(\Gamma)$ is amenable, we have by definition a conditional expectation $E : B(L^2(A)) \rightarrow A$, and the restriction of this conditional expectation is the desired invariant mean $m : l^\infty(\Gamma) \rightarrow \mathbb{C}$.

(3) \implies (2) Assume that we are given a discrete amenable group Γ . In view of Theorem 12.9 above, this means that Γ has an invariant mean, as follows:

$$m : l^\infty(\Gamma) \rightarrow \mathbb{C}$$

Consider now the Hilbert space $H = l^2(\Gamma)$, and for any operator $T \in B(H)$ consider the following map, which is a bounded sesquilinear form:

$$\begin{aligned} \varphi_T : H \times H &\rightarrow \mathbb{C} \\ (\xi, \eta) &\rightarrow m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \xi, \eta \rangle] \end{aligned}$$

By using the Riesz representation theorem, we conclude that there exists a certain operator $E(T) \in B(H)$, such that the following holds, for any two vectors ξ, η :

$$\varphi_T(\xi, \eta) = \langle E(T)\xi, \eta \rangle$$

Summarizing, to any operator $T \in B(H)$ we have associated another operator, denoted $E(T) \in B(H)$, such that the following formula holds, for any two vectors ξ, η :

$$\langle E(T)\xi, \eta \rangle = m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \xi, \eta \rangle]$$

In order to prove now that this linear map E is the desired expectation, observe that for any group element $g \in \Gamma$, and any two vectors $\xi, \eta \in H$, we have:

$$\begin{aligned} \langle \rho_g E(T) \rho_g^* \xi, \eta \rangle &= \langle E(T) \rho_g^* \xi, \rho_g^* \eta \rangle \\ &= m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \rho_g^* \xi, \rho_g^* \eta \rangle] \\ &= m [\gamma \rightarrow \langle \rho_{g\gamma} T \rho_{g\gamma}^* \xi, \eta \rangle] \\ &= m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \xi, \eta \rangle] \\ &= \langle E(T)\xi, \eta \rangle \end{aligned}$$

Since this is valid for any $\xi, \eta \in H$, we conclude that we have, for any $g \in \Gamma$:

$$\rho_g E(T) \rho_g^* = E(T)$$

But this shows that the element $E(T) \in B(H)$ is in the commutant of the right regular representation of Γ , and so belongs to the left regular group algebra of Γ :

$$E(T) \in L(\Gamma)$$

Summarizing, we have constructed a certain linear map $E : B(H) \rightarrow L(\Gamma)$. Now by using the above explicit formula of it, in terms of $m : l^\infty(\Gamma) \rightarrow \mathbb{C}$, which was assumed to be an invariant mean, we conclude that E is indeed an expectation, as desired. \square

As a very concrete application of all this technology, in relation now with the discrete group algebras which are II_1 factors, the results that we have lead to:

THEOREM 12.11. *For a discrete group Γ , the following conditions are equivalent:*

- (1) Γ is amenable, and has the ICC property.
- (2) $A = L(\Gamma)$ is the hyperfinite II_1 factor R .

PROOF. This follows indeed from Theorem 12.10 above, coupled with the standard fact, that we know well from chapter 10 above, that a group algebra $A = L(\Gamma)$ is a factor, and so a II_1 factor, precisely when the group Γ has the ICC property. \square

This was for the basic theory of R . It is possible to further build on the above results, with concrete examples, and also by combining them with the reduction theory results from chapter 11 above. We will discuss this later, at the end of the present chapter.

12c. Quantum groups

With the above basic theory of R understood, or at least explained as above, let us go back now to the various quantum group considerations from chapters 10-11 above. Our main goal will be that of extending to quantum groups what we learned about groups.

As a first question, in relation with the considerations from chapter 10, we would like to understand which discrete quantum groups Γ produce group algebras as follows:

$$L(\Gamma) \simeq R$$

In terms of the compact quantum group duals $G = \widehat{\Gamma}$, the problem is that of understanding which compact quantum groups G produce group algebras as follows:

$$L^\infty(G) \simeq R$$

In order to discuss this, we must first talk about amenability. We have here the following result, basically due to Woronowicz [98], and coming from the Peter-Weyl theory, extending to the discrete quantum groups the standard theory for discrete groups:

THEOREM 12.12. *Let (A, u) with $u \in M_N(A)$ be a Woronowicz algebra, as axiomatized before. Let A_{full} be the enveloping C^* -algebra of $\mathcal{A} = \langle u_{ij} \rangle$, and let A_{red} be the quotient of A by the null ideal of the Haar integration. The following are then equivalent:*

- (1) *The Haar functional of A_{full} is faithful.*
- (2) *The projection map $A_{full} \rightarrow A_{red}$ is an isomorphism.*
- (3) *The counit map $\varepsilon : A \rightarrow \mathbb{C}$ factorizes through A_{red} .*
- (4) *We have $N \in \sigma(\text{Re}(\chi_u))$, the spectrum being taken inside A_{red} .*
- (5) *$\|ax_k - \varepsilon(a)x_k\| \rightarrow 0$ for any $a \in \mathcal{A}$, for certain norm 1 vectors $x_k \in L^2(A)$.*

If this is the case, we say that the underlying discrete quantum group Γ is amenable.

PROOF. Before starting, we should mention that amenability and the present result are a bit like the Spectral Theorem, in the sense that knowing that the result formally holds does not help much, and in practice, one needs to remember the proof as well.

For this reason, we will work out explicitly all the possible implications between (1-5), whenever possible, adding to the global formal proof, which will be linear, as follows:

$$(1) \implies (2) \implies (3) \implies (4) \implies (5) \implies (1)$$

In order to prove these implications, and the other ones too, the general idea is that this is well-known in the group dual case, $A = C^*(\Gamma)$, with Γ being a usual discrete group, and in general, the result follows by adapting the group dual case proof.

(1) \iff (2) This follows from the fact that the GNS construction for the algebra A_{full} with respect to the Haar functional produces the algebra A_{red} .

(2) \implies (3) This is trivial, because we have quotient maps $A_{full} \rightarrow A \rightarrow A_{red}$, and so our assumption $A_{full} = A_{red}$ implies that we have $A = A_{red}$.

(3) \implies (2) Assume indeed that we have a counit map, as follows:

$$\varepsilon : A_{red} \rightarrow \mathbb{C}$$

In order to prove $A_{full} = A_{red}$, we can use the right regular corepresentation. Indeed, we can define such a corepresentation by the following formula:

$$W(a \otimes x) = \Delta(a)(1 \otimes x)$$

This corepresentation is unitary, so we can define a morphism as follows:

$$\begin{aligned} \Delta' : A_{red} &\rightarrow A_{red} \otimes A_{full} \\ a &\rightarrow W(a \otimes 1)W^* \end{aligned}$$

Now by composing with $\varepsilon \otimes id$, we obtain a morphism as follows:

$$\begin{aligned} (\varepsilon \otimes id)\Delta' : A_{red} &\rightarrow A_{full} \\ u_{ij} &\rightarrow u_{ij} \end{aligned}$$

Thus, we have our inverse for the canonical projection $A_{full} \rightarrow A_{red}$, as desired.

(3) \implies (4) This implication is clear, because we have:

$$\begin{aligned} \varepsilon(Re(\chi_u)) &= \frac{1}{2} \left(\sum_{i=1}^N \varepsilon(u_{ii}) + \sum_{i=1}^N \varepsilon(u_{ii}^*) \right) \\ &= \frac{1}{2}(N + N) \\ &= N \end{aligned}$$

Thus the element $N - Re(\chi_u)$ is not invertible in A_{red} , as claimed.

(4) \implies (3) In terms of the corepresentation $v = u + \bar{u}$, whose dimension is $2N$ and whose character is $2Re(\chi_u)$, our assumption $N \in \sigma(Re(\chi_u))$ reads:

$$\dim v \in \sigma(\chi_v)$$

By functional calculus the same must hold for $w = v + 1$, and then once again by functional calculus, the same must hold for any tensor power of w :

$$w_k = w^{\otimes k}$$

Now choose for each $k \in \mathbb{N}$ a state $\varepsilon_k \in A_{red}^*$ having the following property:

$$\varepsilon_k(w_k) = \dim w_k$$

By Peter-Weyl we must have $\varepsilon_k(r) = \dim r$ for any $r \leq w_k$, and since any irreducible corepresentation appears in this way, the sequence ε_k converges to a counit map:

$$\varepsilon : A_{red} \rightarrow \mathbb{C}$$

(4) \implies (5) Consider the following elements of A_{red} , which are positive:

$$a_i = 1 - Re(u_{ii})$$

Our assumption $N \in \sigma(Re(\chi_u))$ tells us that $a = \sum a_i$ is not invertible, and so there exists a sequence x_k of norm one vectors in $L^2(A)$ such that:

$$\langle ax_k, x_k \rangle \rightarrow 0$$

Since the summands $\langle a_i x_k, x_k \rangle$ are all positive, we must have, for any i :

$$\langle a_i x_k, x_k \rangle \rightarrow 0$$

We can go back to the variables u_{ii} by using the following general formula:

$$\|vx - x\|^2 = \|vx\|^2 + 2 \langle (1 - Re(v))x, x \rangle - 1$$

Indeed, with $v = u_{ii}$ and $x = x_k$ the middle term on the right goes to 0, and so the whole term on the right becomes asymptotically negative, and so we must have:

$$\|u_{ii}x_k - x_k\| \rightarrow 0$$

Now let $M_n(A_{red})$ act on $\mathbb{C}^n \otimes L^2(A)$. Since u is unitary we have:

$$\sum_i \|u_{ij}x_k\|^2 = \|u(e_j \otimes x_k)\| = 1$$

From $\|u_{ii}x_k\| \rightarrow 1$ we obtain $\|u_{ij}x_k\| \rightarrow 0$ for $i \neq j$. Thus we have, for any i, j :

$$\|u_{ij}x_k - \delta_{ij}x_k\| \rightarrow 0$$

Now by remembering that we have $\varepsilon(u_{ij}) = \delta_{ij}$, this formula reads:

$$\|u_{ij}x_k - \varepsilon(u_{ij})x_k\| \rightarrow 0$$

By linearity, multiplicativity and continuity, we must have, for any $a \in \mathcal{A}$, as desired:

$$\|ax_k - \varepsilon(a)x_k\| \rightarrow 0$$

(5) \implies (1) This is something well-known, which follows via some standard functional analysis arguments, exactly as in the usual group case.

(1) \implies (5) Once again this is something well-known, which follows via some standard functional analysis arguments, exactly as in the usual group case. \square

Before getting further, with advanced amenability and hyperfiniteness questions, and as a first application of the above, we can now advance on a problem that we left open before, in chapter 7 above, when talking about cocommutative Woronowicz algebras. Indeed, we can now state and prove the following result, which clarifies the situation:

THEOREM 12.13. *The cocommutative Woronowicz algebras are the intermediate quotients of the following type, with $\Gamma = \langle g_1, \dots, g_N \rangle$ being a discrete group,*

$$C^*(\Gamma) \rightarrow C_\pi^*(\Gamma) \rightarrow C_{red}^*(\Gamma)$$

and with π being a unitary representation of Γ , subject to weak containment conditions of type $\pi \otimes \pi \subset \pi$ and $1 \subset \pi$, which guarantee the existence of Δ, ε .

PROOF. We use the various findings from Theorem 12.12 above, following Woronowicz, the idea being to proceed in several steps, as follows:

(1) Theorem 12.12 and standard functional analysis arguments show that the cocommutative Woronowicz algebras should appear as intermediate quotients, as follows:

$$C^*(\Gamma) \rightarrow A \rightarrow C_{red}^*(\Gamma)$$

(2) The existence of $\Delta : A \rightarrow A \otimes A$ requires our intermediate quotient to appear as follows, with π being a unitary representation of Γ , satisfying the condition $\pi \otimes \pi \subset \pi$, taken in a weak containment sense, and with the tensor product \otimes being taken here to be compatible with our usual maximal tensor product \otimes for the C^* -algebras:

$$C^*(\Gamma) \rightarrow C_\pi^*(\Gamma) \rightarrow C_{red}^*(\Gamma)$$

(3) With this condition imposed, the existence of the antipode $S : A \rightarrow A^{opp}$ is then automatic, coming from the group antirepresentation $g \rightarrow g^{-1}$.

(4) The existence of the counit $\varepsilon : A \rightarrow \mathbb{C}$, however, is something non-trivial, related to amenability, and leading to a condition of type $1 \subset \pi$, as in the statement. \square

The above result is of course not the end of the story, because as formulated, with the above highly abstract conditions on π , it comes along with zero non-trivial examples. We refer to Woronowicz's paper [98] for more on these topics.

Getting back now to real life, and concrete mathematics, let us focus on the Kesten amenability criterion, from Theorem 12.12 (4) above, which brings connections with interesting mathematics and physics, and which in practice will be our main amenability criterion. In order to discuss this, we will need the following standard fact:

PROPOSITION 12.14. *Given a Woronowicz algebra (A, u) , with $u \in M_N(A)$, the moments of the main character $\chi = \sum_i u_{ii}$ are given by:*

$$\int_G \chi^k = \dim(\text{Fix}(u^{\otimes k}))$$

In the case $u \sim \bar{u}$ the law of χ is a usual probability measure, supported on $[-N, N]$.

PROOF. There are two assertions here, the proof being as follows:

(1) The first assertion follows from the Peter-Weyl theory, which tells us that we have the following formula, valid for any corepresentation $v \in M_n(A)$:

$$\int_G \chi_v = \dim(\text{Fix}(v))$$

Indeed, with $v = u^{\otimes k}$ the corresponding character is given by:

$$\chi_v = \chi^k$$

Thus, we obtain the result, as a consequence of the above formula.

(2) As for the second assertion, if we assume $u \sim \bar{u}$ then we have:

$$\chi = \chi^*$$

Thus the general spectral measure theory, explained in the above, tells us that $\text{law}(\chi)$ is in this case a real probability measure, supported by the spectrum of χ . But, since the matrix $u \in M_N(A)$ is by definition a unitary, we have:

$$\begin{aligned} uu^* = 1 &\implies \|u_{ij}\| \leq 1, \forall i, j \\ &\implies \|\chi\| \leq N \end{aligned}$$

Thus the spectrum of the character satisfies the following condition:

$$\sigma(\chi) \subset [-N, N]$$

Thus, we are led to the conclusion in the statement. □

In relation now with the notion of amenability, we have:

THEOREM 12.15. *A Woronowicz algebra (A, u) , with $u \in M_N(A)$, is amenable when*

$$N \in \text{supp}(\text{law}(\text{Re}(\chi)))$$

and the support on the right depends only on $\text{law}(\chi)$.

PROOF. There are two assertions here, the proof being as follows:

(1) According to the Kesten amenability criterion, from Theorem 12.12 (4) above, the algebra A is amenable when the following condition is satisfied:

$$N \in \sigma(\text{Re}(\chi))$$

Now since $Re(\chi)$ is self-adjoint, we know from spectral theory that the support of its spectral measure $law(Re(\chi))$ is precisely its spectrum $\sigma(Re(\chi))$, as desired:

$$supp(law(Re(\chi))) = \sigma(Re(\chi))$$

(2) Regarding the second assertion, once again the variable $Re(\chi)$ being self-adjoint, its law depends only on the moments $\int_G Re(\chi)^p$, with $p \in \mathbb{N}$. But, we have:

$$\begin{aligned} \int_G Re(\chi)^p &= \int_G \left(\frac{\chi + \chi^*}{2} \right)^p \\ &= \frac{1}{2^p} \sum_{|k|=p} \int_G \chi^k \end{aligned}$$

Thus $law(Re(\chi))$ depends only on $law(\chi)$, and this gives the result. □

Let us work out now in detail the group dual case. Here we obtain a very interesting measure, called Kesten measure of the group, as follows:

PROPOSITION 12.16. *In the case $A = C^*(\Gamma)$ and $u = diag(g_1, \dots, g_N)$, and with the following normalization made,*

$$1 \in u = \bar{u}$$

the moments of the main character are given by the formula

$$\int_{\hat{\Gamma}} \chi^p = \# \left\{ i_1, \dots, i_p \mid g_{i_1} \dots g_{i_p} = 1 \right\}$$

counting the loops based at 1, having lenght p , on the corresponding Cayley graph.

PROOF. Consider indeed a discrete group $\Gamma = \langle g_1, \dots, g_N \rangle$. The main character of $A = C^*(\Gamma)$, with fundamental corepresentation $u = diag(g_1, \dots, g_N)$, is then:

$$\chi = g_1 + \dots + g_N$$

Given a colored integer $k = e_1 \dots e_p$, the corresponding moment is given by:

$$\begin{aligned} \int_{\hat{\Gamma}} \chi^k &= \int_{\hat{\Gamma}} (g_1 + \dots + g_N)^k \\ &= \# \left\{ i_1, \dots, i_p \mid g_{i_1}^{e_1} \dots g_{i_p}^{e_p} = 1 \right\} \end{aligned}$$

In the self-adjoint case, $u \sim \bar{u}$, we are only interested in the moments with respect to usual integers, $p \in \mathbb{N}$, and the above formula becomes:

$$\int_{\hat{\Gamma}} \chi^p = \# \left\{ i_1, \dots, i_p \mid g_{i_1} \dots g_{i_p} = 1 \right\}$$

Assume now that we have in addition $1 \in u$, so that the condition $1 \in u = \bar{u}$ in the statement is satisfied. At the level of the generating set $S = \{g_1, \dots, g_N\}$ this means:

$$1 \in S = S^{-1}$$

Thus the corresponding Cayley graph is well-defined, with the elements of Γ as vertices, and with the edges $g - h$ appearing when the following condition is satisfied:

$$gh^{-1} \in S$$

A loop on this graph based at 1, having length p , is then a sequence as follows:

$$(1) - (g_{i_1}) - (g_{i_1}g_{i_2}) - \dots - (g_{i_1} \dots g_{i_{p-1}}) - (g_{i_1} \dots g_{i_p} = 1)$$

Thus the moments of χ count indeed such loops, as claimed. □

In order to generalize the above result to arbitrary Woronowicz algebras, we can use the discrete quantum group philosophy. The fundamental result here is as follows:

THEOREM 12.17. *Let (A, u) be a Woronowicz algebra, and assume, by enlarging if necessary u , that we have $1 \in u = \bar{u}$. The following formula*

$$d(v, w) = \min \left\{ k \in \mathbb{N} \mid 1 \subset \bar{v} \otimes w \otimes u^{\otimes k} \right\}$$

defines then a distance on $\text{Irr}(A)$, which coincides with the geodesic distance on the associated Cayley graph. In the group dual case we obtain the usual distance.

PROOF. The fact that the lengths are finite follows from Woronowicz's analogue of Peter-Weyl theory, and the other verifications are as follows:

- (1) The symmetry axiom is clear.
- (2) The triangle inequality is elementary to establish as well.
- (3) Finally, the last assertion is elementary as well.

In the group dual case now, where our Woronowicz algebra is of the form $A = C^*(\Gamma)$, with $\Gamma = \langle S \rangle$ being a finitely generated discrete group, our normalization condition $1 \in u = \bar{u}$ means that the generating set must satisfy:

$$1 \in S = S^{-1}$$

But this is precisely the normalization condition for the discrete groups, and the fact that we obtain the same metric space is clear. □

Summarizing, we have a good understanding of what a discrete quantum group is. We can now formulate a generalization of Proposition 12.16, as follows:

THEOREM 12.18. *Let (A, u) be a Woronowicz algebra, with the normalization assumption $1 \in u = \bar{u}$ made. The moments of the main character,*

$$\int_G \chi^p = \dim(\text{Fix}(u^{\otimes p}))$$

count then the loops based at 1, having length p , on the corresponding Cayley graph.

PROOF. Here the formula of the moments, with $p \in \mathbb{N}$, is the one coming from Proposition 12.14 above, and the Cayley graph interpretation comes from Theorem 12.17. □

As an application of this, we can introduce the notion of growth, as follows:

DEFINITION 12.19. *Given a closed subgroup $G \subset U_N^+$, with $1 \in u = \bar{u}$, consider the series whose coefficients are the ball volumes on the corresponding Cayley graph,*

$$f(z) = \sum_k b_k z^k$$

$$b_k = \sum_{l(v) \leq k} \dim(v)^2$$

and call it growth series of the discrete quantum group \widehat{G} . In the group dual case, $G = \widehat{\Gamma}$, we obtain in this way the usual growth series of Γ .

There are many things that can be said about the growth, and we will be back to this. As a first such result, in relation with the notion of amenability, we have:

THEOREM 12.20. *Polynomial growth implies amenability.*

PROOF. We recall from Theorem 12.17 above that the Cayley graph of \widehat{G} has by definition the elements of $Irr(G)$ as vertices, and the distance is as follows:

$$d(v, w) = \min \left\{ k \in \mathbb{N} \mid 1 \subset \bar{v} \otimes w \otimes u^{\otimes k} \right\}$$

By taking $w = 1$ and by using Frobenius reciprocity, the lengths are given by:

$$l(v) = \min \left\{ k \in \mathbb{N} \mid v \subset u^{\otimes k} \right\}$$

By Peter-Weyl we have then a decomposition as follows, where B_k is the ball of radius k , and where $m_k(v) \in \mathbb{N}$ are certain multiplicities:

$$u^{\otimes k} = \sum_{v \in B_k} m_k(v) \cdot v$$

By using now Cauchy-Schwarz, we obtain the following inequality:

$$\begin{aligned} m_{2k}(1)b_k &= \sum_{v \in B_k} m_k(v)^2 \sum_{v \in B_k} \dim(v)^2 \\ &\geq \left(\sum_{v \in B_k} m_k(v) \dim(v) \right)^2 \\ &= N^{2k} \end{aligned}$$

But shows that if b_k has polynomial growth, then the following happens:

$$\limsup_{k \rightarrow \infty} m_{2k}(1)^{1/2k} \geq N$$

Thus, the Kesten type criterion applies, and gives the result. □

In addition to the above, there are many other things that can be said about the discrete quantum groups, and about amenability in general, at this quantum group level. We will be back to these questions in what follows, on several occasions.

To summarize now, we have a good understanding of what a discrete quantum group is. In relation now with von Neumann algebra questions, we have:

THEOREM 12.21. *For a discrete group von Neumann algebra $A = L(\Gamma)$, the following conditions are equivalent:*

- (1) A is hyperfinite.
- (2) A amenable.
- (3) Γ is amenable.

PROOF. This is something that we already know for the usual discrete groups, from Theorem 12.10 above. The proof for the general discrete quantum groups is similar, by using the theory developed above. To be more precise, the situation is as follows:

(1) \iff (2) The quantum group von Neumann algebras $A = L(\Gamma)$ being by definition finite, Theorem 12.7 above applies, and gives this equivalence.

(2) \implies (3) This is something clear, because if we assume that $A = L(\Gamma)$ is amenable, we have by definition a conditional expectation $E : B(L^2(A)) \rightarrow A$, and the restriction of this conditional expectation is the desired invariant mean $m : l^\infty(\Gamma) \rightarrow \mathbb{C}$.

(3) \implies (2) In the classical group case, assume that we are given a discrete amenable group Γ . This means that Γ has an invariant mean, as follows:

$$m : l^\infty(\Gamma) \rightarrow \mathbb{C}$$

Consider now the Hilbert space $H = l^2(\Gamma)$, and for any operator $T \in B(H)$ consider the following map, which is a bounded sesquilinear form:

$$\varphi_T : H \times H \rightarrow \mathbb{C}$$

$$(\xi, \eta) \rightarrow m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \xi, \eta \rangle]$$

By using the Riesz representation theorem, we conclude that there exists a certain operator $E(T) \in B(H)$, such that the following holds, for any two vectors ξ, η :

$$\varphi_T(\xi, \eta) = \langle E(T)\xi, \eta \rangle$$

Summarizing, to any operator $T \in B(H)$ we have associated another operator, denoted $E(T) \in B(H)$, such that the following formula holds, for any two vectors ξ, η :

$$\langle E(T)\xi, \eta \rangle = m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \xi, \eta \rangle]$$

In order to prove now that this linear map E is the desired expectation, observe that for any group element $g \in \Gamma$, and any two vectors $\xi, \eta \in H$, we have:

$$\begin{aligned} \langle \rho_g E(T) \rho_g^* \xi, \eta \rangle &= \langle E(T) \rho_g^* \xi, \rho_g^* \eta \rangle \\ &= m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \rho_g^* \xi, \rho_g^* \eta \rangle] \\ &= m [\gamma \rightarrow \langle \rho_{g\gamma} T \rho_{g\gamma}^* \xi, \eta \rangle] \\ &= m [\gamma \rightarrow \langle \rho_\gamma T \rho_\gamma^* \xi, \eta \rangle] \\ &= \langle E(T) \xi, \eta \rangle \end{aligned}$$

Since this is valid for any $\xi, \eta \in H$, we conclude that we have, for any $g \in \Gamma$:

$$\rho_g E(T) \rho_g^* = E(T)$$

But this shows that the element $E(T) \in B(H)$ is in the commutant of the right regular representation of Γ , and so belongs to the left regular group algebra of Γ :

$$E(T) \in L(\Gamma)$$

Summarizing, we have constructed a certain linear map, as follows:

$$E : B(H) \rightarrow L(\Gamma)$$

Now by using the above explicit formula of it, in terms of $m : l^\infty(\Gamma) \rightarrow \mathbb{C}$, which was assumed to be an invariant mean, we conclude that E is indeed an expectation, as desired. This finishes the proof in the case where Γ is a classical discrete group, and the proof in general, for discrete quantum groups, is similar. We refer here to [31]. \square

The above result is of course not the end of the story, because there is still a discussion to be made in relation with factoriality, and so with the ICC property, whose quantum formulation is not known yet. In the lack of something better, let us formulate:

THEOREM 12.22. *For a discrete quantum group Γ , the following two conditions are equivalent:*

- (1) Γ is amenable, and has the ICC property, in the sense that $Z(L(\Gamma)) = \mathbb{C}$.
- (2) The group algebra $A = L(\Gamma)$ is the hyperfinite II_1 factor R .

PROOF. This is something that we already know for the usual discrete groups, from Theorem 12.11 above, and with ICC corresponding to the usual ICC property. For the discrete quantum groups this is something which follows from Theorem 12.21 above, with the remark that we have cheated when defining the ICC property as above. \square

Summarizing, and no surprise here, the situation for the discrete quantum groups appears to be quite similar to the situation for the discrete groups, at least when seen from a distance, but in practice there are many interesting questions, still open.

Finally, let us mention that in connection with amenability and hyperfiniteness, we have as well a series of more advanced questions, in relation with the actions of groups,

or more generally quantum groups, which can be either discrete or compact, on the finite von Neumann algebras, and the corresponding crossed products, and fixed point algebras.

To be more precise, the problems that we would like to solve are as follows:

(1) We would like to understand, given a compact group or compact quantum group action on a von Neumann algebra, $G \curvearrowright P$, when the fixed point algebra P^G is a factor.

(2) More generally, we would like to understand under which assumptions on $G \curvearrowright P$, the fixed point algebra $(P \otimes A)^G$ is a factor, for any finite dimensional algebra A .

(3) We are particularly interested in the case of the dual actions, where $G \curvearrowright P$ comes via $P = R \rtimes \Gamma$ from a discrete group or quantum group action $\Gamma \curvearrowright R$.

(4) Finally, in all the above, we would like to understand when the corresponding fixed point algebra P^G , or more generally $(P \otimes A)^G$, is the hyperfinite II_1 factor R .

These questions are all interesting, from a theoretical point of view, and so worth investigating. There are also some deeper reasons for looking into these questions, coming from subfactor theory, the idea here being that the main examples of subfactors, all of integer index, appear as inclusions of II_1 factors of the form $(P \otimes A)^G$, as above.

As a basic result here, in connection with question (4) above, we have:

THEOREM 12.23. *The product type coaction $\gamma_\pi : R \rightarrow R \otimes \widehat{A}$ associated to an inner faithful representation $\pi : A \rightarrow M_k(\mathbb{C})$ containing the counit is outer.*

PROOF. This result, from [2], is something quite technical, which can be proved either by using some subfactor theory from chapters 13-16 below, or directly, as in Vaes' paper [81], by using heavy functional analysis methods. We present below the idea of the original proof, from [2], based on the subfactor material from chapters 13-16 below:

(1) By using standard subfactor results, we have to prove that for any corepresentation $w \in M_n(A)$, the relative commutant of the following inclusion is $\text{End}(w)$:

$$R \subset (M_n(\mathbb{C}) \otimes (R \rtimes_{\gamma_\pi} \widehat{A}))^{\widehat{\gamma}_{\pi w}}$$

(2) We may assume that the coefficients of w generate A . It is easy to see that the above inclusion is isomorphic to the following inclusion, where $w^t \in M_n \otimes A_\sigma$ is the corepresentation equal by definition to the transpose of w :

$$R \subset (M_n(\mathbb{C}) \otimes (R \rtimes_{\gamma_\pi} \widehat{A}_\sigma))^{i_{w^t} \circ \widehat{\gamma}_\pi}$$

(3) Again by using standard subfactor theory results, this latter inclusion is isomorphic to the vertical subfactor associated to the following biunitary matrix:

$$u = (id \otimes \pi)w^t$$

(4) Thus the relative commutant is $End(y)$, where (B, y, ν) is the minimal model for the following biunitary, where $\rho = (\pi \otimes t\pi S)\sigma\Delta$:

$$\begin{aligned} u_{12}u_{13}^{*t} &= ((id \otimes \pi)w^t)_{12}((id \otimes t\pi)\bar{w})_{13} \\ &= (id \otimes \rho)w^t \end{aligned}$$

(5) Since the representation π contains the counit, the spaces of coefficients of both π and $t\pi S$ are contained in the space of coefficients of ρ . It follows that ρ is inner faithful, and so that (A_σ, w^t, ρ) is the minimal model for $u_{12}u_{13}^{*t}$. Thus, we have:

$$(B, y, \nu) = (A_\sigma, w^t, \rho)$$

(6) Thus our commutant that we are interested in is $End(w^t) = End(w)$, as desired, and this leads to the conclusion in the statement. \square

There are many open problems here, especially in connection with the existence of such actions, following Ocneanu and others. We will be back to this, later.

Finally, let us mention that, exactly as it was the case in connection with the ICC property, where things are more complicated for quantum groups than for groups, things are similar in connection with the above questions (1-4). We will be back to this.

12d. Hyperfinite factors

In this final section we discuss the general hyperfinite factors. We must first make some upgrades to our terminology and notations regarding the factors, as follows:

- (1) The matrix algebra $M_N(\mathbb{C})$ is of type I_N .
- (2) The operator algebra $B(H)$ is of type I_∞ .
- (3) The II_1 factors are as before, infinite dimensional, and with a trace.
- (4) The tensor products $A \otimes B(H)$, with A being a II_1 factor, are of type II_∞ .
- (5) As for the factors left, these are called of type III.

It is possible to be more abstract here, but in practice, this is how these factors are best remembered. We will be back to this, with some abstract results as well.

We recall from chapter 11 above that we have the following result:

THEOREM 12.24. *The type III factors basically appear from the type II factors, via various crossed product constructions, and their generalizations.*

PROOF. This is something discussed in chapter 11 above, with the result being basically due to Connes [24], and with the idea, and historical details, being as follows:

(1) First of all, Murray and von Neumann knew of course about such questions, but were quite evasive in their papers about the type III factors, with the brief comment “we don’t know”. Whether they really worked or not on these questions, we’ll never know.

(2) Inspired by the locally compact group theory, it is possible to develop a whole machinery for the study of non-tracial states $\varphi : A \rightarrow \mathbb{C}$, with the main results being the Kubo-Martin-Schwinger (KMS) condition, and the Tomita-Takesaki theory. See [77].

(3) On the other hand, looking at the type II factors and their crossed products by automorphisms, which are not necessarily of type II, leads to a lot of interesting theory as well, leading to large classes of type III factors, appearing from type II factors.

(4) The above results are all basically from the 50s and 60s, and Connes was able to put all this together, in the early 70s, via a series of quick, beautiful and surprising Comptes Rendus notes, eventually leading to his paper [24], which is a must-read. \square

In equivalent terms, and also by remaining a bit informal, we have:

THEOREM 12.25. *The von Neumann algebra factors can be classified as follows,*

$$\begin{aligned} & \text{I}_N, \text{I}_\infty \\ & \text{II}_1, \text{II}_\infty \\ & \text{III}_0, \text{III}_\lambda, \text{III}_1 \end{aligned}$$

with the type II_1 ones being the most important, basically producing the others too.

PROOF. This follows by putting altogether what we have, results of Murray and von Neumann in type I and II, and then of Connes in type III. The last assertion is of course something quite informal, because the situation is not exactly as simple as that. \square

On the other hand, as part of the series of reduction theory results, we have:

THEOREM 12.26. *Given an arbitrary von Neumann algebra $A \subset B(H)$, write its center as follows, with X being a measured space:*

$$Z(A) = L^\infty(X)$$

The whole algebra A decomposes then over this measured space X , as a direct sum of fibers, taken in an appropriate sense,

$$A = \int_X A_x dx$$

with the fibers A_x being von Neumann factors, which can be of type I, II, III.

PROOF. This is something heavy, generalizing our previous results in type I, and then II. The proof however is quite similar, basically using the same ideas. \square

Getting back now to hyperfiniteness, we have here the following result:

THEOREM 12.27. *The hyperfinite factors are as follows, with 1 factor in each class*

$$\begin{aligned} & \text{I}_N, \text{I}_\infty \\ & \text{II}_1, \text{II}_\infty \\ & \text{III}_0, \text{III}_\lambda, \text{III}_1 \end{aligned}$$

with the type II_1 one R being the most important, basically producing the others too.

PROOF. This is something heavy, based on early work of Murray and Von Neumann in type II, and then of heavy work by Connes in type II and III [24], [25], basically finishing the classification, and with a final contribution by Haagerup in type III_1 [37]. \square

It is possible as well to formulate a reduction theory result, as follows:

THEOREM 12.28. *Given a hyperfinite von Neumann algebra $A \subset B(H)$, write its center as follows, with X being a measured space:*

$$Z(A) = L^\infty(X)$$

The whole algebra A decomposes then over this measured space X , as follows,

$$A = \int_X A_x dx$$

with the fibers A_x being hyperfinite von Neumann factors, which can be of type I, II, III.

PROOF. This is something heavy, combining the general reduction theory results with the work of above Connes and others on the hyperfinite case. \square

We should mention that the story with the general classification of factors, and with the subsequent reduction theory results, is not terminated by the above results.

Getting back now to the general classification work for II_1 factors, which is still ongoing, let us mention here that the main open problem is that of deciding whether the free group factors $L(F_2)$ and $L(F_3)$ are isomorphic or not:

$$L(F_2) \simeq? L(F_3)$$

This question can be of course asked in crossed product form, in the spirit of the various crossed product results above, and of advanced ergodic theory in general, with the space in question, producing the crossed product, being the point:

$$\{.\} \rtimes F_2 \simeq? \{.\} \rtimes F_3$$

This formulation has the advantage of putting the above problem into a more conceptual framework, with lots of advanced machinery available around. However, it is not clear whether this formulation simplifies or not the original problem.

Let us mention as well that there are a number of alternative approaches to this question, and notably the Voiculescu one, using the free probability theory explained in chapter 8 above, the idea being that of recapturing the number $N \in \mathbb{N}$ from the knowledge of the von Neumann algebra $L(F_N)$, via an entropy-type invariant.

This latter program, while not solving the original problem, due to serious technical difficulties, is however very successful, in the sense that it has led to a lot of interesting results and computations, in relation with a lot of mathematics and physics.

Is the free group factor problem something belonging to logic, as the difficult problems in functional analysis end up being? No one really knows the answer here.

12e. Exercises

Things have been extremely technical in this chapter, which was more of a survey than something else, and as a unique exercise on all this, we have:

EXERCISE 12.29. Learn some more basic hyperfinite von Neumann algebra theory, from the papers of von Neumann and Murray-von Neumann, then Connes, Haagerup and others, and write down a brief account of what you learned.

In what follows we will avoid ourselves this type of exercise, basically by getting back to the material in chapter 10, and building on that, following Jones.

Part IV

Subfactor theory

*Maria, you've gotta see her
Go insane and out of your mind
Latina, Ave Maria
A million and one candle lights*

CHAPTER 13

Subfactor theory

13a. The Jones tower

In the remainder of this book we discuss the basics of Jones' subfactor theory. The subfactors are quite subtle objects, of very general type, whose understanding requires both a good knowledge of advanced functional analysis, along the lines of chapters 9-12 above, and of more standard mathematics as well, such as combinatorics, geometry and topology, algebra and group theory of all kinds, probability, and so on.

We recall that a II_1 factor is a von Neumann algebra $A \subset B(H)$ which has trivial center, $Z(A) = \mathbb{C}$, is infinite dimensional, and has a trace $tr : A \rightarrow \mathbb{C}$. For a number of reasons, ranging from simple and intuitive to fairly advanced, explained in chapters 9-12 above, such algebras are the core at the whole von Neumann algebra theory.

The world of II_1 factors is a bit similar to the world of the usual matrix algebras $M_N(\mathbb{C})$, which are actually called type I factors, in the sense that it is "self-sufficient", with no need to go further than that. In particular, a nice representation theory for such II_1 factors can be obtained by staying inside the class of II_1 factors, and we have the following definition to start with, which will keep us busy for the rest of this book:

DEFINITION 13.1. *A subfactor is an inclusion of type*

$$A \subset B$$

with both A, B being II_1 factors.

We will see later some examples of such inclusions, along with motivations for their study. In order to get started now, the first thing to be done with such an inclusion is that of defining its index, as a quantity of the following type:

$$[B : A] = \dim_A B$$

Since both A, B are infinite dimensional complex algebras, this is not exactly obvious. In addition, in view of our previous experience with the II_1 factors, from chapters 9-12 above, and notably with their "continuous dimension" features, we can only expect the index to be a quite tricky quantity, ranging continuously, as follows:

$$[B : A] \in [1, \infty]$$

In order to discuss this, let us recall from chapter 10 above that given a representation of a II_1 factor $A \subset B(H)$, we can construct a number as follows, called coupling constant, which for the standard form, where $H = L^2(A)$, takes the value 1, and which in general measures how far is $A \subset B(H)$ from the standard form:

$$\dim_A H \in (0, \infty]$$

Getting now to the subfactors, in the sense of Definition 13.1, we have the following construction, that we know as well from chapter 10 above:

THEOREM 13.2. *Given an inclusion of II_1 factors $A \subset B$, the number*

$$N = \frac{\dim_A H}{\dim_B H}$$

is independent of the ambient Hilbert space H , and is called index.

PROOF. This is something that we know from chapter 10, the idea being that the independence of the index from the choice of the ambient Hilbert space H comes from the various basic properties of the coupling constant. \square

There are many examples of subfactors, and we will discuss this gradually, in what follows. Following Jones, the most basic examples of subfactors are as follows:

PROPOSITION 13.3. *Assuming that G is a compact group, acting on a II_1 factor P in a minimal way, in the sense that we have*

$$(P^G)' \cap P = \mathbb{C}$$

and that $H \subset G$ is a closed subgroup of finite index, we have a subfactor

$$P^G \subset P^H$$

having index $N = [G : H]$, called Jones subfactor.

PROOF. This is something standard, the idea being that the factoriality of P^G, P^H comes from the minimality of the action, and that the index formula is clear. We will be back with full details about this in a moment, directly in a more general setting. \square

In order to study the subfactors, let us start with the following standard result:

PROPOSITION 13.4. *Given a subfactor $A \subset B$, there is a unique linear map*

$$E : B \rightarrow A$$

which is positive, unital, trace-preserving and satisfies the following condition:

$$E(b_1 a b_2) = b_1 E(a) b_2$$

This map is called conditional expectation from B onto A .

PROOF. We make use of the standard representation of the II_1 factor B , with respect to its unique trace $tr : B \rightarrow \mathbb{C}$, as constructed in chapter 10 above:

$$B \subset L^2(B)$$

If we denote by Ω the standard cyclic and separating vector of $L^2(B)$, we have an identification $A\Omega = L^2(A)$. Consider now the following orthogonal projection:

$$e : L^2(B) \rightarrow L^2(A)$$

It follows from definitions that we have an inclusion as follows:

$$e(B\Omega) \subset A\Omega$$

Thus e induces by restriction a certain linear map $E : B \rightarrow A$. This linear map E and the orthogonal projection e are then related by:

$$exe = E(x)e$$

But this shows that the linear map E satisfies the various conditions in the statement, namely positivity, unitality, trace preservation and bimodule property. As for the uniqueness assertion, this follows by using the same argument, applied backwards, the idea being that a map E as in the statement must come from the projection e . \square

Following Jones [43], we will be interested in what follows in the orthogonal projection $e : L^2(B) \rightarrow L^2(A)$ producing the expectation $E : B \rightarrow A$, rather than in E itself:

DEFINITION 13.5. *Associated to any subfactor $A \subset B$ is the orthogonal projection*

$$e : L^2(B) \rightarrow L^2(A)$$

producing the conditional expectation $E : B \rightarrow A$ via the following formula:

$$exe = E(x)e$$

This projection is called Jones projection for the subfactor $A \subset B$.

Quite remarkably, the subfactor $A \subset B$, as well as its commutant, can be recovered from the knowledge of this projection, in the following way:

PROPOSITION 13.6. *Given a subfactor $A \subset B$, with Jones projection e , we have*

$$A = B \cap \{e\}'$$

$$A' = (B' \cap \{e\})''$$

as equalities of von Neumann algebras, acting on the space $L^2(B)$.

PROOF. These formulae basically follow from $exe = E(x)e$, as follows:

(1) Let us first prove that we have $A \subset B \cap \{e\}'$. Given $x \in A$, we have:

$$xe = E(x)e = exe$$

$$x^*e = E(x^*)e = ex^*e$$

Thus, we obtain, as desired, that x commutes with e :

$$ex = (x^*e)^* = (ex^*e)^* = exe = xe$$

(2) Let us prove now that $B \cap \{e\}' \subset A$. Assuming $ex = xe$, we have:

$$E(x)e = exe = xe^2 = xe$$

We conclude from this that we have the following equality:

$$(E(x) - x)\Omega = (E(x) - x)e\Omega = 0$$

Now since Ω is separating for B we have, as desired:

$$x = E(x) \in A$$

(3) In order to prove now $A' = \langle B', e \rangle$, observe that we have:

$$A = B \cap \{e\}' = B'' \cap \{e\}' = (B' \cap \{e\})'$$

Now by taking the commutant, we obtain $A' = (B' \cap \{e\})''$, as desired. \square

We are now ready to formulate a key definition, as follows:

DEFINITION 13.7. *Associated to any subfactor $A \subset B$ is the basic construction*

$$A \subset_e B \subset C$$

with $C = \langle B, e \rangle$ being the algebra generated by B and by the Jones projection

$$e : L^2(B) \rightarrow L^2(A)$$

acting on the Hilbert space $L^2(B)$.

The idea in what follows, going back to [43], will be that $B \subset C$ appears as a kind of “reflection” of $A \subset B$, and also that the basic construction can be iterated, with all this leading to nontrivial results. Let us start by further studying the basic construction:

THEOREM 13.8. *Given a subfactor $A \subset B$ having finite index,*

$$[B : A] < \infty$$

the basic construction $A \subset_e B \subset C$ has the following properties:

- (1) $C = JA'J$.
- (2) $C = \overline{B + Be\bar{b}}$.
- (3) C is a II_1 factor.
- (4) $[C : B] = [B : A]$.
- (5) $eCe = Ae$.
- (6) $\text{tr}(e) = [B : A]^{-1}$.
- (7) $\text{tr}(xe) = \text{tr}(x)[B : A]^{-1}$, for any $x \in B$.

PROOF. All this is standard, the idea being as follows:

(1) We have $JB'J = B$ and $JeJ = e$, which gives:

$$\begin{aligned} JA'J &= J \langle B', e \rangle J \\ &= \langle JB'J, JeJ \rangle \\ &= \langle B, e \rangle \\ &= C \end{aligned}$$

(2) This follows from the fact that the vector space $B + BeB$ is closed under multiplication, and from the fact that we have $exe = E(x)e$.

(3) This follows from the fact, that we know from chapter 10 above, that our finite index assumption $[B : A] < \infty$ is equivalent to the fact that A' is a factor. But this in turn is equivalent to the fact that $C = JA'J$ is a factor, as desired.

(4) We have indeed the following computation:

$$\begin{aligned} [C : B] &= \frac{\dim_B L^2(B)}{\dim_C L^2(B)} \\ &= \frac{1}{\dim_C L^2(B)} \\ &= \frac{1}{\dim_{JA'J} L^2(B)} \\ &= \frac{1}{\dim_{A'} L^2(B)} \\ &= \dim_A L^2(B) \\ &= [B : A] \end{aligned}$$

(5) This follows indeed from (2) and from the formula $exe = E(x)e$.

(6) We have the following computation:

$$\begin{aligned} 1 &= \dim_A L^2(A) \\ &= \dim_A(eL^2(B)) \\ &= \text{tr}_{A'}(e) \dim_A(L^2(B)) \\ &= \text{tr}_{A'}(a)[B : A] \end{aligned}$$

Now since $C = JA'J$ and $JeJ = e$, we obtain from this, as desired:

$$\begin{aligned} \text{tr}(e) &= \text{tr}_{JA'J}(JeJ) \\ &= \text{tr}_{A'}(e) \\ &= [B : A]^{-1} \end{aligned}$$

(7) We already know from (6) that the formula in the statement holds for $x = 1$. In order to discuss the general case, observe first that for $x, y \in A$ we have:

$$\operatorname{tr}(xye) = \operatorname{tr}(yex) = \operatorname{tr}(yxex)$$

Thus the linear map $x \rightarrow \operatorname{tr}(xe)$ is a trace on A , and by uniqueness of the trace on A , we must have, for a certain constant $c > 0$:

$$\operatorname{tr}(xe) = c \cdot \operatorname{tr}(x)$$

Now by using (6) we obtain $c = [B : A]^{-1}$, so we have proved the formula in the statement for $x \in A$. The passage to the general case $x \in B$ can be done as follows:

$$\begin{aligned} \operatorname{tr}(xe) &= \operatorname{tr}(exe) \\ &= \operatorname{tr}(E(x)e) \\ &= \operatorname{tr}(E(x))c \\ &= \operatorname{tr}(x)c \end{aligned}$$

Thus, we have proved the formula in the statement, in general. \square

The above result is quite interesting, so let us perform now twice the basic construction, and see what we get. The result here, which is more technical, is as follows:

PROPOSITION 13.9. *Associated to $A \subset B$ is the double basic construction*

$$A \subset_e B \subset_f C \subset D$$

with e, f being the following orthogonal projections,

$$e : L^2(B) \rightarrow L^2(A)$$

$$f : L^2(C) \rightarrow L^2(B)$$

having the following properties:

$$fef = [B : A]^{-1}f$$

$$efe = [B : A]^{-1}e$$

PROOF. We have two formulae to be proved, the idea being as follows:

(1) The first formula is clear, because we have:

$$\begin{aligned} fef &= E(e)f \\ &= \operatorname{tr}(e)f \\ &= [B : A]^{-1}f \end{aligned}$$

(2) Regarding now the second formula, it is enough to check it on the dense subset $(B + BeB)\Omega$. Thus, we must show that for any $x, y, z \in B$, we have:

$$efe(x + yez)\Omega = [B : A]^{-1}e(x + yez)\Omega$$

For this purpose, we will prove that we have, for any $x, y, z \in B$:

$$efex\Omega = [B : A]^{-1}ex\Omega$$

$$efeyez\Omega = [B : A]^{-1}eyez\Omega$$

The first formula can be established as follows:

$$\begin{aligned} efex\Omega &= efexf\Omega \\ &= eE(ex)f\Omega \\ &= eE(e)xf\Omega \\ &= [B : A]^{-1}exf\Omega \\ &= [B : A]^{-1}ex\Omega \end{aligned}$$

The second formula can be established as follows:

$$\begin{aligned} efeyez\Omega &= efeyezf\Omega \\ &= eE(eyez)f\Omega \\ &= eE(eye)zf\Omega \\ &= eE(E(y)e)zf\Omega \\ &= eE(y)E(e)zf\Omega \\ &= [B : A]^{-1}eE(y)zf\Omega \\ &= [B : A]^{-1}eyezf\Omega \\ &= [B : A]^{-1}eyez\Omega \end{aligned}$$

Thus, we are led to the conclusion in the statement. □

We can in fact perform the basic construction by recurrence, and we obtain:

THEOREM 13.10. *Associated to any subfactor $A_0 \subset A_1$ is the Jones tower*

$$A_0 \subset_{e_1} A_1 \subset_{e_2} A_2 \subset_{e_3} A_3 \subset \dots$$

with the Jones projections having the following properties:

- (1) $e_i^2 = e_i = e_i^*$.
- (2) $e_i e_j = e_j e_i$ for $|i - j| \geq 2$.
- (3) $e_i e_{i \pm 1} e_i = [B : A]^{-1} e_i$.
- (4) $tr(we_{n+1}) = [B : A]^{-1} tr(w)$, for any word $w \in \langle e_1, \dots, e_n \rangle$.

PROOF. This follows from Theorem 13.8 and Proposition 13.9 above, because the triple basic construction does not need in fact any further study. □

13b. Temperley-Lieb

The relations found in Theorem 13.10 are in fact well-known, from the standard theory of the Temperley-Lieb algebra. This algebra, discovered by Temperley and Lieb in the context of statistical mechanics [80], has a very simple definition, as follows:

DEFINITION 13.11. *The Temperley-Lieb algebra of index $N \in [1, \infty)$ is defined as*

$$TL_N(k) = \text{span}(NC_2(k, k))$$

with product given by vertical concatenation, with the rule

$$\bigcirc = N$$

for the closed circles that might appear when concatenating.

In other words, the algebra $TL_N(k)$, depending on parameters $k \in \mathbb{N}$ and $N \in [1, \infty)$, is the formal linear span of the pairings $\pi \in NC_2(k, k)$. The product operation is obtained by linearity, for the pairings which span $TL_N(k)$ this being the usual vertical concatenation, with the conventions that things go “from top to bottom”, and that each circle that might appear when concatenating is replaced by a scalar factor, equal to N .

In order to make the connection with subfactors, let us start with:

PROPOSITION 13.12. *The Temperley-Lieb algebra $TL_N(k)$ is generated by the diagrams*

$$\varepsilon_1 = \begin{array}{c} \cup \\ \cap \end{array}, \quad \varepsilon_2 = \begin{array}{c} \cup \\ | \\ \cap \end{array}, \quad \varepsilon_3 = \begin{array}{c} \cup \\ || \\ \cap \end{array}, \quad \dots$$

which are all multiples of projections, in the sense that their rescaled versions

$$e_i = N^{-1}\varepsilon_i$$

satisfy the abstract projection relations $e_i^2 = e_i = e_i^*$.

PROOF. We have two assertions here, the idea being as follows:

(1) The fact that the algebra $TL_N(k)$ is indeed generated by the sequence of diagrams $\varepsilon_1, \varepsilon_2, \varepsilon_3, \dots$ follows by drawing pictures, and more specifically by graphically decomposing each basis element $\pi \in NC_2(k, k)$ as a product of such elements ε_i .

(2) Regarding now the projection assertion, when composing ε_i with itself we obtain ε_i itself, times a circle. Thus, according to our multiplication conventions, we have:

$$\varepsilon_i^2 = N\varepsilon_i$$

Also, when turning upside-down ε_i , we obtain ε_i itself. Thus, according to our involution convention for the Temperley-Lieb algebra, we have:

$$\varepsilon_i^* = \varepsilon_i$$

We conclude that the rescalings $e_i = N^{-1}\varepsilon_i$ satisfy $e_i^2 = e_i = e_i^*$, as desired. □

As a second result now, making the link with Theorem 13.10, we have:

PROPOSITION 13.13. *The standard generators $e_i = N^{-1}\varepsilon_i$ of the Temperley-Lieb algebra $TL_N(k)$ have the following properties, where tr is the trace obtained by closing:*

- (1) $e_i e_j = e_j e_i$ for $|i - j| \geq 2$.
- (2) $e_i e_{i\pm 1} e_i = [B : A]^{-1} e_i$.
- (3) $tr(we_{n+1}) = [B : A]^{-1} tr(w)$, for any word $w \in \langle e_1, \dots, e_n \rangle$.

PROOF. This follows indeed by doing some elementary computations with diagrams, in the spirit of those performed in the proof of Proposition 13.12 above. Indeed:

- (1) This is clear from the definition of the diagrams ε_i .
- (2) This is clear as well from the definition of the diagrams ε_i .
- (3) This is something which is clear too, from the definition of ε_{n+1} . □

With the above results in hand, we can now reformulate our main finding about subfactors, namely Theorem 13.10 above, into something more conceptual, as follows:

THEOREM 13.14. *Given a finite index subfactor $A_0 \subset A_1$, with Jones tower*

$$A_0 \subset_{e_1} A_1 \subset_{e_2} A_2 \subset_{e_3} A_3 \subset \dots$$

the rescaled sequence of projections $e_1, e_2, e_3, \dots \in B(H)$ produces a representation

$$TL_N \subset B(H)$$

of the Temperley-Lieb algebra of index $N = [A_1 : A_0]$.

PROOF. The idea here is that Theorem 13.10, coming from the study of the basic construction, tells us that the rescaled sequence of projections $e_1, e_2, e_3, \dots \in B(H)$ behaves algebraically exactly as the sequence of diagrams $\varepsilon_1, \varepsilon_2, \varepsilon_3, \dots \in TL_N$ given by:

$$\varepsilon_1 = \cup \quad , \quad \varepsilon_2 = \mid \cup \quad , \quad \varepsilon_3 = \parallel \cup \quad , \quad \dots$$

But these diagrams generate TL_N , and so we have an embedding $TL_N \subset B(H)$, where H is the Hilbert space where our subfactor $A_0 \subset A_1$ lives, as claimed. □

Before going further, with some examples, more theory, and consequences of Theorem 13.14 above, let us make the following key observation, also from [43]:

THEOREM 13.15. *Given a finite index subfactor $A_0 \subset A_1$, the graded algebra*

$$P = (P_k)$$

formed by the sequence of higher relative commutants

$$P_k = A'_0 \cap A_k$$

contains the copy of the Temperley-Lieb algebra constructed above:

$$TL_N \subset P$$

This graded algebra $P = (P_k)$ is called “planar algebra” of the subfactor.

PROOF. As a first observation, since the Jones projection $e_1 : A_1 \rightarrow A_0$ commutes with A_0 , as was previously established in the above, we have:

$$e_1 \in P'_2$$

By translation we obtain from this that we have, for any $k \in \mathbb{N}$:

$$e_1, \dots, e_{k-1} \in P_k$$

Thus we have indeed an inclusion of graded algebras $TL_N \subset P$, as claimed. \square

The point with the above result, which explains among others the terminology at the end, is that, in the context of Theorem 13.14 above, the planar algebra structure of TL_N , obtained by composing diagrams, extends into an abstract planar algebra structure of P . See [47]. We will discuss all this, with full details, in the next chapter.

13c. Basic examples

Let us discuss now some basic examples of subfactors, with concrete illustrations for all the above notions, constructions, and general theory. These examples will all come from group actions $G \curvearrowright P$, which are assumed to be minimal, in the sense that:

$$(P^G)' \cap P = \mathbb{C}$$

As a starting point, we have the following result, that we know from the above:

PROPOSITION 13.16. *Assuming that G is a compact group, acting minimally on a II_1 factor P , and that $H \subset G$ is a subgroup of finite index, we have a subfactor*

$$P^G \subset P^H$$

having index $N = [G : H]$, called Jones subfactor.

PROOF. This is something that we already know, the idea being that the factoriality of P^G, P^H comes from the minimality of the action, and that the index formula is clear. \square

Along the same lines, we have the following result:

PROPOSITION 13.17. *Assuming that G is a finite group, acting minimally on a II_1 factor P , we have a subfactor as follows,*

$$P \subset P \rtimes G$$

having index $N = |G|$, called Ocneanu subfactor.

PROOF. This is standard as well, the idea being that the factoriality of $P \rtimes G$ comes from the minimality of the action, and that the index formula is clear. \square

We have as well a third result of the same type, as follows:

PROPOSITION 13.18. *Assuming that G is a compact group, acting minimally on a II_1 factor P , and that $G \rightarrow \text{PU}_n$ is a projective representation, we have a subfactor*

$$P^G \subset (M_n(\mathbb{C}) \otimes P)^G$$

having index $N = n^2$, called Wassermann subfactor.

PROOF. As before, the idea is that the factoriality of $P^G, (M_n(\mathbb{C}) \otimes P)^G$ comes from the minimality of the action, and the index formula is clear. \square

The above subfactors seem to be quite related, and indeed they are, due to:

THEOREM 13.19. *The Jones, Ocneanu and Wassermann subfactors are all of the same nature, and can be written as follows,*

$$\begin{aligned} (P^G \subset P^H) &\simeq ((\mathbb{C} \otimes P)^G \subset (l^\infty(G/H) \otimes P)^G) \\ (P \subset P \rtimes G) &\simeq ((l^\infty(G) \otimes P)^G \subset (\mathcal{L}(l^2(G)) \otimes P)^G) \\ (P^G \subset (M_n(\mathbb{C}) \otimes P)^G) &\simeq ((\mathbb{C} \otimes P)^G \subset (M_n(\mathbb{C}) \otimes P)^G) \end{aligned}$$

with standard identifications for the various tensor products and fixed point algebras.

PROOF. This is something very standard, modulo all kinds of standard identifications. We will explain all this more in detail later, after unifying these subfactors. \square

In order to unify now the above constructions of subfactors, the idea is quite clear. Given a compact group G , acting minimally on a II_1 factor P , and an inclusion of finite dimensional algebras $B_0 \subset B_1$, endowed as well with an action of G , we would like to construct a kind of generalized Wassermann subfactor, as follows:

$$(B_0 \otimes P)^G \subset (B_1 \otimes P)^G$$

In order to do this, we must talk first about the finite dimensional algebras B , and about inclusions of such algebras $B_0 \subset B_1$. Let us start with the following definition:

DEFINITION 13.20. *Associated to any finite dimensional algebra B is its canonical trace, obtained by composing the left regular representation with the trace of $\mathcal{L}(B)$:*

$$\text{tr} : B \subset \mathcal{L}(B) \rightarrow \mathbb{C}$$

We say that an inclusion of finite dimensional algebras $B_0 \subset B_1$ is Markov if it commutes with the canonical traces of B_0, B_1 .

In what regards the first notion, that of the canonical trace, this is something that we know well, from chapter 5 above. Indeed, as explained there, we can formally write $B = C(X)$, with X being a finite quantum space, and the canonical trace $\text{tr} : B \rightarrow \mathbb{C}$ is then precisely the integration with respect to the “counting measure” on X .

In what regards the second notion, that of a Markov inclusion, this is something very natural too, and as a first example here, any inclusion of type $\mathbb{C} \subset B$ is Markov. In

general, if we write $B_0 = C(X_0)$ and $B_1 = C(X_1)$, then the inclusion $B_0 \subset B_1$ must come from a certain fibration $X_1 \rightarrow X_0$, and the inclusion $B_0 \subset B_1$ is Markov precisely when the fibration $X_1 \rightarrow X_0$ commutes with the respective counting measures.

We will be back to Markov inclusions and their various properties on several occasions, in what follows. For our next purposes here, we just need the following result:

PROPOSITION 13.21. *Given a Markov inclusion of finite dimensional algebras $B_0 \subset B_1$ we can perform to it the basic construction, as to obtain a Jones tower*

$$B_0 \subset_{e_1} B_1 \subset_{e_2} B_2 \subset_{e_3} B_3 \subset \dots$$

exactly as we did in the above for the inclusions of II_1 factors.

PROOF. This is something quite routine, by following the computations in the above, from the case of the II_1 factors, and with everything extending well. It is of course possible to do something more general here, unifying the constructions for the inclusions of II_1 factors $A_0 \subset A_1$, and for the inclusions of Markov inclusions of finite dimensional algebras $B_0 \subset B_1$, but we will not need this degree of generality, in what follows. \square

With these ingredients in hand, getting back now to the Jones, Ocneanu and Wassermann subfactors, from Theorem 13.19 above, the point is that these constructions can be unified, and then further studied, the final result on the subject being as follows:

THEOREM 13.22. *Let G be a compact group, and $G \rightarrow \text{Aut}(P)$ be a minimal action on a II_1 factor. Consider a Markov inclusion of finite dimensional algebras*

$$B_0 \subset B_1$$

and let $G \rightarrow \text{Aut}(B_1)$ be an action which leaves invariant B_0 , and which is such that its restrictions to the centers of B_0 and B_1 are ergodic. We have then a subfactor

$$(B_0 \otimes P)^G \subset (B_1 \otimes P)^G$$

of index $N = [B_1 : B_0]$, called generalized Wassermann subfactor, whose Jones tower is

$$(B_1 \otimes P)^G \subset (B_2 \otimes P)^G \subset (B_3 \otimes P)^G \subset \dots$$

where $\{B_i\}_{i \geq 1}$ are the algebras in the Jones tower for $B_0 \subset B_1$, with the canonical actions of G coming from the action $G \rightarrow \text{Aut}(B_1)$, and whose planar algebra is given by:

$$P_k = (B'_0 \cap B_k)^G$$

These subfactors generalize the Jones, Ocneanu and Wassermann subfactors.

PROOF. There are several things to be proved, the idea being as follows:

(1) As before on various occasions, the idea is that the factoriality of the algebras $(B_i \otimes P)^G$ comes from the minimality of the action $G \rightarrow \text{Aut}(P)$, and that the index formula is clear as well, from the definition of the coupling constant and of the index.

(2) Regarding the Jones tower assertion, the precise thing to be checked here is that if $A \subset B \subset C$ is a basic construction, then so is the following sequence of inclusions:

$$(A \otimes P)^G \subset (B \otimes P)^G \subset (C \otimes P)^G$$

But this is something standard, which follows by verifying the basic construction conditions. We will be back to this in a moment, directly in a more general setting.

(3) Next, regarding the planar algebra assertion, we have to prove here that for any indices $i \leq j$, we have the following equality between subalgebras of $B_j \otimes P$:

$$((B_i \otimes P)^G)' \cap (B_j \otimes P)^G = (B_i' \cap B_j^G) \otimes 1$$

But this is something which is routine too, following Wassermann [93], and we will be back to this in a moment, with full details, directly in a more general setting.

(4) Finally, the last assertion, regarding the main examples of such subfactors, which are those of Jones, Ocneanu, Wassermann, follows from Theorem 13.19 above. \square

In addition to the Jones, Ocneanu and Wassermann subfactors, discussed and unified in the above, we have the Popa subfactors, which are constructed as follows:

PROPOSITION 13.23. *Given a discrete group $\Gamma = \langle g_1, \dots, g_n \rangle$, acting faithfully via outer automorphisms on a II_1 factor Q , we have the following “diagonal” subfactor*

$$\left\{ \left(\begin{array}{ccc} g_1(q) & & \\ & \ddots & \\ & & g_n(q) \end{array} \right) \Big| q \in Q \right\} \subset M_n(Q)$$

having index $N = n^2$, called Popa subfactor.

PROOF. This is something standard, a bit as for the Jones, Ocneanu and Wassermann subfactors, with the result basically coming from the work of Popa, who was the main user of such subfactors. We will come in a moment with a more general result in this direction, involving discrete quantum groups, along with a complete proof. \square

In order to unify now Theorem 13.22 and Proposition 13.23, observe that the diagonal subfactors can be written in the following way, by using a group dual:

$$(Q \rtimes \Gamma)^{\widehat{\Gamma}} \subset (M_n(\mathbb{C}) \otimes (Q \rtimes \Gamma))^{\widehat{\Gamma}}$$

Here the group dual $\widehat{\Gamma}$ acts on $P = Q \rtimes \Gamma$ via the dual of the action $\Gamma \subset \text{Aut}(Q)$, and on $M_n(\mathbb{C})$ via the adjoint action of the following representation:

$$\oplus g_i : \widehat{\Gamma} \rightarrow \mathbb{C}^n$$

Summarizing, we are led into quantum groups. Our plan in what follows will be that of discussing the quantum extension of Theorem 13.22, covering the diagonal subfactors as well, and this time with full details, and with examples and illustrations as well.

We follow [2], [3], [4], [6], where this extension of the Wassermann construction [93] was developed. Let us start our discussion with some basic theory. We first have:

DEFINITION 13.24. *A coaction of a Woronowicz algebra A on a finite von Neumann algebra P is an injective morphism $\Phi : P \rightarrow P \otimes A''$ satisfying the following conditions:*

- (1) *Coassociativity: $(\Phi \otimes id)\Phi = (id \otimes \Delta)\Phi$.*
- (2) *Trace equivariance: $(tr \otimes id)\Phi = tr(\cdot)1$.*
- (3) *Smoothness: $\overline{\mathcal{P}}^w = P$, where $\mathcal{P} = \Phi^{-1}(P \otimes_{alg} \mathcal{A})$.*

The above conditions come from what happens in the commutative case, $A = C(G)$, where they correspond to the usual associativity, trace equivariance and smoothness of the corresponding action $G \curvearrowright P$. Along the same lines, we have as well:

DEFINITION 13.25. *A coaction $\Phi : P \rightarrow P \otimes A''$ as above is called:*

- (1) *Ergodic, if the algebra $P^\Phi = \{p \in P \mid \Phi(p) = p \otimes 1\}$ reduces to \mathbb{C} .*
- (2) *Faithful, if the span of $\{(f \otimes id)\Phi(P) \mid f \in P_*\}$ is dense in A'' .*
- (3) *Minimal, if it is faithful, and satisfies $(P^\Phi)' \cap P = \mathbb{C}$.*

Observe that the minimality of the action implies in particular that the fixed point algebra P^Φ is a factor. Thus, we are getting here to the case that we are interested in, actions producing factors, via their fixed point algebras. More on this later.

In order to prove our subfactor results, we need of some general theory regarding the minimal actions. Following Wassermann [93], let us start with the following definition:

DEFINITION 13.26. *Let $\Phi : P \rightarrow P \otimes A''$ be a coaction. An eigenmatrix for a corepresentation $u \in B(H) \otimes A$ is an element $M \in B(H) \otimes P$ satisfying:*

$$(id \otimes \Phi)M = M_{12}u_{13}$$

A coaction is called semidual if each corepresentation has a unitary eigenmatrix.

As a basic example here, the canonical coaction $\Delta : A \rightarrow A \otimes A$ is semidual. We will prove in what follows, following the work of Wassermann in the usual compact group case, that the minimal coactions of Woronowicz algebras are semidual. We first have:

PROPOSITION 13.27. *If $\Phi : P \rightarrow P \otimes A''$ is a minimal coaction and $u \in Irr(A)$ is a corepresentation, then u has a unitary eigenmatrix precisely when $P^u \neq \{0\}$.*

PROOF. Given $u \in M_n(A)$, consider the following unitary corepresentation:

$$u^+ = (n \otimes 1) \oplus u = \begin{pmatrix} 1 & 0 \\ 0 & u \end{pmatrix} \in M_2(M_n(\mathbb{C}) \otimes \mathcal{A}) = M_2(\mathbb{C}) \otimes M_n(\mathbb{C}) \otimes \mathcal{A}$$

It is then routine to check, exactly as in [93], with the computation being explained in [2], that if the following algebra is a factor, then u has a unitary eigenmatrix:

$$X_u = (M_2(\mathbb{C}) \otimes M_n(\mathbb{C}) \otimes P)^{\pi_{u^+}}$$

So, let us prove that X_u is a factor. For this purpose, let $x \in Z(X_u)$. We have then $1 \otimes 1 \otimes P^\Phi \subset X_u$, and from the irreducibility of the inclusion $P^\pi \subset P$ we obtain that:

$$x \in M_2(\mathbb{C}) \otimes M_n(\mathbb{C}) \otimes 1$$

On the other hand, we have the following formula:

$$\begin{aligned} X_u \cap M_2(\mathbb{C}) \otimes M_n(\mathbb{C}) \otimes 1 &= (M_2(\mathbb{C}) \otimes M_n(\mathbb{C}))^{i_{u^+}} \otimes 1 \\ &= \text{End}(u^+) \otimes 1 \end{aligned}$$

Since our corepresentation u was chosen to be irreducible, it follows that x must be of the following form, with $y \in M_n(\mathbb{C})$, and with $\lambda \in \mathbb{C}$:

$$x = \begin{pmatrix} y & 0 \\ 0 & \lambda I \end{pmatrix} \otimes 1$$

Now let us pick a nonzero element $p \in P^u$, and write:

$$\Phi(p) = \sum_{ij} p_{ij} \otimes u_{ij}$$

Then $\Phi(p_{ij}) = \sum_k p_{kj} \otimes u_{ki}$ for any i, j , and so each column of $(p_{ij})_{ij}$ is a u -eigenvector. Choose such a nonzero column l and let m^i be the matrix having the i -th row equal to l , and being zero elsewhere. Then m^i is a u -eigenmatrix for any i , and this implies that:

$$\begin{pmatrix} 0 & m^i \\ 0 & 0 \end{pmatrix} \in X_u$$

The commutation relation of this matrix with x is as follows:

$$\begin{pmatrix} y & 0 \\ 0 & \lambda I \end{pmatrix} \begin{pmatrix} 0 & m^i \\ 0 & 0 \end{pmatrix} = \begin{pmatrix} 0 & m^i \\ 0 & 0 \end{pmatrix} \begin{pmatrix} y & 0 \\ 0 & \lambda I \end{pmatrix}$$

But this gives $(y - \lambda I)m^i = 0$. Now by definition of m^i , this shows that the i -th column of $y - \lambda I$ is zero. Thus $y - \lambda I = 0$, and so $x = \lambda 1$, as desired. \square

We can now prove a main result about minimal coactions, as follows:

THEOREM 13.28. *The minimal coactions are semidual.*

PROOF. Let K be the set of finite dimensional unitary corepresentations of A which have unitary eigenmatrices. Then, according to the above, the following happen:

(1) K is stable under taking tensor products. Indeed, if M, N are unitary eigenmatrices for u, w , then $M_{13}N_{23}$ is a unitary eigenmatrix for $u \otimes w$.

(2) K is stable under taking sums. Indeed, if M_i are unitary eigenmatrices for u_i , then $\text{diag}(M_i)$ is a unitary eigenmatrix for $\oplus u_i$.

(3) K is stable under substractions. Indeed, if M is an eigenmatrix for $U = \oplus_{i=1}^n u_i$, then the first $\dim(u_1)$ columns of M are formed by elements of P^{u_1} , the next $\dim(u_2)$ columns of M are formed by elements of P^{u_2} , and so on. Now if M is unitary, it is in

particular invertible, so all P^{u_i} are different from $\{0\}$, and we may conclude that we can indeed substract corepresentations from U , by using Proposition 13.27 above.

(4) K is stable under complex conjugation. Indeed, by the above results we may restrict attention to irreducible corepresentations. Now if $u \in \text{Irr}(A)$ has a nonzero eigenmatrix M then \overline{M} is an eigenmatrix for \bar{u} . By Proposition 13.27 above we obtain from this that $P^{\bar{u}} \neq \{0\}$, and we may conclude by using Proposition 13.27.

With this in hand, by using Peter-Weyl, we obtain the result. See [2]. \square

Let us construct now the fixed point subfactors. We first have:

PROPOSITION 13.29. *Consider a Woronowicz algebra $A = (A, \Delta, S)$, and denote by A_σ the Woronowicz algebra $(A, \sigma\Delta, S)$, where σ is the flip. Given coactions*

$$\begin{aligned}\beta &: B \rightarrow B \otimes A \\ \pi &: P \rightarrow P \otimes A_\sigma\end{aligned}$$

with B being finite dimensional, the following linear map, while not being multiplicative in general, is coassociative with respect to the comultiplication $\sigma\Delta$ of A_σ ,

$$\begin{aligned}\beta \odot \pi &: B \otimes P \rightarrow B \otimes P \otimes A_\sigma \\ b \otimes p &\rightarrow \pi(p)_{23}((\text{id} \otimes S)\beta(b))_{13}\end{aligned}$$

and its fixed point space, which is by definition the following linear space,

$$(B \otimes P)^{\beta \odot \pi} = \left\{ x \in B \otimes P \mid (\beta \odot \pi)x = x \otimes 1 \right\}$$

is then a von Neumann subalgebra of $B \otimes P$.

PROOF. This is something standard, which follows from a straightforward algebraic verification, explained in [2], [3]. As mentioned in the statement, to be noted is that the tensor product coaction $\beta \odot \pi$ is not multiplicative in general. See [2], [3]. \square

Our first task is to investigate the factoriality of such algebras, and we have here:

THEOREM 13.30. *If $\beta : B \rightarrow B \otimes A$ is a coaction and $\pi : P \rightarrow P \otimes A_\sigma$ is a minimal coaction, then the following conditions are equivalent:*

- (1) *The von Neumann algebra $(B \otimes P)^{\beta \odot \pi}$ is a factor.*
- (2) *The coaction β is centrally ergodic, $Z(B) \cap B^\beta = \mathbb{C}$.*

PROOF. This is something standard, from [2], [3], the idea being as follows:

(1) Our first claim, which is something whose proof is a routine verification, explained in [3], based on the semiduality of the minimal coaction π , that we know from Theorem 13.28 above, is that the following diagram is a non-degenerate commuting square:

$$\begin{array}{ccc} P & \subset & B \otimes P \\ \cup & & \cup \\ P^\pi & \subset & (B \otimes P)^{\beta \odot \pi} \end{array}$$

(2) In order to prove now the result, it is enough to check the following equality, between von Neumann subalgebras of the algebra $B \otimes P$:

$$Z((B \otimes P)^{\beta \circ \pi}) = (Z(B) \cap B^\beta) \otimes 1$$

So, let x be in the algebra on the left. Then x commutes with $1 \otimes P^\pi$, so it has to be of the form $b \otimes 1$. Thus x commutes with $1 \otimes P$. But x commutes with $(B \otimes P)^{\beta \circ \pi}$, and from the non-degeneracy of the above square, x commutes with $B \otimes P$, and in particular with $B \otimes 1$. Thus we have $b \in Z(B) \cap B^\beta$. As for the other inclusion, this is obvious. \square

In view of the above result, we can talk about subfactors of type $(B_0 \otimes P)^G \subset (B_1 \otimes P)^G$. In order to investigate such subfactors, we will need the following technical result:

PROPOSITION 13.31. *Consider two commuting squares, as follows:*

$$\begin{array}{ccccc} F & \subset & E & \subset & D \\ \cup & & \cup & & \cup \\ A & \subset & B & \subset & C \end{array}$$

- (1) *If the square on the left and the big square are non-degenerate, then so is the square on the right.*
- (2) *If both squares are non-degenerate, $F \subset E \subset D$ is a basic construction, and the Jones projection $e \in D$ for this basic construction belongs to C , then the square on the right is the basic construction for the square on the left.*

PROOF. We have several things to be proved, the idea being as follows:

- (1) This assertion is clear from the following computation:

$$D = \overline{sp}^w CF = \overline{sp}^w CBF = \overline{sp}^w CE$$

- (2) Let $\Psi : D \rightarrow C$ be the expectation. By non-degeneracy, we have that:

$$E = \overline{sp}^w FB = \overline{sp}^w BF$$

We also have $D = \overline{sp}^w EeE$ by the basic construction, so we get that:

$$\begin{aligned} C &= \Psi(D) \\ &= \Psi(\overline{sp}^w EeE) \\ &= \Psi(\overline{sp}^w BFeFB) \\ &= \Psi(\overline{sp}^w BeFB) \\ &= \overline{sp}^w Be\Psi(F)B \\ &= \overline{sp}^w BeAB \\ &= \overline{sp}^w BeB \end{aligned}$$

Thus the algebra C is generated by B and e , and this gives the result. \square

Next in line, we have the following key technical result:

PROPOSITION 13.32. *If $\beta : B \rightarrow B \otimes A$ is a coaction then*

$$\begin{array}{ccc} A & \subset & B \otimes A \\ \cup & & \uparrow \beta \\ \mathbb{C} & \subset & B \end{array}$$

is a non-degenerate commuting square.

PROOF. From the β -equivariance of the trace we get that the inclusion on the left commutes with the traces, so that the above is a commuting diagram of finite von Neumann algebras. From the formula of the expectation $E_\beta = (id \otimes \int_A) \beta$ we get that this diagram is a commuting square. Choose now an orthonormal basis $\{b_i\}$ of B , write $\beta : b_i \rightarrow \sum_j b_j \otimes u_{ji}$, and consider the corresponding unitary corepresentation:

$$u_\beta = \sum e_{ij} \otimes u_{ij}$$

Then for any k and any $a \in A$ we have the following computation:

$$\begin{aligned} \sum_i \beta(b_i)(1 \otimes u_{ki}^* a) &= \sum_{ij} b_j \otimes u_{ji} u_{ki}^* a \\ &= \sum_{ij} b_j \otimes \delta_{jk} a \\ &= b_k \otimes a \end{aligned}$$

Thus our commuting square is non-degenerate, as claimed. \square

Getting now to the generalized Wassermann subfactors, we first have:

PROPOSITION 13.33. *Given a Markov inclusion of finite dimensional algebras $B_0 \subset B_1$, construct its Jones tower, and denote it as follows:*

$$B_0 \subset B_1 \subset_{e_1} B_2 = \langle B_1, e_1 \rangle_{e_2} B_3 = \langle B_2, e_2 \rangle_{e_3} \dots$$

If $\beta_1 : B_1 \rightarrow B_1 \otimes A$ is a coaction/anticoaction leaving B_0 invariant then there exists a unique sequence $\{\beta_i\}_{i \geq 0}$ of coactions/anticoactions

$$\beta_i : B_i \rightarrow B_i \otimes A$$

such that each β_i extends β_{i-1} and leaves invariant the Jones projection e_{i-1} .

PROOF. By taking opposite inclusions we see that the assertion for anticoactions is equivalent to the one for coactions, that we will prove now. The uniqueness is clear from $B_i = \langle B_{i-1}, e_{i-1} \rangle$. For the existence, we can apply Proposition 13.32 to:

$$\begin{array}{ccccc} A & \subset & B_0 \otimes A & \subset & B_1 \otimes A \\ \cup & & \uparrow \beta_0 & & \uparrow \beta_1 \\ \mathbb{C} & \subset & B_0 & \subset & B_1 \end{array}$$

Indeed, we get in this way that the square on the right is a non-degenerate. Now by performing basic constructions to it, we get a sequence as follows:

$$\begin{array}{ccccccccc} B_0 \otimes A & \subset & B_1 \otimes A & \subset & B_2 \otimes A & \subset & B_3 \otimes A & \subset & \dots \\ \uparrow \beta_0 & & \uparrow \beta_1 & & \uparrow \beta_2 & & \uparrow \beta_3 & & \\ B_0 & \subset & B_1 & \subset & B_2 & \subset & B_3 & \subset & \dots \end{array}$$

It is easy to see from definitions that the β_i are coactions, that they extend each other, and that they leave invariant the Jones projections. But this gives the result. \square

With the above technical results in hand, we can now formulate our main theorem regarding the fixed point subfactors, of the most possible general type, as follows:

THEOREM 13.34. *Let G be a compact quantum group, and $G \rightarrow \text{Aut}(P)$ be a minimal action on a II_1 factor. Consider a Markov inclusion of finite dimensional algebras*

$$B_0 \subset B_1$$

and let $G \rightarrow \text{Aut}(B_1)$ be an action which leaves invariant B_0 and which is such that its restrictions to the centers of B_0 and B_1 are ergodic. We have then a subfactor

$$(B_0 \otimes P)^G \subset (B_1 \otimes P)^G$$

of index $N = [B_1 : B_0]$, called generalized Wassermann subfactor, whose Jones tower is

$$(B_1 \otimes P)^G \subset (B_2 \otimes P)^G \subset (B_3 \otimes P)^G \subset \dots$$

where $\{B_i\}_{i \geq 1}$ are the algebras in the Jones tower for $B_0 \subset B_1$, with the canonical actions of G coming from the action $G \rightarrow \text{Aut}(B_1)$, and whose planar algebra is given by:

$$P_k = (B'_0 \cap B_k)^G$$

These subfactors generalize the Jones, Ocneanu, Wassermann and Popa subfactors.

PROOF. We have several things to be proved, the idea being as follows:

(1) The first part of the statement, regarding the factoriality, the index and the Jones tower assertions, is something that follows exactly as in the classical group case.

(2) In order to prove now the planar algebra assertion, consider the following diagram, with $i < j$ being arbitrary integers:

$$\begin{array}{ccccc} P & \subset & B_i \otimes P & \subset & B_j \otimes P \\ \cup & & \cup & & \cup \\ P^\pi & \subset & (B_i \otimes P)^{\beta_i \otimes \pi} & \subset & (B_j \otimes P)^{\beta_j \otimes \pi} \end{array}$$

We know from Proposition 13.32 that the big square and the square on the left are both non-degenerate commuting squares. Thus Proposition 13.31 applies, and shows that the square on the right is a non-degenerate commuting square.

(3) Consider now the following sequence of non-degenerate commuting squares:

$$\begin{array}{ccccccc} B_0 \otimes P & \subset & B_1 \otimes P & \subset & B_2 \otimes P & \subset & \dots \\ \cup & & \cup & & \cup & & \\ (B_0 \otimes P)^{\beta_0 \otimes \pi} & \subset & (B_1 \otimes P)^{\beta_1 \otimes \pi} & \subset & (B_2 \otimes P)^{\beta_2 \otimes \pi} & \subset & \dots \end{array}$$

Since the Jones projections live in the lower line, Proposition 13.32 applies and shows that this is a sequence of basic constructions for non-degenerate commuting squares. In particular the lower line is a sequence of basic constructions, as desired.

(4) Finally, we already know from Theorem 13.22 above that our construction generalizes the Jones, Ocneanu and Wassermann subfactors. As for the Popa subfactors, the result here follows from the discussion made after Proposition 13.23 above. \square

13d. The index theorem

Let us go back now to the arbitrary subfactors, with Theorem 13.14 being our main result. As an interesting consequence of the above results, somehow contradicting the “continuous geometry” philosophy that has been going on so far, in relation with the II_1 factors, we have the following surprising result, also from Jones’ original paper [43]:

THEOREM 13.35. *The index of subfactors $A \subset B$ is “quantized” in the $[1, 4]$ range,*

$$N \in \left\{ 4 \cos^2 \left(\frac{\pi}{n} \right) \mid n \geq 3 \right\} \cup [4, \infty]$$

with the obstruction coming from the existence of the representation $TL_N \subset B(H)$.

PROOF. This comes from the basic construction, and more specifically from the combinatorics of the Jones projections e_1, e_2, e_3, \dots , the idea being as follows:

(1) In order to best comment on what happens, when iterating the basic construction, let us record the first few values of the numbers in the statement:

$$\begin{aligned} 4 \cos^2 \left(\frac{\pi}{3} \right) &= 1 \quad , \quad 4 \cos^2 \left(\frac{\pi}{4} \right) = 2 \\ 4 \cos^2 \left(\frac{\pi}{5} \right) &= \frac{3 + \sqrt{5}}{2} \quad , \quad 4 \cos^2 \left(\frac{\pi}{6} \right) = 3 \\ &\dots \end{aligned}$$

(2) When performing a basic construction, we obtain, by trace manipulations on e_1 :

$$N \notin (1, 2)$$

With a double basic construction, we obtain, by trace manipulations on $\langle e_1, e_2 \rangle$:

$$N \notin \left(2, \frac{3 + \sqrt{5}}{2} \right)$$

With a triple basic construction, we obtain, by trace manipulations on $\langle e_1, e_2, e_3 \rangle$:

$$N \notin \left(\frac{3 + \sqrt{5}}{2}, 3 \right)$$

Thus, we are led to the conclusion in the statement, by a kind of recurrence, involving a certain family of orthogonal polynomials.

(3) In practice now, the most elegant way of proving the result is by using the fundamental fact, explained in Theorem 13.14 above, that that sequence of Jones projections $e_1, e_2, e_3, \dots \subset B(H)$ generate a copy of the Temperley-Lieb algebra of index N :

$$TL_N \subset B(H)$$

With this result in hand, we must prove that such a representation cannot exist in index $N < 4$, unless we are in the following special situation:

$$N = 4 \cos^2 \left(\frac{\pi}{n} \right)$$

But this can be proved by using some suitable trace and positivity manipulations on TL_N , as in (2) above. For full details here, we refer [34], [35], [43], [52]. \square

The above result raises the question of understanding if there are further restrictions on the index of subfactors $A \subset B$, in the range found there, namely:

$$N \in \left\{ 4 \cos^2 \left(\frac{\pi}{n} \right) \mid n \geq 3 \right\} \cup [4, \infty]$$

This question is quite tricky, because it depends on the ambient factor $B \subset B(H)$, and also on the irreducibility assumption on the subfactor, namely $A' \cap B = \mathbb{C}$, which is something quite natural, and can be added to the problem.

All this is quite technical, to be discussed later on, when doing more advanced subfactor theory. In the simplest formulation of the question, the answer is generally “no”, as shown by the following result, also from Jones’ original paper [43]:

THEOREM 13.36. *Consider the Murray-von Neumann hyperfinite II_1 factor R . Its subfactors $R_0 \subset R$ are then as follows:*

- (1) *They exist for all admissible index values, $N \in \{4 \cos^2(\frac{\pi}{n}) \mid n \geq 3\} \cup [4, \infty]$.*
- (2) *In index $N \leq 4$, they can be realized as irreducible subfactors, $R'_0 \cap R = \mathbb{C}$.*
- (3) *In index $N > 4$, they can be realized as arbitrary subfactors.*

PROOF. This is something quite tricky, worked out in Jones’ original paper [43], and requiring some advanced algebra methods, the idea being as follows:

(1) This basically follows by taking a copy of the Temperley-Lieb algebra TL_N , and then building a subfactor out of it, first by constructing a certain inclusion of inductive

limits of finite dimensional algebras, $\mathcal{A} \subset \mathcal{B}$, and then by taking the weak closure, which produces copies of the Murray-von Neumann hyperfinite II_1 factor, $A \simeq B \simeq R$.

(2) This follows by examining and fine-tuning the construction in (1), which can be performed as to have control over the relative commutant.

(3) This follows as well from (1), and with the simplest proof here being in fact quite simple, based on a projection trick. \square

As another application now, which is more theoretical, let us go back to the question of defining the index of a subfactor in a purely algebraic manner, which was open since chapter 10 above. The answer here, due to Pimsner and Popa, is as follows:

THEOREM 13.37. *Any finite index subfactor $A \subset B$ has an algebraic orthonormal basis, called Pimsner-Popa basis, which is constructed as follows:*

- (1) *In integer index, $N \in \mathbb{N}$, this is a usual basis, of type $\{b_1, \dots, b_N\}$, whose length is exactly the index.*
- (2) *In non-integer index, $N \notin \mathbb{N}$, this is something of type $\{b_1, \dots, b_n, c\}$, having length $n + 1$, with $n = [N]$, and with $N - n \in (0, 1)$ being related to c .*

PROOF. This is something quite technical, which follows from the basic theory of the basic construction. We refer here to the paper of Pimsner and Popa [68]. \square

13e. Exercises

There has been a lot of exciting theory in this chapter, leading us from functional analysis to concrete combinatorics, and as an exercise on all this, we have:

EXERCISE 13.38. *Clarify all the details for the Jones index theorem, stating that*

$$N \in \left\{ 4 \cos^2 \left(\frac{\pi}{n} \right) \mid n \geq 3 \right\} \cup [4, \infty]$$

with the obstruction coming from the existence of the representation $TL_N \subset B(H)$.

This is something that we already discussed in the above, but with a few details missing. Time to have this understood, along the above lines.

CHAPTER 14

Planar algebras

14a. Planar algebras

We have seen in the previous chapter the foundations of subfactor theory, and the main examples of subfactors, all having integer index. In this chapter we go into the core of the theory, with the idea in mind of axiomatizing the combinatorics of a given subfactor $A_0 \subset A_1$, via an object similar to the tensor categories for the quantum groups.

So, our starting point will be an arbitrary subfactor $A_0 \subset A_1$, assumed to have finite index, $[A_1 : A_0] < \infty$. Let us first review first what can be said about it, by using the Jones basic construction. We recall from chapter 13 that we have the following result:

THEOREM 14.1. *Given an inclusion of II_1 factors $A_0 \subset A_1$, with Jones tower*

$$A_0 \subset_{e_1} A_1 \subset_{e_2} A_2 \subset_{e_3} A_3 \subset \dots$$

the sequence of projections $e_1, e_2, e_3, \dots \in B(H)$ produces a representation

$$TL_N \subset B(H)$$

of the Temperley-Lieb algebra of index $N = [A_1 : A_0]$. Moreover, we have

$$TL_N \subset P$$

where $P = (P_k)$ is the graded algebra formed by the commutants $P_k = A'_0 \cap A_k$.

PROOF. There are two statements here, that we both know from chapter 13 above, the idea for this, in short, being as follows:

(1) A detailed study of the basic construction, performed in chapter 13, shows that the rescaled sequence of Jones projections $e_1, e_2, e_3, \dots \in B(H)$ behaves algebraically exactly as the sequence of standard generators $\varepsilon_1, \varepsilon_2, \varepsilon_3, \dots \in TL_N$. Thus we have an embedding $TL_N \subset B(H)$, where H is the Hilbert space where our subfactor $A_0 \subset A_1$ lives.

(2) Once again by carefully looking at the Jones basic construction, the more precise conclusion is that the image of the representation $TL_N \subset B(H)$ constructed above lives indeed in the graded algebra $P = (P_k)$ formed by the commutants $P_k = A'_0 \cap A_k$. \square

Quite remarkably, the planar algebra structure of TL_N , taken in an intuitive sense, that of composing planar diagrams, extends to a planar algebra structure on P .

In order to discuss this key result, let us start with the axioms for the planar algebras. Following Jones' paper [47], we have the following definition:

DEFINITION 14.2. *The planar algebras are defined as follows:*

- (1) *We consider rectangles in the plane, with the sides parallel to the coordinate axes, and taken up to planar isotopy, and we call such rectangles boxes.*
- (2) *A labelled box is a box with $2k$ marked points on its boundary, k on its upper side, and k on its lower side, for some integer $k \in \mathbb{N}$.*
- (3) *A tangle is labelled box, containing a number of labelled boxes, with all marked points, on the big and small boxes, being connected by noncrossing strings.*
- (4) *A planar algebra is a sequence of finite dimensional vector spaces $P = (P_k)$, together with linear maps $P_{k_1} \otimes \dots \otimes P_{k_r} \rightarrow P_k$, one for each tangle, such that the gluing of tangles corresponds to the composition of linear maps.*

In this definition we are using rectangles, but everything being up to isotopy, we could have used instead circles with marked points, as in [47], which is the same thing. Our choice for using rectangles comes from the main examples that we have in mind, to be discussed below, where the planar algebra structure is best viewed by using rectangles.

Let us also mention that Definition 14.2 above is something quite simplified, based on [47]. As explained in [47], in order for subfactors to produce planar algebras and vice versa, there are quite a number of supplementary axioms that must be added, and in view of this, it is perhaps better to start with something stronger than Definition 14.2, as basic axioms. However, as before with rectangles vs circles, our axiomatic choices here are mainly motivated by the concrete examples that we have in mind.

As a basic example of a planar algebra, we have the Temperley-Lieb algebra:

THEOREM 14.3. *The Temperley-Lieb algebra TL_N , viewed as sequence of finite dimensional vector spaces*

$$TL_N = (TL_N(k))_{k \in \mathbb{N}}$$

is a planar algebra in the above sense, with the corresponding linear maps associated to the planar tangles

$$TL_N(k_1) \otimes \dots \otimes TL_N(k_r) \rightarrow TL_N(k)$$

appearing by putting the various $TL_N(k_i)$ diagrams into the small boxes of the given tangle, which produces a $TL_N(k)$ diagram.

PROOF. This is something trivial, which follows from definitions:

- (1) Assume indeed that we are given a planar tangle π in the sense of Definition 14.2, consisting of a box having $2k$ marked points on its boundary, and containing r small boxes, having respectively $2k_1, \dots, 2k_r$ marked points on their boundaries, and then a total of $k + \sum k_i$ noncrossing strings, connecting the various $2k + \sum 2k_i$ marked points.

(2) We want to associate to this planar tangle π a linear map as follows:

$$T_\pi : TL_N(k_1) \otimes \dots \otimes TL_N(k_r) \rightarrow TL_N(k)$$

For this purpose, by linearity, it is enough to construct elements as follows, for any choice of Temperley-Lieb diagrams $\sigma_i \in TL_N(k_i)$, with $i = 1, \dots, r$:

$$T_\pi(\sigma_1 \otimes \dots \otimes \sigma_r) \in TL_N(k)$$

(3) But constructing such an element is obvious, simply by putting the various diagrams $\sigma_i \in TL_N(k_i)$ into the small boxes the given tangle π . Indeed, this procedure produces a certain diagram in $TL_N(k)$, that we can call $T_\pi(\sigma_1 \otimes \dots \otimes \sigma_r)$, as above.

(4) Finally, we still have to check that everything is well-defined up to planar isotopy, and that the gluing of tangles corresponds to the composition of linear maps. But both these checks are trivial, coming from the definition of TL_N , and we are done. \square

As a conclusion to the above, $P = TL_N$ is indeed a planar algebra, and of somewhat “trivial” type, with the triviality coming from the fact that, in this case, the elements of P are planar diagrams themselves, and so the planar structure appears trivially.

The Temperley-Lieb planar algebra TL_N is an important planar algebra, because it is the “smallest” one, appearing inside the planar algebra of any subfactor. But more on this later, when talking about the relation between planar algebras and subfactors.

Moving ahead now, here is our second basic example of a planar algebra, which is still “trivial” in the above sense, with the elements of the planar algebra being planar diagrams themselves, but which appears in a bit more complicated way:

THEOREM 14.4. *The Fuss-Catalan algebra $FC_{N,M}$, which appears by coloring the Temperley-Lieb diagrams with black/white colors, clockwise, as follows*



and keeping those diagrams whose strings connect either $\circ - \circ$ or $\bullet - \bullet$, is a planar algebra, with again the corresponding linear maps associated to the planar tangles

$$FC_{N,M}(k_1) \otimes \dots \otimes FC_{N,M}(k_r) \rightarrow FC_{N,M}(k)$$

appearing by putting the various $FC_{N,M}(k_i)$ diagrams into the small boxes of the given tangle, which produces a $FC_{N,M}(k)$ diagram.

PROOF. The proof here is nearly identical to the proof of Theorem 14.3 above, with the only change appearing at the level of the colors. To be more precise:

(1) Forgetting about upper and lower sequences of points, which must be joined by strings, a Temperley-Lieb diagram can be thought of as being just a collection of strings, say black strings, which compose in the obvious way, with the rule that the value of the

circle, which is now a black circle, is N . And it is this obvious composition rule that gives the planar algebra structure, as explained in the proof of Theorem 14.3 above.

(2) Similarly, forgetting about sequences of points, a Fuss-Catalan diagram can be thought of as being a collection of strings, which come now in two colors, black and white. These Fuss-Catalan diagrams compose then in the obvious way, with the rule that the value of the black circle is N , and the value of the white circle is M . And it is this obvious composition rule that gives the planar algebra structure, as before for TL_N . \square

The same comments as those for TL_N apply. On one hand, $FC_{N,M}$ is by definition a “trivial” planar algebra, with the triviality coming from the fact that its elements are planar diagrams themselves. On the other hand, $FC_{N,M}$ is an important planar algebra, because it can be shown to appear inside the planar algebra of any subfactor $A \subset B$, assuming that an intermediate subfactor $A \subset C \subset B$ is present. But more on this later, when talking about the relation between planar algebras and subfactors.

Getting back now to generalities, and to Definition 14.2 above, that of a general planar algebra, we have so far two illustrations for it, which, while both important, are both “trivial”, with the planar structure simply coming from the fact that, in both these illustrations, the elements of the planar algebra are planar diagrams themselves.

In general, the planar algebras are more complicated than this, and we will see some further examples in a moment. However, the idea is very simple, namely “the elements of a planar algebra are not necessarily diagrams, but they behave like diagrams”.

Let us begin with the construction of the tensor planar algebra \mathcal{T}_N , which is the third most important planar algebra, in our series of examples. This algebra is as follows:

DEFINITION 14.5. *The tensor planar algebra \mathcal{T}_N is the sequence of vector spaces*

$$P_k = M_N(\mathbb{C})^{\otimes k}$$

with the multilinear maps associated to the various k -tangles

$$T_\pi : P_{k_1} \otimes \dots \otimes P_{k_r} \rightarrow P_k$$

being given by the following formula, in multi-index notation,

$$T_\pi(e_{i_1} \otimes \dots \otimes e_{i_r}) = \sum_j \delta_\pi(i_1, \dots, i_r : j) e_j$$

with the Kronecker symbols δ_π being 1 if the indices fit, and being 0 otherwise.

In other words, we are using here a construction which is very similar to the construction $\pi \rightarrow T_\pi$ from easy quantum groups. We put the indices of the basic tensors on the marked points of the small boxes, in the obvious way, and the coefficients of the output tensor are then given by Kronecker symbols, exactly as in the easy case.

The fact that we have indeed a planar algebra, in the sense that the gluing of tangles corresponds to the composition of linear maps, as required by Definition 14.2, is something elementary, in the same spirit as the verification of the functoriality properties of the correspondence $\pi \rightarrow T_\pi$, discussed in chapter 8 above, and we refer here to [47].

Let us discuss now a second planar algebra of the same type, which is important as well for various reasons, namely the spin planar algebra \mathcal{S}_N . This planar algebra appears somewhat as a “square root” of the tensor planar algebra \mathcal{T}_N , and its construction is quite similar, but by using this time the algebra \mathbb{C}^N instead of the algebra $M_N(\mathbb{C})$.

There is one subtlety, however, coming from the fact that the general planar algebra formalism, from Definition 14.2 above, requires the tensors to have even length. Note that this was automatic for \mathcal{T}_N , where the tensors of $M_N(\mathbb{C})$ have length 2.

In the case of the spin planar algebra \mathcal{S}_N , we want the vector spaces to be:

$$P_k = (\mathbb{C}^N)^{\otimes k}$$

Thus, we must double the indices of the tensors, in the following way:

DEFINITION 14.6. *We write the standard basis of $(\mathbb{C}^N)^{\otimes k}$ in $2 \times k$ matrix form,*

$$e_{i_1 \dots i_k} = \begin{pmatrix} i_1 & i_1 & i_2 & i_2 & i_3 & \dots & \dots \\ i_k & i_k & i_{k-1} & \dots & \dots & \dots & \dots \end{pmatrix}$$

by duplicating the indices, and then writing them clockwise, starting from top left.

Now with this convention in hand for the tensors, we can formulate the construction of the spin planar algebra \mathcal{S}_N , also from [47], as follows:

DEFINITION 14.7. *The spin planar algebra \mathcal{S}_N is the sequence of vector spaces*

$$P_k = (\mathbb{C}^N)^{\otimes k}$$

written as above, with the multilinear maps associated to the various k -tangles

$$T_\pi : P_{k_1} \otimes \dots \otimes P_{k_r} \rightarrow P_k$$

being given by the following formula, in multi-index notation,

$$T_\pi(e_{i_1} \otimes \dots \otimes e_{i_r}) = \sum_j \delta_\pi(i_1, \dots, i_r : j) e_j$$

with the Kronecker symbols δ_π being 1 if the indices fit, and being 0 otherwise.

In other words, we are using exactly the same construction as for the tensor planar algebra \mathcal{T}_N , which was itself very related to the easy quantum group formalism, but with $M_N(\mathbb{C})$ replaced by \mathbb{C}^N , with the indices doubled, as in Definition 14.6.

As before with the tensor planar algebra \mathcal{T}_N , the fact that the spin planar algebra \mathcal{S}_N is indeed a planar algebra is something rather trivial, coming from definitions.

Observe however that, unlike our previous planar algebras TL_N and $FC_{N,M}$, which were “trivial” planar algebras, their elements being planar diagrams themselves, the planar algebras \mathcal{T}_N and \mathcal{S}_N are not trivial, their elements being not exactly planar diagrams.

Let us also mention that the algebras \mathcal{T}_N and \mathcal{S}_N are important for a number of reasons, in the context of the fixed point subfactors, to be discussed later on.

Getting back now to the planar algebra structure of \mathcal{T}_N and \mathcal{S}_N , which is something quite fundamental, worth being well understood, let us have here some more discussion. Generally speaking, the planar calculus for tensors is quite simple, and does not really require diagrams. Indeed, it suffices to imagine that the way various indices appear, travel around and disappear is by following some obvious strings connecting them.

Here are some illustrations for this principle, for the spin planar algebra \mathcal{S}_N :

EXAMPLE 14.8. *Identity, multiplication, inclusion.*

The identity 1_k is the (k, k) -tangle having vertical strings only. The solutions of $\delta_{1_k}(x, y) = 1$ being the pairs of the form (x, x) , this tangle 1_k acts by the identity:

$$1_k \begin{pmatrix} j_1 & \cdots & j_k \\ i_1 & \cdots & i_k \end{pmatrix} = \begin{pmatrix} j_1 & \cdots & j_k \\ i_1 & \cdots & i_k \end{pmatrix}$$

The multiplication M_k is the (k, k, k) -tangle having 2 input boxes, one on top of the other, and vertical strings only. It acts in the following way:

$$M_k \left(\begin{pmatrix} j_1 & \cdots & j_k \\ i_1 & \cdots & i_k \end{pmatrix} \otimes \begin{pmatrix} l_1 & \cdots & l_k \\ m_1 & \cdots & m_k \end{pmatrix} \right) = \delta_{j_1 m_1} \cdots \delta_{j_k m_k} \begin{pmatrix} l_1 & \cdots & l_k \\ i_1 & \cdots & i_k \end{pmatrix}$$

The inclusion I_k is the $(k, k + 1)$ -tangle which looks like 1_k , but has one more vertical string, at right of the input box. Given x , the solutions of $\delta_{I_k}(x, y) = 1$ are the elements y obtained from x by adding to the right a vector of the form $\binom{l}{i}$, and so:

$$I_k \begin{pmatrix} j_1 & \cdots & j_k \\ i_1 & \cdots & i_k \end{pmatrix} = \sum_l \begin{pmatrix} j_1 & \cdots & j_k & l \\ i_1 & \cdots & i_k & l \end{pmatrix}$$

Observe that I_k is an inclusion of algebras, and that the various I_k are compatible with each other. The union of the algebras $\mathcal{S}_N(k)$ is a graded algebra, denoted \mathcal{S}_N .

Along the same lines, some other important tangles are as follows:

EXAMPLE 14.9. *Expectation, Jones projection.*

The expectation U_k is the $(k+1, k)$ -tangle which looks like 1_k , but has one more string, connecting the extra 2 input points, both at right of the input box:

$$U_k \begin{pmatrix} j_1 & \cdots & j_k & j_{k+1} \\ i_1 & \cdots & i_k & i_{k+1} \end{pmatrix} = \delta_{i_{k+1}j_{k+1}} \begin{pmatrix} j_1 & \cdots & j_k \\ i_1 & \cdots & i_k \end{pmatrix}$$

Observe that U_k is a bimodule morphism with respect to I_k .

The Jones projection E_k is a $(0, k+2)$ -tangle, having no input box. There are k vertical strings joining the first k upper points to the first k lower points, counting from left to right. The remaining upper 2 points are connected by a semicircle, and the remaining lower 2 points are also connected by a semicircle. We have the following formula:

$$E_k(1) = \sum_{ijl} \begin{pmatrix} i_1 & \cdots & i_k & j & j \\ i_1 & \cdots & i_k & l & l \end{pmatrix}$$

The elements $e_k = N^{-1}E_k(1)$ are projections, and define a representation of the infinite Temperley-Lieb algebra of index N inside the inductive limit algebra \mathcal{S}_N .

Finally, again along the same lines, we have the following key tangle:

EXAMPLE 14.10. *Rotation.*

The rotation R_k is the (k, k) -tangle which looks like 1_k , but the first 2 input points are connected to the last 2 output points, and the same happens at right:

$$R_k = \begin{array}{c} \cap \quad | \quad | \quad | \\ \parallel \quad \quad \parallel \\ \parallel \quad | \quad | \quad \cup \end{array}$$

The action of R_k on the standard basis is by rotation of the indices, as follows:

$$R_k(e_{i_1 \dots i_k}) = e_{i_2 i_3 \dots i_k i_1}$$

Thus R_k acts by an order k linear automorphism of $\mathcal{S}_N(k)$, also called rotation.

There are many other interesting examples of k -tangles, but in view of our present purposes, we can actually stop here, due to the following key fact, which basically reduces everything to the study of the above particular tangles, and that we will use many times in what follows, for the various planar algebra results that we will prove:

THEOREM 14.11. *The multiplications, inclusions, expectations, Jones projections, and rotations generate the set of all tangles, via the gluing operation.*

PROOF. This is something elementary, obtained by “chopping” the various planar tangles into small pieces, belonging to the above list. See [47]. □

Finally, in order for things to be complete, we must talk as well about the $*$ -structure. Once again this is constructed as in the easy quantum group calculus, as follows:

$$\begin{pmatrix} j_1 & \cdots & j_k \\ i_1 & \cdots & i_k \end{pmatrix}^* = \begin{pmatrix} i_1 & \cdots & i_k \\ j_1 & \cdots & j_k \end{pmatrix}$$

Summarizing, the sequence of vector spaces $\mathcal{S}_N(k) = (\mathbb{C}^N)^{\otimes k}$ has a planar $*$ -algebra structure, called spin planar algebra of index $N = |X|$. See [47].

As a conclusion to all this, we have so far an abstract definition for the planar algebras, then two very basic examples, namely TL_N and $FC_{N,M}$, where the elements of the planar algebra are actual diagrams, composing as the diagrams do, by gluing, and then two examples which are slightly more complicated, namely \mathcal{T}_N and \mathcal{S}_N , where the planar algebra elements are tensors, composing according to the usual rules for the tensors.

14b. Higher commutants

In relation now with subfactors, the result, which extends Theorem 14.1 above, and which was found by Jones in [44], almost 20 years after [43], is as follows:

THEOREM 14.12. *Given a subfactor $A_0 \subset A_1$, the collection $P = (P_k)$ of linear spaces*

$$P_k = A'_0 \cap A_k$$

has a planar algebra structure, extending the planar algebra structure of TL_N .

PROOF. We know from Theorem 14.1 that we have an inclusion as follows, coming from the basic construction, and with TL_N itself being a planar algebra:

$$TL_N \subset P$$

Thus, the whole point is that of proving that the planar algebra structure of TL_N , obtained by composing diagrams, extends into a planar algebra structure of P . But this can be done via a long algebraic study, basically focusing on the basic tangles from Theorem 14.11 above, and with the whole thing, while being quite routine, being something quite long. For the full computation here, we refer to Jones' paper [47]. \square

As a first illustration for the above result, we have:

THEOREM 14.13. *We have the following universality results:*

- (1) *The Temperley-Lieb algebra TL_N appears inside the planar algebra of any subfactor $A \subset B$ having index N .*
- (2) *The Fuss-Catalan algebra $FC_{N,M}$ appears inside the planar algebra of any subfactor $A \subset B$, in the presence of an intermediate subfactor $A \subset C \subset B$.*

PROOF. Here the first assertion is something that we already know, from Theorem 14.1, and the second assertion is something quite standard as well, by carefully working out the basic construction for $A \subset B$, in the presence of an intermediate subfactor $A \subset C \subset B$. For details here, we refer to the paper of Bisch and Jones [17]. \square

It is possible to prove as well that the tensor planar algebra \mathcal{T}_N and the spin planar algebra \mathcal{S}_N have similar universality properties, but this time being the biggest possible instead of the smallest possible, in the framework of some suitable fixed point subfactors. We will discuss all this in a moment, in the general context of fixed point subfactors.

All the above results raise the question on whether any planar algebra produces a subfactor. The answer here is yes, but with many subtleties, as follows:

THEOREM 14.14. *We have the following results:*

- (1) *Any planar algebra with positivity produces a subfactor.*
- (2) *In particular, we have TL and FC type subfactors.*
- (3) *In the amenable case, and with $A_1 = R$, the correspondence is bijective.*
- (4) *In general, we must take $A_1 = L(F_\infty)$, and we do not have bijectivity.*
- (5) *The axiomatization of P , in the case $A_1 = R$, is not known.*

PROOF. All this is quite heavy, mainly coming from the work of Popa in the 90s, using heavy functional analysis, the idea being as follows:

(1) As already mentioned in the comments after Definition 14.2, our planar algebra axioms here are something quite simplified, based on [47]. However, by getting back to Theorem 14.12, and carefully looking at the planar algebras there, appearing from subfactors, the conclusion is that these subfactor planar algebras satisfy a number of supplementary “positivity” conditions, basically coming from the positivity of the II_1 factor trace. And the point now is that, with these positivity conditions axiomatized, we reach to something which is equivalent to Popa’s axiomatization of the lattice of higher relative commutants $A'_i \cap A_j$ of the finite index subfactors [72], obtained in the 90s via heavy functional analysis. For the story here, and details, we refer to Jones’ paper [47].

(2) We have been a bit quick in the above, and before anything, let us mention that our 4 main examples of planar algebras, namely TL_N and $FC_{N,M}$, and then \mathcal{T}_N and \mathcal{S}_N too, do satisfy the positivity requirements needed in (1). Thus, there are subfactors associated to all of them. In practice now, the existence of the TL_N subfactors, also known as “ A_∞ subfactors”, is something which was known for some time, since some early work of Popa on the subject. As for the existence of the $FC_{N,M}$ subfactors, this can be shown by using the intermediate subfactor picture, $A \subset C \subset B$, by composing two A_∞ subfactors of suitable indices, $A \subset C$ and $C \subset B$. For the story here, we refer to [17], [47].

(3) This is something fairly heavy, as it is always the case with operator algebra results about hyperfiniteness and amenability, due to Popa [70], [71].

(4) This is something more fashionable and recent, obtained by further building on the above-mentioned old constructions of Popa, and we refer here to [73], [36]

(5) This is the big open question in subfactors. The story here goes back to Jones' original paper [43], which contains at the end the question, due to Connes, of finding the possible values of the index for the irreducible subfactors of R . This question, which certainly looks much easier than (5) in the statement, is in fact still open, now 40 years after its formulation, and with on one having any valuable idea of dealing with it. \square

We refer to the original papers of Popa, and then to more recent papers by Jones, Popa and their collaborators for details on the above, which is quite heavy material.

14c. Fixed points

We discuss now the connection of all the above with the main examples of subfactors. We recall from chapter 13 above that the main examples of subfactors are all of integer index, and appear as fixed point subfactors, according to the following result:

THEOREM 14.15. *Let G be a compact quantum group, and $G \rightarrow \text{Aut}(P)$ be a minimal action on a II_1 factor. Consider a Markov inclusion of finite dimensional algebras*

$$B_0 \subset B_1$$

and let $G \rightarrow \text{Aut}(B_1)$ be an action which leaves invariant B_0 and which is such that its restrictions to the centers of B_0 and B_1 are ergodic. We have then a subfactor

$$(B_0 \otimes P)^G \subset (B_1 \otimes P)^G$$

of index $N = [B_1 : B_0]$, called generalized Wassermann subfactor, whose Jones tower is

$$(B_1 \otimes P)^G \subset (B_2 \otimes P)^G \subset (B_3 \otimes P)^G \subset \dots$$

where $\{B_i\}_{i \geq 1}$ are the algebras in the Jones tower for $B_0 \subset B_1$, with the canonical actions of G coming from the action $G \rightarrow \text{Aut}(B_1)$, and whose planar algebra is given by:

$$P_k = (B'_0 \cap B_k)^G$$

These subfactors generalize the Jones, Ocneanu, Wassermann and Popa subfactors.

PROOF. This is something that we know well from chapter 13, whose proof basically comes by generalizing, several times, the results of Wassermann in [93]. \square

In view of the above result, what we have to do in relation with such subfactors is to further interpret the last formula there, that of the planar algebra, namely:

$$P_k = (B'_0 \cap B_k)^G$$

To be more precise, we will show here that, under suitable assumptions on the original inclusion $B_0 \subset B_1$, we can associate a certain combinatorial planar algebra $P(B_0 \subset B_1)$

to this inclusion, and then the planar algebra associated to the fixed point subfactor itself appears as a fixed point subalgebra of this planar algebra, as follows:

$$P = P(B_0 \subset B_1)^G$$

More in detail, the idea will be that $P(B_0 \subset B_1)$ appears as a joint generalization of the spin and tensor planar algebras, discussed in the above, which appear as follows:

$$\mathcal{S}_N = P(\mathbb{C} \subset \mathbb{C}^N)$$

$$\mathcal{T}_N = P(\mathbb{C} \subset M_N(\mathbb{C}))$$

Thus, our first task will be that of getting back to the Markov inclusions $B_0 \subset B_1$, introduced in chapter 13 above, and further discuss the combinatorics of their basic construction, with planar algebra ideas in mind. As in chapter 13 above, it is most convenient to denote such inclusions by $A \subset B$, at least at a first stage of their study.

Following [35], which is the standard reference for such things, we first have:

DEFINITION 14.16. *Associated to an inclusion $A \subset B$ of finite dimensional algebras are the following objects:*

- (1) *The column vector $(a_i) \in \mathbb{N}^s$ given by $A = \bigoplus_{i=1}^s M_{a_i}(\mathbb{C})$.*
- (2) *The column vector $(b_j) \in \mathbb{N}^t$ given by $B = \bigoplus_{j=1}^t M_{b_j}(\mathbb{C})$.*
- (3) *The inclusion matrix $(m_{ij}) \in M_{s \times t}(\mathbb{N})$, satisfying $m^t a = b$.*

To be more precise here, in what regards the inclusion matrix, each minimal idempotent in $M_{a_i}(\mathbb{C}) \subset A$ splits, when regarded as an element of B , as a sum of minimal idempotents of B , and $m_{ij} \in \mathbb{N}$ is the number of such idempotents from $M_{b_j}(\mathbb{C})$.

We have the following result, bringing traces into picture:

PROPOSITION 14.17. *For an inclusion $A \subset B$, the following are equivalent:*

- (1) *$A \subset B$ commutes with the canonical traces.*
- (2) *We have $mb = ra$, where $r = \|b\|^2 / \|a\|^2$.*

PROOF. The weight vectors of the canonical traces of A, B are given by:

$$\tau_i = \frac{a_i^2}{\|a\|^2} \quad , \quad \tau_j = \frac{b_j^2}{\|b\|^2}$$

We can plug these values into the following standard compatibility formula:

$$\frac{\tau_i}{a_i} = \sum_j m_{ij} \cdot \frac{\tau_j}{b_j}$$

We obtain in this way the condition in the statement. □

We will need as well the following basic facts, also from [35]:

DEFINITION 14.18. Associated to an inclusion of finite dimensional algebras $A \subset B$, with inclusion matrix $m \in M_{s \times t}(\mathbb{N})$, are:

- (1) *The Bratteli diagram: this is the bipartite graph Γ having as vertices the sets $\{1, \dots, s\}$ and $\{1, \dots, t\}$, the number of edges between i, j being m_{ij} .*
- (2) *The basic construction: this is the inclusion $B \subset A_1$ obtained from $A \subset B$ by reflecting the Bratteli diagram.*
- (3) *The Jones tower: this is the tower of algebras $A \subset B \subset A_1 \subset B_1 \subset \dots$ obtained by iterating the basic construction.*

We know that for a Markov inclusion $A \subset B$ we have $m^t a = b$ and $mb = ra$, and so $mm^t a = ra$, which gives an eigenvector for the square matrix $mm^t \in M_s(\mathbb{N})$. When this latter matrix has positive entries, by Perron-Frobenius we obtain:

$$\|mm^t\| = r$$

This equality holds in fact without assumptions on m , and we have:

THEOREM 14.19. Let $A \subset B$ be Markov, with inclusion matrix $m \in M_{s \times t}(\mathbb{N})$.

- (1) $r = \dim(B)/\dim(A)$ is an integer.
- (2) $\|m\| = \|m^t\| = \sqrt{r}$.
- (3) $\|\dots mm^t mm^t \dots\| = r^{k/2}$, for any product of length k .

PROOF. Consider the vectors a, b , as in Definition 14.16. We know from definitions and from Proposition 14.17 that we have:

$$b = m^t a \quad , \quad mb = ra \quad , \quad r = \|b\|^2/\|a\|^2$$

- (1) If we construct as above the Jones tower for $A \subset B$, we have, for any k :

$$\frac{\dim B_k}{\dim A_k} = \frac{\dim A_k}{\dim B_{k-1}} = r$$

On the other hand, we have as well the following well-known formula:

$$\lim_{k \rightarrow \infty} (\dim A_k)^{1/2k} = \lim_{k \rightarrow \infty} (\dim B_k)^{1/2k} = \|mm^t\|$$

By combining these two formulae we obtain the following formula:

$$\|mm^t\| = r$$

But from $r \in \mathbb{Q}$ and $(mm^t)^k a = r^k a$ for any $k \in \mathbb{N}$, we get $r \in \mathbb{N}$, and we are done.

(2) This follows from the above equality $\|mm^t\| = r$, and from the following standard equalities, for any real rectangular matrix r :

$$\|m\|^2 = \|m^t\|^2 = \|mm^t\|$$

(3) Let n be the length k word in the statement. First, by applying the norm and by using the formula $\|m\| = \|m^t\| = \sqrt{r}$, we obtain the following inequality:

$$\|n\| \leq r^{k/2}$$

For the converse inequality, assume first that k is even. Then n has either a or b as eigenvector, depending on whether n begins with a m or with a m^t , in both cases with eigenvalue $r^{k/2}$, and this gives the desired inequality, namely:

$$\|n\| \geq r^{k/2}$$

Assume now that k is odd, and let $\circ \in \{1, t\}$ be such that $n' = m^\circ n$ is alternating. Since n' has even length, we already know that we have:

$$\|n'\| = r^{(k+1)/2}$$

On the other hand, we have as well the following estimate:

$$\|n'\| \leq \|m^\circ\| \cdot \|n\| = \sqrt{r}\|n\|$$

But this gives the following inequality:

$$\|n\| \geq r^{k/2}$$

Thus, we have obtained the desired reverse inequality, and we are done. □

The point now is that for a Markov inclusion, the basic construction and the Jones tower have a particularly simple form. Let us first work out the basic construction:

PROPOSITION 14.20. *The basic construction for a Markov inclusion $i : A \subset B$ of index $r \in \mathbb{N}$ is the inclusion $j : B \subset A_1$ obtained as follows:*

- (1) $A_1 = M_r(\mathbb{C}) \otimes A$, as an algebra.
- (2) $j : B \subset A_1$ is given by $mb = ra$.
- (3) $ji : A \subset A_1$ is given by $(mm^t)a = ra$.

PROOF. With notations from the above, the weight vector of the algebra A_1 appearing from the basic construction is $mb = ra$, and this gives the result. □

We fix a Markov inclusion $i : A \subset B$. We have the following result:

PROPOSITION 14.21. *The Jones tower associated to a Markov inclusion $i : A \subset B$, denoted as follows, with alternating letters,*

$$A \subset B \subset A_1 \subset B_1 \subset \dots$$

is given by the following formulae:

- (1) $A_k = M_r(\mathbb{C})^{\otimes k} \otimes A$.
- (2) $B_k = M_r(\mathbb{C})^{\otimes k} \otimes B$.
- (3) $A_k \subset B_k$ is $id_k \otimes i$.
- (4) $B_k \subset A_{k+1}$ is $id_k \otimes j$.

PROOF. This follows from Proposition 14.20, with the remark that if $i : A \subset B$ is Markov, then so is its basic construction $j : B \subset A_1$. □

Regarding now the relative commutants for this tower, we have here:

PROPOSITION 14.22. *The relative commutants for the Jones tower*

$$A \subset B \subset A_1 \subset B_1 \subset \dots$$

associated to a Markov inclusion $A \subset B$ are given by:

- (1) $A'_s \cap A_{s+k} = M_r(\mathbb{C})^{\otimes k} \otimes (A' \cap A)$.
- (2) $A'_s \cap B_{s+k} = M_r(\mathbb{C})^{\otimes k} \otimes (A' \cap B)$.
- (3) $B'_s \cap A_{s+k} = M_r(\mathbb{C})^{\otimes k} \otimes (B' \cap A)$.
- (4) $B'_s \cap B_{s+k} = M_r(\mathbb{C})^{\otimes k} \otimes (B' \cap B)$.

PROOF. The above assertions are all elementary, as follows:

(1,2) These assertions both follow from Proposition 14.21 above, and from the general properties of the Markov inclusions.

(3) In order to prove the formula in the statement, observe first that we have:

$$\begin{aligned} B' \cap A_1 &= (B' \cap B_1) \cap A_1 \\ &= (M_r(\mathbb{C}) \otimes Z(B)) \cap (M_r(\mathbb{C}) \otimes A) \\ &= M_r(\mathbb{C}) \otimes (B' \cap A) \end{aligned}$$

But this proves the assertion at $s = 0, k = 1$, and the general case follows from it.

(4) This is again clear, once again coming from Proposition 14.21 above, and from the general properties of the Markov inclusions. \square

In order to further refine all this, let us formulate the following key definition:

DEFINITION 14.23. *We say that a Markov inclusion $A \subset B$ is abelian if*

$$[A, B] = 0$$

with the commutant being computed inside B .

In other words, we are asking for the commutation relation $ab = ba$, for any $a \in A$ and $b \in B$. Note that this is the same as asking that B is an A -algebra, $A \subset Z(B)$. As basic examples, observe that all inclusions with $A = \mathbb{C}$ or with $B = \mathbb{C}^n$ are abelian.

The point with the above notion is that it leads to the following simple statement:

PROPOSITION 14.24. *With $\tilde{B}_k = M_r(\mathbb{C})^{\otimes k} \otimes Z(B)$, the relative commutants for the Jones tower $A \subset B \subset A_1 \subset B_1 \subset \dots$ of an abelian inclusion are given by:*

- (1) $A'_s \cap A_{s+k} = A_k$.
- (2) $A'_s \cap B_{s+k} = B_k$.
- (3) $B'_s \cap A_{s+k} = A_k$.
- (4) $B'_s \cap B_{s+k} = \tilde{B}_k$.

PROOF. This follows by comparing Proposition 14.21 and Proposition 14.22, and by using the fact that for an abelian inclusion we have:

$$Z(A) = A \quad , \quad A' \cap B = B \quad , \quad B' \cap A = A$$

Thus, we are led to the conclusion in the statement. □

Getting back now to the fixed point subfactors, from Theorem 14.15 above, we can improve the planar algebra computation there, in the abelian case, as follows:

THEOREM 14.25. *The relative commutants for the Jones tower*

$$N \subset M \subset N_1 \subset M_1 \subset \dots$$

associated to an abelian fixed point subfactor

$$(P \otimes A)^G \subset (P \otimes B)^G$$

are given by the following formulae:

- (1) $N'_s \cap N_{s+k} = A_k^G.$
- (2) $N'_s \cap M_{s+k} = B_k^G.$
- (3) $M'_s \cap N_{s+k} = A_k^G.$
- (4) $M'_s \cap M_{s+k} = \tilde{B}_k^G.$

PROOF. This follows indeed by combining the planar algebra computation from Theorem 14.15 with the result about the abelian inclusions from Proposition 14.24. □

In order to further advance now, the idea will be that of associating to the original inclusion $B_0 \subset B_1$ a certain combinatorial planar algebra $P(B_0 \subset B_1)$, as for the planar algebra associated to the fixed point subfactor itself to appear as follows:

$$P = P(B_0 \subset B_1)^G$$

More in detail, the idea will be that $P(B_0 \subset B_1)$ appears as a joint generalization of the spin and tensor planar algebras, discussed in the above, which appear as follows:

$$\mathcal{S}_N = P(\mathbb{C} \subset \mathbb{C}^N)$$

$$\mathcal{T}_N = P(\mathbb{C} \subset M_N(\mathbb{C}))$$

In practice now, we will need for all this the notion of planar algebra of a bipartite graph, generalizing the algebras $\mathcal{S}_N, \mathcal{T}_N$, constructed by Jones in [48].

So, let Γ be a bipartite graph, with vertex set $\Gamma_a \cup \Gamma_b$. It is useful to think of Γ as being the Bratteli diagram of an inclusion $A \subset B$, in the sense of Definition 14.16.

Our first task is to define the graded vector space P . Since the elements of P will be subject to a planar calculus, it is convenient to introduce them “in boxes”, as follows:

DEFINITION 14.26. Associated to Γ is the abstract vector space P_k spanned by the $2k$ -loops based at points of Γ_a . The basis elements of P_k will be denoted

$$x = \begin{pmatrix} e_1 & e_2 & \dots & e_k \\ e_{2k} & e_{2k-1} & \dots & e_{k+1} \end{pmatrix}$$

where e_1, e_2, \dots, e_{2k} is the sequence of edges of the corresponding $2k$ -loop.

Consider now the adjacency matrix of Γ , which is of the following type:

$$M = \begin{pmatrix} 0 & m \\ m^t & 0 \end{pmatrix}$$

We pick an M -eigenvalue $\gamma \neq 0$, and then a γ -eigenvector, as follows:

$$\eta : \Gamma_a \cup \Gamma_b \rightarrow \mathbb{C} - \{0\}$$

With this data in hand, we have the following construction, due to Jones [48]:

DEFINITION 14.27. Associated to any tangle is the multilinear map

$$T(x_1 \otimes \dots \otimes x_r) = \gamma^c \sum_x \delta(x_1, \dots, x_r, x) \prod_m \mu(e_m)^{\pm 1} x$$

where the objects on the right are as follows:

- (1) The sum is over the basis of P_k , and c is the number of circles of T .
- (2) $\delta = 1$ if all strings of T join pairs of identical edges, and $\delta = 0$ if not.
- (3) The product is over all local maxima and minima of the strings of T .
- (4) e_m is the edge of Γ labelling the string passing through m (when $\delta = 1$).
- (5) $\mu(e) = \sqrt{\eta(e_f)/\eta(e_i)}$, where e_i, e_f are the initial and final vertex of e .
- (6) The \pm sign is $+$ for a local maximum, and $-$ for a local minimum.

In other words, we plug the loops x_1, \dots, x_r into the input boxes of T , and then we construct the “output”: this is the sum of all loops x satisfying the compatibility condition $\delta = 1$, altered by certain normalization factors, coming from the eigenvector η .

Let us work out now the precise formula of the action, for 6 carefully chosen tangles, which are of key importance for the considerations to follow. This study will be useful as well as an introduction to Jones’ result in [48], stating that P is a planar algebra:

DEFINITION 14.28. We have the following examples of tangles:

- (1) Identity 1_k : the (k, k) -tangle having $2k$ vertical strings.
- (2) Multiplication M_k : the (k, k, k) -tangle having $3k$ vertical strings.
- (3) Inclusion I_k : the $(k, k + 1)$ -tangle like 1_k , with an extra string at right.
- (4) Shift J_k : the $(k, k + 2)$ -tangle like 1_k , with two extra strings at left.
- (5) Expectation U_k : the $(k + 1, k)$ -tangle like 1_k , with a curved string at right.
- (6) Jones projection E_k : the $(k + 2)$ -tangle having two semicircles at right.

Let us look first at the identity 1_k . Since the solutions of $\delta(x, y) = 1$ are the pairs of the form (x, x) , this tangle acts by the identity:

$$1_k \begin{pmatrix} f_1 & \cdots & f_k \\ e_1 & \cdots & e_k \end{pmatrix} = \begin{pmatrix} f_1 & \cdots & f_k \\ e_1 & \cdots & e_k \end{pmatrix}$$

A similar argument applies to the multiplication M_k , which acts as follows:

$$M_k \left(\begin{pmatrix} f_1 & \cdots & f_k \\ e_1 & \cdots & e_k \end{pmatrix} \otimes \begin{pmatrix} h_1 & \cdots & h_k \\ g_1 & \cdots & g_k \end{pmatrix} \right) = \delta_{f_1 g_1} \cdots \delta_{f_k g_k} \begin{pmatrix} h_1 & \cdots & h_k \\ e_1 & \cdots & e_k \end{pmatrix}$$

Regarding now the inclusion I_k , the solutions of $\delta(x_0, x) = 1$ being the elements x obtained from x_0 by adding to the right a vector of the form $\begin{pmatrix} g \\ g \end{pmatrix}$, we have:

$$I_k \begin{pmatrix} f_1 & \cdots & f_k \\ e_1 & \cdots & e_k \end{pmatrix} = \sum_g \begin{pmatrix} f_1 & \cdots & f_k & g \\ e_1 & \cdots & e_k & g \end{pmatrix}$$

The same method applies to the shift J_k , whose action is given by:

$$J_k \begin{pmatrix} f_1 & \cdots & f_k \\ e_1 & \cdots & e_k \end{pmatrix} = \sum_{gh} \begin{pmatrix} g & h & f_1 & \cdots & f_k \\ g & h & e_1 & \cdots & e_k \end{pmatrix}$$

Let us record some partial conclusions, coming from the above formulae:

PROPOSITION 14.29. *The graded vector space $P = (P_k)$ constructed above becomes a graded algebra, with the multiplication map given by*

$$xy = M_k(x \otimes y)$$

on each vector space P_k , and with the above inclusion maps I_k . Also, the shift tangle J_k acts as an injective morphism of algebras $P_k \rightarrow P_{k+2}$.

PROOF. The fact that the multiplication is associative follows from its above formula, which is nothing but a generalization of the usual matrix multiplication. The assertions about the inclusions and shifts follow as well by using their above explicit formula. \square

Let us go back now to the remaining tangles in Definition 14.28. The usual method applies to the expectation tangle U_k , which acts with a spin factor, as follows:

$$U_k \begin{pmatrix} f_1 & \cdots & f_k & h \\ e_1 & \cdots & e_k & g \end{pmatrix} = \delta_{gh} \mu(g)^2 \begin{pmatrix} f_1 & \cdots & f_k \\ e_1 & \cdots & e_k \end{pmatrix}$$

As for the Jones projection E_k , this tangle has no input box, so we can only apply it to the unit of \mathbb{C} . And when doing so, we obtain the following element:

$$E_k(1) = \sum_{egh} \mu(g)\mu(h) \begin{pmatrix} e_1 & \cdots & e_k & h & h \\ e_1 & \cdots & e_k & g & g \end{pmatrix}$$

Once again, let us record now some partial conclusions, coming from these formulae:

PROPOSITION 14.30. *The elements $e_k = \gamma^{-1}E_k(1)$ are projections, and define a representation of the Temperley-Lieb algebra, as follows:*

$$TL(\gamma) \rightarrow P$$

The maps U_k are bimodule morphisms with respect to I_k , and their composition is the canonical trace on the image of $TL(\gamma)$.

PROOF. The proof of all the assertions is standard, by using the fact that η is a γ -eigenvector of the adjacency matrix. Note that the statement itself is just a generalization of the usual Temperley-Lieb algebra representation on tensors, from [47]. \square

In fact, more generally, we have the following result, from Jones' paper [48]:

THEOREM 14.31. *The graded linear space $P = (P_k)$, together with the action of the planar tangles given above, is a planar algebra.*

PROOF. This is something which is quite routine, starting from the above detailed study of the main planar algebra tangles, which can be proved for instance by using Theorem 14.11 above, with the rotations R_k there being replaced by the shifts J_k . Also, let us mention that all this generalizes the previous constructions of the spin and tensor planar algebras $\mathcal{S}_N, \mathcal{T}_N$, which appear respectively from the Bratteli diagrams of the inclusions $\mathbb{C} \subset \mathbb{C}^N$ and $\mathbb{C} \subset M_N(\mathbb{C})$. For full details on all this, we refer to Jones' paper [48]. \square

Let us go back now to the Markov inclusions $A \subset B$, as before. We have here the following result, regarding such inclusions, also from Jones' paper [48]:

THEOREM 14.32. *The planar algebra associated to the graph of $A \subset B$, with eigenvalue $\gamma = \sqrt{r}$ and eigenvector $\eta(i) = a_i/\sqrt{\dim A}$, $\eta(j) = b_j/\sqrt{\dim B}$, is as follows:*

- (1) *The graded algebra structure is given by $P_{2k} = A' \cap A_k$, $P_{2k+1} = A' \cap B_k$.*
- (2) *The elements e_k are the Jones projections for $A \subset B \subset A_1 \subset B_1 \subset \dots$*
- (3) *The expectation and shift are given by the above formulae.*

PROOF. As a first observation, η is indeed a γ -eigenvector for the adjacency matrix of the graph. Indeed, we have the following formulae:

$$m^t a = b \quad , \quad mb = ra \quad , \quad \sqrt{r} = \|b\|/\|a\|$$

By using these formulae, we have the following computation:

$$\begin{aligned} \begin{pmatrix} 0 & m \\ m^t & 0 \end{pmatrix} \begin{pmatrix} a/\|a\| \\ b/\|b\| \end{pmatrix} &= \begin{pmatrix} \gamma^2 a/\|b\| \\ b/\|a\| \end{pmatrix} \\ &= \gamma \begin{pmatrix} \gamma a/\|b\| \\ b/\gamma\|a\| \end{pmatrix} \\ &= \gamma \begin{pmatrix} a/\|a\| \\ b/\|b\| \end{pmatrix} \end{aligned}$$

Since the algebra A was supposed abelian, the Jones tower algebras A_k, B_k are simply the span of the $4k$ -paths, respectively $4k + 2$ -paths on Γ , starting at points of Γ_a . With this description in hand, when taking commutants with A we have to just have to restrict attention from paths to loops, and we obtain the above spaces P_{2k}, P_{2k+1} . See [48]. \square

In the particular case of the inclusions satisfying $[A, B] = 0$, we have:

PROPOSITION 14.33. *The “bipartite graph” planar algebra $P(A \subset B)$ associated to an abelian inclusion $A \subset B$ can be described as follows:*

- (1) *As a graded algebra, this is the Jones tower $A \subset B \subset A_1 \subset B_1 \subset \dots$*
- (2) *The Jones projections and expectations are the usual ones for this tower.*
- (3) *The shifts correspond to the canonical identifications $A'_1 \cap P_{k+2} = P_k$.*

PROOF. The first assertion is a reformulation of Theorem 14.32 in the abelian case, by using the identifications $A' \cap A_k = A_k$ and $A' \cap B_k = B_k$ from Proposition 14.24.

The assertion on Jones projections follows as well from Theorem 14.32, and the assertion on expectations follows from the fact that their composition is the usual trace.

Regarding now the third assertion, let us recall first from Proposition 14.24 that we have indeed identifications $A'_1 \cap A_{k+1} = A_k$ and $A'_1 \cap B_{k+1} = B_k$. By using the path model for these algebras, as in the proof of Theorem 14.32, we obtain the result. \square

In order to formulate now our main result, regarding the subfactors associated to the compact quantum groups G , we will need a few abstract notions. Let us start with:

DEFINITION 14.34. *Let P_1, P_2 be two finite dimensional algebras, coming with coactions $\alpha_i : P_i \rightarrow P_i \otimes L^\infty(G)$, and let $T : P_1 \rightarrow P_2$ be a linear map.*

- (1) *We say that T is G -equivariant if $(T \otimes id)\alpha_1 = \alpha_2 T$.*
- (2) *We say that T is weakly G -equivariant if $T(P_1^G) \subset P_2^G$.*

Consider now a planar algebra $P = (P_k)$. The annular category over P is the collection of maps $T : P_k \rightarrow P_l$ coming from the “annular” tangles, having at most one input box. These maps form sets $Hom(k, l)$, and these sets form a category [45]. We have:

DEFINITION 14.35. *A coaction of $L^\infty(G)$ on P is a graded algebra coaction*

$$\alpha : P \rightarrow P \otimes L^\infty(G)$$

such that the annular tangles are weakly G -equivariant.

This definition is something a bit technical, but this is what comes out of the known examples that we have, coming from the fixed point subfactors.

In fact, as we will show below, the examples are basically those coming from actions of quantum groups on Markov inclusions $A \subset B$, under the assumption $[A, B] = 0$.

For the moment, at the generality level of Definition 14.35, we have:

PROPOSITION 14.36. *If G acts on a planar algebra P , then P^G is a planar algebra.*

PROOF. The weak equivariance condition tells us that the annular category is contained in the suboperad $\mathcal{P}' \subset \mathcal{P}$ consisting of tangles which leave invariant P^G . On the other hand the multiplicativity of α gives $M_k \in \mathcal{P}'$, for any k . Now since \mathcal{P} is generated by multiplications and annular tangles, we get $\mathcal{P}' = \mathcal{P}$, and we are done. \square

Let us go back now to the abelian inclusions. We have the following key result:

PROPOSITION 14.37. *If G acts on an abelian inclusion $A \subset B$, the canonical extension of this coaction to the Jones tower is a coaction of G on the planar algebra $P(A \subset B)$.*

PROOF. We know from the above that, as a graded algebra, $P = P(A \subset B)$ coincides with the Jones tower for our inclusion, denoted as follows:

$$A \subset B \subset A_1 \subset B_1 \subset \dots$$

Thus the coaction in the statement can be regarded as a graded coaction, as follows:

$$\alpha : P \rightarrow P \otimes L^\infty(G)$$

In order to finish, we have to prove that the annular tangles are weakly equivariant, as in Definition 14.35 above, and this can be done as follows:

(1) First, since the annular category is generated by I_k, E_k, U_k, J_k , we just have to prove that these 4 particular tangles are weakly equivariant. Now since I_k, E_k, U_k are plainly equivariant, by construction of the coaction of G on the Jones tower, it remains to prove that the shift J_k is weakly equivariant.

(2) We know that the image of the fixed point subfactor shift J'_k is formed by the G -invariant elements of the relative commutant $A'_1 \cap P_{k+2} = P_k$. Now since this commutant is the image of the planar shift J_k , we have $Im(J_k) = Im(J'_k)$, and this gives the result. \square

With the above result in hand, we can now prove:

PROPOSITION 14.38. *Assume that G acts on an abelian inclusion $A \subset B$. Then the graded vector space of fixed points $P(A \subset B)^G$ is a planar subalgebra of $P(A \subset B)$.*

PROOF. This follows indeed from Proposition 14.36 and Proposition 14.37. \square

We are now in position of stating and proving our main result:

THEOREM 14.39. *In the abelian case, the planar algebra of the fixed point subfactor*

$$(P \otimes A)^G \subset (P \otimes B)^G$$

is the fixed point algebra $P(A \subset B)^G$ of the bipartite graph algebra $P(A \subset B)$.

PROOF. This basically follows from what we have, as follows:

(1) Let $P = P(A \subset B)$, and let Q be the planar algebra of the fixed point subfactor. We know that we have an equality of graded algebras $Q = P^G$. Thus, it remains to prove that the planar algebra structure on Q coming from the fixed point subfactor agrees with the planar algebra structure of P , coming from Proposition 14.33.

(2) Since \mathcal{P} is generated by the annular category \mathcal{A} and by the multiplication tangles M_k , we just have to check that the annular tangles agree on P, Q . Moreover, since \mathcal{A} is generated by I_k, E_k, U_k, J_k , we just have to check that these tangles agree on P, Q .

(3) We know that $Q \subset P$ is an inclusion of graded algebras, that all the Jones projections for P are contained in Q , and that the conditional expectations agree. Thus the tangles I_k, E_k, U_k agree on P, Q , and the only verification left is that for the shift J_k .

(4) Now by using either the axioms of Popa in [72], or the construction of Jones in [48], it is enough to show that the image of the subfactor shift J'_k coincides with that of the planar shift J_k . But this follows as in the proof of Proposition 14.37. \square

14d. Tannakian results

We discuss here some converses to the above results, which are rather specialized results, of Tannakian nature. We will first prove that any quantum permutation group $G \subset S_N^+$ produces a planar subalgebra of \mathcal{S}_N . In order to do so, we first have:

THEOREM 14.40. *Given a quantum permutation group $G \subset S_N^+$, consider the associated coaction map on $C(X)$, where $X = \{1, \dots, N\}$,*

$$\Phi : C(X) \rightarrow C(X) \otimes C(G) \quad , \quad e_j \rightarrow \sum_j e_j \otimes u_{ji}$$

and then consider the tensor powers of this coaction, which are the following linear maps:

$$\Phi^k : C(X^k) \rightarrow C(X^k) \otimes C(G) \quad , \quad e_{i_1 \dots i_k} \rightarrow \sum_{j_1 \dots j_k} e_{j_1 \dots j_k} \otimes u_{j_1 i_1} \dots u_{j_k i_k}$$

The fixed point spaces of these latter coactions are then given by the formula

$$P_k = \text{Fix}(u^{\otimes k})$$

and form a planar subalgebra of the spin planar algebra \mathcal{S}_N .

PROOF. This can be done in several steps, as follows:

(1) Since the map Φ is a coaction, coming from the corepresentation u , its tensor powers Φ^k are coactions too, coming from the corepresentations $u^{\otimes k}$, and at the level of the fixed point algebras we have the following formula, which is standard:

$$P_k = \text{Fix}(u^{\otimes k})$$

(2) In order to prove now the planar algebra assertion, we use the presentation result for the spin planar algebras established before, involving the multiplications, inclusions, expectations, Jones projections and rotations.

(3) Consider the rotation R_k . Rotating, then applying Φ^k , and rotating backwards by R_k^{-1} is the same as applying Φ^k , then rotating each k -fold product of coefficients of Φ .

(4) Thus the elements obtained by rotating, then applying Φ^k , or by applying Φ^k , then rotating, differ by a sum of Dirac masses tensored with commutators in $A = C(G)$:

$$\Phi^k R_k(x) - (R_k \otimes id)\Phi^k(x) \in C(X^k) \otimes [A, A]$$

(5) Now let \int_A be the Haar functional of A , and consider the conditional expectation onto the fixed point algebra P_k , which is given by the following formula:

$$\phi_k = \left(id \otimes \int_A \right) \Phi^k$$

The square of the antipode being the identity, the Haar integration \int_A is a trace, so it vanishes on commutators. Thus R_k commutes with ϕ_k :

$$\phi_k R_k = R_k \phi_k$$

(6) The commutation relation $\phi_k T = T \phi_l$ holds in fact for any (l, k) -tangle T . These tangles are called annular, and the proof is by verification on generators of the annular category. In particular we obtain, for any annular tangle T :

$$\phi_k T \phi_l = T \phi_l$$

(7) We conclude from this that the annular category is contained in the suboperad $\mathcal{P}' \subset \mathcal{P}$ of the planar operad consisting of tangles T satisfying the following condition, where $\phi = (\phi_k)$, and where $i(\cdot)$ is the number of input boxes:

$$\phi T \phi^{\otimes i(T)} = T \phi^{\otimes i(T)}$$

On the other hand the multiplicativity of Φ^k gives $M_k \in \mathcal{P}'$. Since \mathcal{P} is generated by multiplications and annular tangles, it follows that we have $\mathcal{P}' = \mathcal{P}$.

(8) Thus for any tangle T the corresponding multilinear map between spaces $P_k(X)$ restricts to a multilinear map between spaces P_k . In other words, the action of the planar operad \mathcal{P} restricts to P , and makes it a subalgebra of \mathcal{S}_N , as claimed. □

As a second result now, completing our study, we have:

THEOREM 14.41. *Given a subalgebra $Q \subset \mathcal{S}_N$, there is a unique quantum group*

$$G \subset S_N^+$$

whose associated planar algebra is Q .

PROOF. The idea is that this will follow by applying Tannakian duality to the annular category over Q . Let n, m be positive integers. To any element $T_{n+m} \in Q_{n+m}$ we can associate a linear map $L_{nm}(T_{n+m}) : P_n(X) \rightarrow P_m(X)$ in the following way:

$$L_{nm} \left(\begin{array}{c} | | | \\ T_{n+m} \\ | | | \end{array} \right) : \left(\begin{array}{c} | \\ a_n \\ | \end{array} \right) \rightarrow \left(\begin{array}{c} | | \cap \\ T_{n+m} \\ | | | \\ a_n \\ \cup | | \end{array} \right)$$

That is, we consider the planar $(n, n + m, m)$ -tangle having an small input n -box, a big input $n + m$ -box and an output m -box, with strings as on the picture of the right. This defines a certain multilinear map, as follows:

$$P_n(X) \otimes P_{n+m}(X) \rightarrow P_m(X)$$

Now let us put the element T_{n+m} in the big input box. We obtain in this way a certain linear map $P_n(X) \rightarrow P_m(X)$, that we call L_{nm} . Now let us set:

$$Q_{nm} = \left\{ L_{nm}(T_{n+m}) : P_n(X) \rightarrow P_m(X) \mid T_{n+m} \in Q_{n+m} \right\}$$

These spaces form a Tannakian category, and so by [99] we obtain a Woronowicz algebra (A, u) , such that the following equalities hold, for any m, n :

$$Hom(u^{\otimes m}, u^{\otimes n}) = Q_{mn}$$

We prove that u is a magic unitary. We have $Hom(1, u^{\otimes 2}) = Q_{02} = Q_2$, so the unit of Q_2 must be a fixed vector of $u^{\otimes 2}$. But $u^{\otimes 2}$ acts on the unit of Q_2 as follows:

$$\begin{aligned} u^{\otimes 2}(1) &= u^{\otimes 2} \left(\sum_i \begin{pmatrix} i & i \\ i & i \end{pmatrix} \right) \\ &= \sum_{ikl} \begin{pmatrix} k & k \\ l & l \end{pmatrix} \otimes u_{ki}u_{li} \\ &= \sum_{kl} \begin{pmatrix} k & k \\ l & l \end{pmatrix} \otimes (uu^t)_{kl} \end{aligned}$$

From $u^{\otimes 2}(1) = 1 \otimes 1$ we get that uu^t is the identity matrix. Together with the unitarity of u , this gives the following formulae:

$$u^t = u^* = u^{-1}$$

Consider the Jones projection $E_1 \in Q_3$. After isotoping, $L_{21}(E_1)$ looks as follows:

$$L_{21} \left(\begin{array}{c} | \cup \\ | \cap \end{array} \right) : \left(\begin{array}{c} | | | \\ i \quad i \\ j \quad j \\ | | | \end{array} \right) \rightarrow \left(\begin{array}{c} | \\ i \quad i \\ j \quad j \\ | \supset \end{array} \right) = \delta_{ij} \left(\begin{array}{c} | \\ i \\ | \end{array} \right)$$

In other words, the linear map $M = L_{21}(E_1)$ is the multiplication $\delta_i \otimes \delta_j \rightarrow \delta_{ij}\delta_i$:

$$M \begin{pmatrix} i & i \\ j & j \end{pmatrix} = \delta_{ij} \begin{pmatrix} i \\ i \end{pmatrix}$$

Consider now the following element of $C(X) \otimes A$:

$$\begin{aligned} (M \otimes id)u^{\otimes 2} \left(\begin{pmatrix} i & i \\ j & j \end{pmatrix} \otimes 1 \right) &= (M \otimes id) \left(\sum_{kl} \begin{pmatrix} k & k \\ l & l \end{pmatrix} \otimes u_{ki}u_{lj} \right) \\ &= \sum_k \begin{pmatrix} k \\ k \end{pmatrix} \delta_k \otimes u_{ki}u_{kj} \end{aligned}$$

Since $M \in Q_{21} = Hom(u^{\otimes 2}, u)$, this equals the following element of $C(X) \otimes A$:

$$\begin{aligned} u(M \otimes id) \left(\begin{pmatrix} i & i \\ j & j \end{pmatrix} \otimes 1 \right) &= u \left(\delta_{ij} \begin{pmatrix} i \\ i \end{pmatrix} \delta_i \otimes 1 \right) \\ &= \sum_k \begin{pmatrix} k \\ k \end{pmatrix} \delta_k \otimes \delta_{ij}u_{ki} \end{aligned}$$

Thus $u_{ki}u_{kj} = \delta_{ij}u_{ki}$ for any i, j, k . With $i = j$ we get $u_{ki}^2 = u_{ki}$, and together with the formula $u^t = u^*$ this shows that all entries of u are self-adjoint projections. With $i \neq j$ we get $u_{ki}u_{kj} = 0$, so the projections on each row of u are orthogonal to each other. Together with $u^t = u^{-1}$, this shows that each row of u is a partition of unity with self-adjoint projections. The antipode is given by the formula $(id \otimes S)u = u^*$. But u^* is the transpose of u , so we can apply S to the formulae saying that rows of u are partitions of unity, and we get that columns of u are also partitions of unity. Thus u is a magic unitary.

Now if P is the planar algebra associated to u , we have $Hom(1, v^{\otimes n}) = P_n = Q_n$, as desired. As for the uniqueness, this is clear from the Peter-Weyl theory. \square

The results established above, regarding the subgroups $G \subset S_N^+$, have several generalizations, to the subgroups $G \subset O_N^+$ and $G \subset U_N^+$, as well as subfactor versions, going beyond the purely combinatorial level. We refer here to the literature on the subject.

14e. Exercises

There were several things formulated with proof in this chapter, and as a main exercise regarding all this, more advanced material, we have:

EXERCISE 14.42. *Look up the theorem stating that any planar algebra produces a subfactor, and write down a brief account of what you learned.*

As already mentioned in the above, there are several theorems here, which are all non-trivial. And there is a big open question too, concerning hyperfiniteness.

CHAPTER 15

Commuting squares

15a. Commuting squares

In this chapter and in the next one we discuss a number of more specialized aspects of subfactor theory, making the link with several advanced topics, such as quantum permutation groups, noncommutative geometry, free probability, and more.

A first question, to be discussed in the present chapter, is the explicit construction of subfactors by using some suitable combinatorial data, encoded in a structure called “commuting square”. Let us start with the following definition:

DEFINITION 15.1. *A commuting square in the sense of subfactor theory is a commuting diagram of finite dimensional algebras with traces, as follows,*

$$\begin{array}{ccc} C_{01} & \longrightarrow & C_{11} \\ \uparrow & & \uparrow \\ C_{00} & \longrightarrow & C_{10} \end{array}$$

having the property that the conditional expectations $C_{11} \rightarrow C_{01}$ and $C_{11} \rightarrow C_{10}$ commute, and their product is the conditional expectations $C_{11} \rightarrow C_{00}$.

There are many examples of such commuting squares, the idea being always that such squares must come from subtle combinatorial data. As an illustration for this principle, in relation with the complex Hadamard matrices that we met in chapter 11, we have:

PROPOSITION 15.2. *Given an Hadamard matrix $H \in M_N(\mathbb{C})$, the diagram formed by the associated pair of orthogonal maximal commutative subalgebras of $M_N(\mathbb{C})$,*

$$\begin{array}{ccc} \Delta & \longrightarrow & M_N(\mathbb{C}) \\ \uparrow & & \uparrow \\ \mathbb{C} & \longrightarrow & H\Delta H^* \end{array}$$

where $\Delta \subset M_N(\mathbb{C})$ are the diagonal matrices, is a commuting square.

PROOF. The expectation $E_\Delta : M_N(\mathbb{C}) \rightarrow \Delta$ is the operation $M \rightarrow M_\Delta$ which consists in keeping the diagonal, and erasing the rest. Consider now the other expectation:

$$E_{H\Delta H^*} : M_N(\mathbb{C}) \rightarrow H\Delta H^*$$

It is better to identify this with the following expectation, with $U = H/\sqrt{N}$:

$$E_{U\Delta U^*} : M_N(\mathbb{C}) \rightarrow U\Delta U^*$$

This must be given by a formula of type $M \rightarrow UX_\Delta U^*$, with X satisfying:

$$\langle M, UDU^* \rangle = \langle UX_\Delta U^*, UDU^* \rangle, \quad \forall D \in \Delta$$

The scalar products being given by $\langle a, b \rangle = \text{tr}(ab^*)$, this condition reads:

$$\text{tr}(MUD^*U^*) = \text{tr}(X_\Delta D^*) \quad , \quad \forall D \in \Delta$$

Thus $X = U^*MU$, and the formulae of our two expectations are as follows:

$$\begin{aligned} E_\Delta(M) &= M_\Delta \\ E_{U\Delta U^*}(M) &= U(U^*MU)_\Delta U^* \end{aligned}$$

With these formulae in hand, we have the following computation:

$$\begin{aligned} (E_\Delta E_{U\Delta U^*} M)_{ij} &= \delta_{ij} (U(U^*MU)_\Delta U^*)_{ii} \\ &= \delta_{ij} \sum_k U_{ik} (U^*MU)_{kk} \bar{U}_{ik} \\ &= \delta_{ij} \sum_k \frac{1}{N} \cdot (U^*MU)_{kk} \\ &= \delta_{ij} \text{tr}(U^*MU) \\ &= \delta_{ij} \text{tr}(M) \\ &= (E_{\mathbb{C}} M)_{ij} \end{aligned}$$

As for the other composition, the computation here is similar, as follows:

$$\begin{aligned} (E_{U\Delta U^*} E_\Delta M)_{ij} &= (U(U^*M_\Delta U)_\Delta U^*)_{ij} \\ &= \sum_k U_{ik} (U^*M_\Delta U)_{kk} \bar{U}_{jk} \\ &= \sum_{kl} U_{ik} \bar{U}_{lk} M_{ll} U_{lk} \bar{U}_{jk} \\ &= \frac{1}{N} \sum_{kl} U_{ik} M_{ll} \bar{U}_{jk} \\ &= \delta_{ij} \text{tr}(M) \\ &= (E_{\mathbb{C}} M)_{ij} \end{aligned}$$

Thus, we have indeed a commuting square, as claimed. \square

There are many other explicit examples of commuting squares, all coming from subtle combinatorial data, and more on this later. Getting back now to the general case, let us consider an arbitrary commuting square, as in Definition 15.1 above:

$$\begin{array}{ccc}
 C_{01} & \longrightarrow & C_{11} \\
 \uparrow & & \uparrow \\
 C_{00} & \longrightarrow & C_{10}
 \end{array}$$

The point is that, under some suitable extra mild assumptions, any such square C produces a subfactor of the hyperfinite II_1 factor R . Indeed, by performing the basic construction, in finite dimensions, we obtain a whole array, as follows:

$$\begin{array}{ccccccc}
 & A_0 & & A_1 & & A_2 & \\
 & \uparrow & & \uparrow & & \uparrow & \\
 & \vdots & & \vdots & & \vdots & \\
 C_{02} & \longrightarrow & C_{12} & \longrightarrow & C_{22} & \cdots \longrightarrow & B_2 \\
 \uparrow & & \uparrow & & \uparrow & & \\
 C_{01} & \longrightarrow & C_{11} & \longrightarrow & C_{21} & \cdots \longrightarrow & B_1 \\
 \uparrow & & \uparrow & & \uparrow & & \\
 C_{00} & \longrightarrow & C_{10} & \longrightarrow & C_{20} & \cdots \longrightarrow & B_0
 \end{array}$$

To be more precise, by performing the basic construction in both possible directions, namely to the right and upwards, we obtain a whole array of finite dimensional algebras with traces, that we can denote $(C_{ij})_{i,j \geq 0}$, as above. Once this done, we can further consider the von Neumann algebras obtained in the limit, on each vertical and horizontal line, and denote them A_i, B_j , as above, and we have the following result:

THEOREM 15.3. *In the context of the above diagram, the limiting von Neumann algebras A_i, B_j are all isomorphic to the hyperfinite II_1 factor R , and:*

- (1) $A_0 \subset A_1$ is a subfactor, and $\{A_i\}$ is the Jones tower for it.
- (2) The corresponding planar algebra is given by $A'_0 \cap A_k = C'_{01} \cap C_{k0}$.
- (3) A similar result holds for the “horizontal” subfactor $B_0 \subset B_1$.

PROOF. This is something very standard, with the factoriality of the limiting von Neumann algebras A_i, B_j coming as a consequence of the general commutant computation in (2), which is independent from it, with the hyperfiniteness of the same A_i, B_j algebras being clear by definition, and with the idea for the rest being as follows:

(1) This is somewhat clear from definitions, or rather from a quick verification of the basic construction axioms, as formulated in chapter 13 above, because the tower of algebras $\{A_i\}$ appears by definition as the $j \rightarrow \infty$ limit of the towers of algebras $\{C_{ij}\}$, which are all Jones towers. Thus the limiting tower $\{A_i\}$ is also a Jones tower.

(2) This is the non-trivial result, called Ocneanu compactness theorem, and whose proof is by doing some linear algebra. To be more precise, in one sense the result is clear, because by definition of the algebras $\{A_i\}$, we have inclusions as follows:

$$A'_0 \cap A_k \supset C'_{01} \cap C_{k0}$$

In the other sense things are more tricky, mixing standard linear algebra with some functional analysis too, and we refer here to Ocneanu’s lecture notes [64], [65].

(3) This follows from (1,2), by transposing the whole diagram. Indeed, given a commuting square as in Definition 15.1, its transpose is a commuting square as well:

$$\begin{array}{ccc} C_{10} & \longrightarrow & C_{11} \\ \uparrow & & \uparrow \\ C_{00} & \longrightarrow & C_{01} \end{array}$$

Thus we can apply (1,2) above to this commuting square, and we obtain in this way Jones tower and planar algebra results for the “horizontal” subfactor $B_0 \subset B_1$. \square

In relation with the examples of commuting squares that we have so far, namely those coming from the Hadamard matrices, from Proposition 15.2 above, we have:

THEOREM 15.4. *Given a complex Hadamard matrix $H \in M_N(\mathbb{C})$, the diagram formed by the associated pair of orthogonal maximal commutative subalgebras, namely*

$$\begin{array}{ccc} \Delta & \longrightarrow & M_N(\mathbb{C}) \\ \uparrow & & \uparrow \\ \mathbb{C} & \longrightarrow & H\Delta H^* \end{array}$$

is a commuting square in the sense of subfactor theory, and the associated planar algebra $P = (P_k)$ is given by the following formula, in terms of H itself,

$$T \in P_k \iff T^\circ G^2 = G^{k+2} T^\circ$$

where the objects on the right are constructed as follows:

- (1) $T^\circ = id \otimes T \otimes id$.
- (2) $G_{ia}^{jb} = \sum_k H_{ik} \bar{H}_{jk} \bar{H}_{ak} H_{bk}$.
- (3) $G_{i_1 \dots i_k, j_1 \dots j_k}^k = G_{i_k i_{k-1}}^{j_k j_{k-1}} \cdots G_{i_2 i_1}^{j_2 j_1}$.

PROOF. We have several assertions here, the idea being as follows:

(1) The fact that we have indeed a commuting square is something quite elementary, that we already know, from Proposition 15.2 above.

(2) The computation of the associated planar algebra, directly in terms of H , is something which is definitely possible, thanks to the formula in Theorem 15.3 (2).

(3) As for the precise formula of the planar algebra, which emerges by doing the computation, we will be back to it, with full details, later on. \square

Summarizing, we have so far an interesting combinatorial notion, that of a commuting square, and a method of producing subfactors and planar algebras out of it. We will further explore all the possibilities that this opens up, in what follows.

15b. Matrix models

Before getting into more complicated theory and examples, let us clarify the computation for the Hadamard matrix commuting squares, from Theorem 15.4 above. Our claim is that all this is related, and in a beautiful way, to the quantum permutation groups.

In order to discuss this, with full details, and to present as well some generalizations, afterwards, we will need some substantial preliminaries on the quantum permutation groups, and their matrix models. Let us start with a general definition, as follows:

DEFINITION 15.5. *A matrix model for a Woronowicz algebra,*

$$A = C(G)$$

is a morphism of C^ -algebras of the following type,*

$$\pi : C(G) \rightarrow M_K(C(T))$$

with T being a compact space, and $K \geq 1$ being an integer.

In other words, a matrix model for a Woronowicz algebra is simply a morphism from it into another algebra, and with the target algebra being assumed to be something very familiar, so that we can do computations inside it, namely a random matrix algebra.

We would like of course our model to be faithful, in some appropriate sense, in order to recover information about G , via computations in the model. We first have:

DEFINITION 15.6. *A matrix model $\pi : C(G) \rightarrow M_K(C(T))$ is called stationary when*

$$\int_G = \left(\text{tr} \otimes \int_T \right) \pi$$

where \int_T is the integration with respect to a given probability measure on T .

Here the term “stationary” comes from a functional analytic interpretation of all this, with a certain Cesàro limit being needed to be stationary, and this will be explained later on. Yet another explanation comes from a certain relation with the lattice models, but this relation is rather something folklore, not axiomatized yet. We will be back to this.

As a first result now, the stationarity property implies the faithfulness:

THEOREM 15.7. *Assuming that $G \subset U_N^+$ has a stationary model,*

$$\begin{aligned} \pi : C(G) &\rightarrow M_K(C(T)) \\ \int_G &= \left(\text{tr} \otimes \int_T \right) \pi \end{aligned}$$

it follows that G must be coamenable, and that the model is faithful.

PROOF. We have two assertions to be proved, the idea being as follows:

(1) Assume that we have a stationary model, as in the statement. By performing the GNS construction with respect to \int_G , we obtain a factorization as follows, which commutes with the respective canonical integration functionals:

$$\pi : C(G) \rightarrow C(G)_{\text{red}} \subset M_K(C(T))$$

Thus, in what regards the coamenability question, we can assume that π is faithful. With this assumption made, observe that we have embeddings as follows:

$$C^\infty(G) \subset C(G) \subset M_K(C(T))$$

The point now is that the GNS construction gives a better embedding, as follows:

$$L^\infty(G) \subset M_K(L^\infty(T))$$

Now since the von Neumann algebra on the right is of type I, so must be its subalgebra $A = L^\infty(G)$. This means that, when writing the center of this latter algebra as $Z(A) = L^\infty(X)$, the whole algebra decomposes over X , as an integral of type I factors:

$$L^\infty(G) = \int_X M_{K_x}(\mathbb{C}) \, dx$$

In particular, we can see from this that $C^\infty(G) \subset L^\infty(G)$ has a unique C^* -norm, and so G is coamenable. Thus we have proved our first assertion.

(2) The second assertion follows as well from the above, because our factorization of π consists of the identity, and of an inclusion. \square

As a conclusion to all this, we have a potentially interesting notion of a matrix model, but the most straightforward notion of faithfulness of such models, namely the plain faithfulness, or rather the slightly modified notion of stationarity, from Definition 15.6 above, leads to type I algebras, and so to nothing interesting. Indeed, recall from chapters 11-12 above that the liberation procedure conjecturally leads to II_1 factors.

There is however a way of improving all this. For this purpose, consider a usual discrete group Γ , and a matrix model for its usual group algebra:

$$\pi : C^*(\Gamma) \rightarrow M_K(C(T))$$

This model must come from a unitary group representation, as follows:

$$\rho : \Gamma \rightarrow C(T, U_K)$$

Now observe that when this unitary group representation ρ is faithful, the associated algebra representation π is in general not faithful. This is for instance because when the underlying compact space is $T = \{.\}$, the target algebra of π is finite dimensional, while the group algebra $C^*(\Gamma)$ is infinite dimensional, as long as Γ is an infinite group.

On the other hand, in this case, when ρ is faithful, the algebra representation π obviously “reminds” Γ , and so can be used in order to fully understand Γ .

Summarizing, we have a new idea here, basically saying that, for practical purposes, the faithfulness property can be replaced with something much weaker.

This weaker notion is called “inner faithfulness”, and the theory here is as follows:

DEFINITION 15.8. *Let $\pi : C(G) \rightarrow M_K(C(T))$ be a matrix model.*

- (1) *The Hopf image of π is the smallest quotient Hopf C^* -algebra $C(G) \rightarrow C(H)$ producing a factorization of type $\pi : C(G) \rightarrow C(H) \rightarrow M_K(C(T))$.*
- (2) *When the inclusion $H \subset G$ is an isomorphism, i.e. when there is no non-trivial factorization as above, we say that π is inner faithful.*

As a basic illustration for these notions, in the case where $G = \widehat{\Gamma}$ is a group dual, π must come from a group representation, as follows:

$$\rho : \Gamma \rightarrow C(T, U_K)$$

We conclude that in this case, the minimal factorization constructed in Definition 15.8 is simply the one obtained by taking the image:

$$\rho : \Gamma \rightarrow \Lambda \subset C(T, U_K)$$

Thus π is inner faithful when our group satisfies the following condition:

$$\Gamma \subset C(T, U_K)$$

As a second illustration now, given a compact group G , and elements $g_1, \dots, g_K \in G$, we have a representation $\pi : C(G) \rightarrow \mathbb{C}^K$, given by:

$$f \rightarrow (f(g_1), \dots, f(g_K))$$

The minimal factorization of π is then via $C(H)$, with $H \subset G$ being given by:

$$H = \overline{\langle g_1, \dots, g_K \rangle}$$

Thus π is inner faithful precisely when our group satisfies:

$$G = H$$

In general, the existence and uniqueness of the Hopf image comes from dividing $C(G)$ by a suitable ideal. In Tannakian terms, we have the following result:

THEOREM 15.9. *Assuming $G \subset U_N^+$, with fundamental corepresentation $u = (u_{ij})$, the Hopf image of $\pi : C(G) \rightarrow M_K(C(T))$ comes from the following Tannakian category,*

$$C_{kl} = \text{Hom}(U^{\otimes k}, U^{\otimes l})$$

where $U_{ij} = \pi(u_{ij})$, and where the spaces on the right are taken in a formal sense.

PROOF. Since the morphisms increase the intertwining spaces, when defined either in a representation theory sense, or just formally, we have inclusions as follows:

$$\text{Hom}(u^{\otimes k}, u^{\otimes l}) \subset \text{Hom}(U^{\otimes k}, U^{\otimes l})$$

More generally, we have such inclusions when replacing (G, u) with any pair producing a factorization of π . Thus, by Tannakian duality, the Hopf image must be given by the fact that the intertwining spaces must be the biggest, subject to the above inclusions. On the other hand, since u is biunitary, so is U , and it follows that the spaces on the right form a Tannakian category. Thus, we have a quantum group (H, v) given by:

$$\text{Hom}(v^{\otimes k}, v^{\otimes l}) = \text{Hom}(U^{\otimes k}, U^{\otimes l})$$

By the above discussion, $C(H)$ follows to be the Hopf image of π , as claimed. \square

The inner faithful models $\pi : C(G) \rightarrow M_K(C(T))$ are a very interesting notion, because they are not subject to the coamenability condition on G , as it was the case with the stationary models, as explained in Theorem 15.7. In fact, there are no known restrictions on the class of subgroups $G \subset U_N^+$ which can be modelled in an inner faithful way. Thus, our modelling theory applies a priori to any compact quantum group.

Regarding now the study of the inner faithful models, a key problem is that of computing the Haar integration functional. The result here is as follows:

THEOREM 15.10. *Given an inner faithful model $\pi : C(G) \rightarrow M_K(C(T))$, we have*

$$\int_G = \lim_{k \rightarrow \infty} \frac{1}{k} \sum_{r=1}^k \int_G^r$$

where $\int_G^r = (\varphi \circ \pi)^{*r}$, with $\varphi = \text{tr} \otimes \int_T$ being the random matrix trace.

PROOF. As a first observation, there is an obvious similarity here with the Woronowicz construction of the Haar measure, explained in chapter 7 above. In fact, the above result holds more generally for any model $\pi : C(G) \rightarrow B$, with $\varphi \in B^*$ being a faithful trace.

With this picture in hand, the Woronowicz construction simply corresponds to the case $\pi = id$, and the result itself is therefore a generalization of Woronowicz's result.

In order to prove now the result, we can proceed as in chapter 7. If we denote by \int'_G the limit in the statement, we must prove that this limit converges, and that we have:

$$\int'_G = \int_G$$

It is enough to check this on the coefficients of corepresentations, and if we let $v = u^{\otimes k}$ be one of the Peter-Weyl corepresentations, we must prove that we have:

$$\left(id \otimes \int'_G \right) v = \left(id \otimes \int_G \right) v$$

We know from chapter 7 that the matrix on the right is the orthogonal projection onto $Fix(v)$. Regarding now the matrix on the left, this is the orthogonal projection onto the 1-eigenspace of $(id \otimes \varphi\pi)v$. Now observe that, if we set $V_{ij} = \pi(v_{ij})$, we have:

$$(id \otimes \varphi\pi)v = (id \otimes \varphi)V$$

Thus, as in chapter 7, we conclude that the 1-eigenspace that we are interested in equals $Fix(V)$. But, according to Theorem 15.9, we have:

$$Fix(V) = Fix(v)$$

Thus, we have proved that we have $\int'_G = \int_G$, as desired. □

Summarizing, we have so far a notion of matrix model, and a notion of inner faithfulness which is quite broad, and that we can study via both algebra and analysis.

15c. Hadamard models

With this theory in hand, let us go back now to our von Neumann algebra and subfactor questions. In relation with the complex Hadamard matrices, the connection with the quantum permutations is immediate, coming from the following observation:

PROPOSITION 15.11. *If $H \in M_N(\mathbb{C})$ is Hadamard, the rank one projections*

$$P_{ij} = Proj \left(\begin{pmatrix} H_i \\ H_j \end{pmatrix} \right)$$

where $H_1, \dots, H_N \in \mathbb{T}^N$ are the rows of H , form a magic unitary.

PROOF. This is clear, the verification for the rows being as follows:

$$\begin{aligned} \left\langle \frac{H_i}{H_j}, \frac{H_i}{H_k} \right\rangle &= \sum_l \frac{H_{il}}{H_{jl}} \cdot \frac{H_{kl}}{H_{il}} \\ &= \sum_l \frac{H_{kl}}{H_{jl}} \\ &= N\delta_{jk} \end{aligned}$$

As for the verification for the columns, this is similar, as follows:

$$\begin{aligned} \left\langle \frac{H_i}{H_j}, \frac{H_k}{H_j} \right\rangle &= \sum_l \frac{H_{il}}{H_{jl}} \cdot \frac{H_{kl}}{H_{jl}} \\ &= \sum_l \frac{H_{il}}{H_{kl}} \\ &= N\delta_{ik} \end{aligned}$$

Thus, we have indeed a magic unitary, as claimed. \square

We are led in this way into the following notion:

DEFINITION 15.12. *To any Hadamard matrix $H \in M_N(\mathbb{C})$ we associate the quantum permutation group $G \subset S_N^+$ given by the following Hopf image factorization,*

$$\begin{array}{ccc} C(S_N^+) & \xrightarrow{\pi} & M_N(\mathbb{C}) \\ & \searrow & \nearrow \\ & C(G) & \end{array}$$

where $\pi(u_{ij}) = \text{Proj}(H_i/H_j)$, with $H_1, \dots, H_N \in \mathbb{T}^N$ being the rows of H .

Our claim now is that this construction $H \rightarrow G$ is something really useful, with the quantum group G encoding the combinatorics of H . To be more precise, the idea will be that “ H can be thought of as being a kind of Fourier matrix for G ”.

As an illustration for this principle, we first have the following result:

THEOREM 15.13. *The construction $H \rightarrow G$ has the following properties:*

- (1) *For a Fourier matrix $H = F_G$ we obtain the group G itself, acting on itself.*
- (2) *For $H \notin \{F_G\}$, the quantum group G is not classical, nor a group dual.*
- (3) *For a tensor product $H = H' \otimes H''$ we obtain a product, $G = G' \times G''$.*

PROOF. All this material is standard, and elementary, as follows:

(1) Let us first discuss the cyclic group case, $H = F_N$. Here the rows of H are given by $H_i = \rho^i$, where $\rho = (1, w, w^2, \dots, w^{N-1})$. Thus, we have the following formula:

$$\frac{H_i}{H_j} = \rho^{i-j}$$

It follows that the corresponding rank 1 projections $P_{ij} = Proj(H_i/H_j)$ form a circulant matrix, all whose entries commute. Since the entries commute, the corresponding quantum group must satisfy the following condition:

$$G \subset S_N$$

Now by taking into account the circulant property of $P = (P_{ij})$ as well, we are led to the conclusion that we have, as claimed:

$$G = \mathbb{Z}_N$$

In the general case now, where $H = F_G$, with G being an arbitrary finite abelian group, the result can be proved either by extending the above proof, or by decomposing $G = \mathbb{Z}_{N_1} \times \dots \times \mathbb{Z}_{N_k}$ and using (3) below, whose proof is independent from the rest.

(2) This is something more tricky, needing some general study of the representations whose Hopf images are commutative, or cocommutative. For details here, along with a number of supplementary facts on the construction $H \rightarrow G$, we refer to the literature.

(3) Assume that we have a tensor product $H = H' \otimes H''$, and let G, G', G'' be the associated quantum permutation groups. We have then a diagram as follows:

$$\begin{array}{ccccc} C(S_{N'}^+) \otimes C(S_{N''}^+) & \longrightarrow & C(G') \otimes C(G'') & \longrightarrow & M_{N'}(\mathbb{C}) \otimes M_{N''}(\mathbb{C}) \\ \uparrow & & & & \downarrow \\ C(S_N^+) & \longrightarrow & C(G) & \longrightarrow & M_N(\mathbb{C}) \end{array}$$

Here all the maps are the canonical ones, with those on the left and on the right coming from $N = N'N''$. At the level of standard generators, the diagram is as follows:

$$\begin{array}{ccccc} u'_{ij} \otimes u''_{ab} & \longrightarrow & w'_{ij} \otimes w''_{ab} & \longrightarrow & P'_{ij} \otimes P''_{ab} \\ \uparrow & & & & \downarrow \\ u_{ia,jb} & \longrightarrow & w_{ia,jb} & \longrightarrow & P_{ia,jb} \end{array}$$

Now observe that this diagram commutes. We conclude that the representation associated to H factorizes indeed through $C(G') \otimes C(G'')$, and this gives the result. \square

In order to discuss now the relation with the commuting squares and the subfactors, let us compute the representation theory of the quantum group G from Definition 15.12 above. At the Tannakian level, the result is as follows:

THEOREM 15.14. *The Tannakian category of the quantum group $G \subset S_N^+$ associated to a complex Hadamard matrix $H \in M_N(\mathbb{C})$ is given by*

$$T \in \text{Hom}(u^{\otimes k}, u^{\otimes l}) \iff T^\circ G^{k+2} = G^{l+2} T^\circ$$

where the objects on the right are constructed as follows:

- (1) $T^\circ = id \otimes T \otimes id.$
- (2) $G_{ia}^{jb} = \sum_k H_{ik} \bar{H}_{jk} \bar{H}_{ak} H_{bk}.$
- (3) $G_{i_1 \dots i_k, j_1 \dots j_k}^k = G_{i_k i_{k-1}}^{j_k j_{k-1}} \dots G_{i_2 i_1}^{j_2 j_1}.$

PROOF. With the notations in Theorem 15.9, we have the following formula:

$$\text{Hom}(u^{\otimes k}, u^{\otimes l}) = \text{Hom}(U^{\otimes k}, U^{\otimes l})$$

The vector space on the right consists by definition of the complex $N^l \times N^k$ matrices T , satisfying the following relation:

$$TU^{\otimes k} = U^{\otimes l}T$$

If we denote this equality by $L = R$, the left term L is given by:

$$\begin{aligned} L_{ij} &= (TU^{\otimes k})_{ij} \\ &= \sum_a T_{ia} U_{aj}^{\otimes k} \\ &= \sum_a T_{ia} U_{a_1 j_1} \dots U_{a_k j_k} \end{aligned}$$

As for the right term R , this is given by:

$$\begin{aligned} R_{ij} &= (U^{\otimes l}T)_{ij} \\ &= \sum_b U_{ib}^{\otimes l} T_{bj} \\ &= \sum_b U_{i_1 b_1} \dots U_{i_l b_l} T_{bj} \end{aligned}$$

Consider now the vectors $\xi_{ij} = H_i/H_j$. Since these vectors span the ambient Hilbert space, the equality $L = R$ is equivalent to the following equality:

$$\langle L_{ij} \xi_{pq}, \xi_{rs} \rangle = \langle R_{ij} \xi_{pq}, \xi_{rs} \rangle$$

We use now the following well-known formula, expressing a product of rank one projections P_1, \dots, P_k in terms of the corresponding image vectors ξ_1, \dots, ξ_k :

$$\langle P_1 \dots P_k x, y \rangle = \langle x, \xi_k \rangle \langle \xi_k, \xi_{k-1} \rangle \dots \langle \xi_2, \xi_1 \rangle \langle \xi_1, y \rangle$$

This gives the following formula for the left term L :

$$\begin{aligned}
 \langle L_{ij}\xi_{pq}, \xi_{rs} \rangle &= \sum_a T_{ia} \langle P_{a_1j_1} \dots P_{a_kj_k} \xi_{pq}, \xi_{rs} \rangle \\
 &= \sum_a T_{ia} \langle \xi_{pq}, \xi_{a_kj_k} \rangle \dots \langle \xi_{a_1j_1}, \xi_{rs} \rangle \\
 &= \sum_a T_{ia} G_{pa_k}^{qj_k} G_{a_k a_{k-1}}^{j_k j_{k-1}} \dots G_{a_2 a_1}^{j_2 j_1} G_{a_1 r}^{j_1 s} \\
 &= \sum_a T_{ia} G_{rap, sjq}^{k+2} \\
 &= (T^\circ G^{k+2})_{rip, sjq}
 \end{aligned}$$

As for the right term R , this is given by the following formula:

$$\begin{aligned}
 \langle R_{ij}\xi_{pq}, \xi_{rs} \rangle &= \sum_b \langle P_{i_1b_1} \dots P_{i_l b_l} \xi_{pq}, \xi_{rs} \rangle T_{bj} \\
 &= \sum_b \langle \xi_{pq}, \xi_{i_l b_l} \rangle \dots \langle \xi_{i_1 b_1}, \xi_{rs} \rangle T_{bj} \\
 &= \sum_b G_{pi_l}^{qb_l} G_{i_l i_{l-1}}^{b_l b_{l-1}} \dots G_{i_2 i_1}^{b_2 b_1} G_{i_1 r}^{b_1 s} T_{bj} \\
 &= \sum_b G_{rip, sbq}^{l+2} T_{bj} \\
 &= (G^{l+2} T^\circ)_{rip, sjq}
 \end{aligned}$$

Thus, we obtain the formula in the statement. □

The point now is that, with $k = 0$, we obtain in this way precisely the planar algebra spaces P_l computed by Jones in [44], for the corresponding commuting square, described in Theorem 15.4 above. Thus, we are led in this way to the following result:

THEOREM 15.15. *Let $H \in M_N(\mathbb{C})$ be a complex Hadamard matrix.*

- (1) *The planar algebra associated to H is given by the formula*

$$P_k = \text{Fix}(u^{\otimes k})$$

where $G \subset S_N^+$ is the associated quantum permutation group.

- (2) *The Poincaré series $\sum_k \dim(P_k)z^k$ equals the Stieltjes transform*

$$f(z) = \int_G \frac{1}{1 - z\chi}$$

of the law of the main character $\chi = \sum_i u_{ii}$.

PROOF. This follows as indicated above, by putting together what we have:

(1) As already mentioned above, this simply follows by comparing Theorem 15.14 with the subfactor computation in [47], discussed in Theorem 15.4.

(2) This follows from (1) and from the Peter-Weyl theory, with the statement itself being a nice and concrete application of our main result, (1) above. \square

Summarizing, in connection with commuting square problematics from the beginning of this chapter, the conclusion is that for the simplest such commuting squares, namely those coming from the Hadamard matrices, the combinatorics ultimately comes from quantum permutation groups. This is something nice, and exploring improvements and generalizations of this fact will be our main purpose, in the remainder of this chapter.

15d. Fixed points

We know from the above that the planar algebra associated to an Hadamard matrix $H \in M_N(\mathbb{C})$ appears in fact as the planar algebra associated to a certain related quantum permutation group $G \subset S_N^+$. In view of the various results from chapters 13-14 above, this suggests that the subfactor itself associated to H should appear as a fixed point subfactor associated to G . We will prove here that this is indeed the case.

To be more precise, regarding the subfactor itself, the result here is as follows, making the link with the general constructions of subfactors from chapters 13-14 above:

THEOREM 15.16. *The subfactor associated to $H \in M_N(\mathbb{C})$ is of the form*

$$A^G \subset (\mathbb{C}^N \otimes A)^G$$

with $A = R \rtimes \widehat{G}$, where $G \subset S_N^+$ is the associated quantum permutation group.

PROOF. This is something more technical, the idea being that the basic construction procedure for the commuting squares, explained before Theorem 15.3, can be performed in an “equivariant setting”, for commuting squares having components as follows:

$$D \otimes_G E = (D \otimes (E \rtimes \widehat{G}))^G$$

To be more precise, starting with a commuting square formed by such algebras, we obtain by basic construction a whole array of commuting squares as follows, with $\{D_i\}, \{E_i\}$

being by definition Jones towers, and with D_∞, E_∞ being their inductive limits:

$$\begin{array}{ccccccc}
 & D_0 \otimes_G E_\infty & & D_1 \otimes_G E_\infty & & D_2 \otimes_G E_\infty & \\
 & \uparrow \cdots & & \uparrow \cdots & & \uparrow \cdots & \\
 D_0 \otimes_G E_2 & \longrightarrow & D_1 \otimes_G E_2 & \longrightarrow & D_2 \otimes_G E_2 & \cdots \longrightarrow & D_\infty \otimes_G E_2 \\
 \uparrow & & \uparrow & & \uparrow & & \\
 D_0 \otimes_G E_1 & \longrightarrow & D_1 \otimes_G E_1 & \longrightarrow & D_2 \otimes_G E_1 & \cdots \longrightarrow & D_\infty \otimes_G E_1 \\
 \uparrow & & \uparrow & & \uparrow & & \\
 D_0 \otimes_G E_0 & \longrightarrow & D_1 \otimes_G E_0 & \longrightarrow & D_2 \otimes_G E_0 & \cdots \longrightarrow & D_\infty \otimes_G E_0
 \end{array}$$

The point now is that this quantum group picture works in fact for any commuting square having \mathbb{C} in the lower left corner. In the Hadamard matrix case, that we are interested in here, the corresponding commuting square is as follows:

$$\begin{array}{ccc}
 \mathbb{C} \otimes_G \mathbb{C}^N & \longrightarrow & \mathbb{C}^N \otimes_G \mathbb{C}^N \\
 \uparrow & & \uparrow \\
 \mathbb{C} \otimes_G \mathbb{C} & \longrightarrow & \mathbb{C}^N \otimes_G \mathbb{C}
 \end{array}$$

Thus, the subfactor obtained by vertical basic construction appears as follows:

$$\mathbb{C} \otimes_G E_\infty \subset \mathbb{C}^N \otimes_G E_\infty$$

But this gives the conclusion in the statement, with the II_1 factor appearing there being by definition $A = E_\infty \rtimes \widehat{G}$, and with the remark that we have $E_\infty \simeq R$. \square

All the above was of course quite brief, but we will discuss now all this with more details, directly in a more general setting, covering the Hadamard matrix situation.

To be more precise, our claim is that the above fixed point subfactor techniques apply, more generally, to the commuting squares having \mathbb{C} in the lower left corner:

$$\begin{array}{ccc}
 E & \longrightarrow & X \\
 \uparrow & & \uparrow \\
 \mathbb{C} & \longrightarrow & D
 \end{array}$$

In order to discuss this, let us go back to the fixed point subfactors, from chapter 13 above. In what concerns the fixed point factors, we know from there that we have:

THEOREM 15.17. Consider a Woronowicz algebra $A = (A, \Delta, S)$, and denote by A_σ the Woronowicz algebra $(A, \sigma\Delta, S)$, where σ is the flip. Given coactions

$$\beta : B \rightarrow B \otimes A$$

$$\pi : P \rightarrow P \otimes A_\sigma$$

with B being finite dimensional, the following linear map, while not being multiplicative in general, is coassociative with respect to the comultiplication $\sigma\Delta$ of A_σ ,

$$\beta \odot \pi : B \otimes P \rightarrow B \otimes P \otimes A_\sigma$$

$$b \otimes p \rightarrow \pi(p)_{23}((id \otimes S)\beta(b))_{13}$$

and its fixed point space, which is by definition the following linear space,

$$(B \otimes P)^{\beta \odot \pi} = \left\{ x \in B \otimes P \mid (\beta \odot \pi)x = x \otimes 1 \right\}$$

is then a von Neumann subalgebra of $B \otimes P$. Moreover, such algebras can be used in order to construct the generalized Wassermann subfactors, $(B_0 \otimes P)^G \subset (B_1 \otimes P)^G$.

PROOF. This is something that we know from chapter 13, and for details, and comments in relation with the non-multiplicativity of $\beta \odot \pi$, we refer to the material there. \square

Let $\int_A : A \rightarrow \mathbb{C}$ be the Haar functional, let $l^2(A)$ be its l^2 -space and let $\widehat{A} \subset B(l^2(A))$ be the dual algebra. If $\alpha : E \rightarrow E \otimes \widehat{A}$ is a coaction of \widehat{A} on a finite von Neumann algebra E , the crossed product $E \rtimes_\alpha \widehat{A}$ is the von Neumann subalgebra of $E \otimes B(l^2(A))$ generated by $\alpha(E)$ and by $1 \otimes A$. There exists a unique coaction $\widehat{\alpha}$ of A on $E \rtimes_\alpha \widehat{A}$ such that $(E \rtimes_\alpha \widehat{A})^{\widehat{\alpha}} = \alpha(E)$, and such that the copy $1 \otimes A$ of A is equivariant.

With these conventions, we have the following result:

PROPOSITION 15.18. Let A be a Woronowicz algebra. If $\beta : D \rightarrow D \otimes A$ is a coaction on a finite dimensional finite von Neumann algebra and $\alpha : E \rightarrow E \otimes \widehat{A}_\sigma$ is a coaction on a finite von Neumann algebra then we have the equality

$$(D \otimes (E \rtimes_\alpha \widehat{A}_\sigma))^{\beta \odot \widehat{\alpha}} = \overline{sp}^w \left\{ \beta(D)_{13} \cdot \alpha(E)_{23} \right\}$$

as linear subspaces of $D \otimes E \otimes B(l^2(A_\sigma))$. Moreover, the following diagram

$$\begin{array}{ccc} \alpha(E)_{23} & \subset & (D \otimes (E \rtimes_\alpha \widehat{A}_\sigma))^{\beta \odot \widehat{\alpha}} \\ \cup & & \cup \\ \mathbb{C} & \subset & \beta(D)_{13} \end{array}$$

is a non-degenerate commuting square of finite von Neumann algebras.

PROOF. By definition of the crossed product $E \rtimes_{\alpha} \widehat{A}_{\sigma}$, we have the following equalities between subalgebras of $D \otimes E \otimes B(l^2(A_{\sigma}))$:

$$\begin{aligned} D \otimes (E \rtimes_{\alpha} \widehat{A}_{\sigma}) &= D \otimes (\overline{sp}^w \{ \alpha(E) \cdot (1 \otimes A_{\sigma}) \}) \\ &= \overline{sp}^w \{ (D \otimes A_{\sigma})_{13} \cdot \alpha(E)_{23} \} \end{aligned}$$

On the other hand, since the coactions on the finite dimensional algebras are automatically non-degenerate, we have as well the following equality:

$$D \otimes A_{\sigma} = sp \{ (1 \otimes A_{\sigma}) \cdot \beta(D) \}$$

Thus, we have the following equality of algebras:

$$D \otimes (E \rtimes_{\alpha} A_{\sigma}) = \overline{sp}^w \{ (1 \otimes 1 \otimes A_{\sigma}) \cdot \beta(D)_{13} \cdot \alpha(E)_{23} \}$$

Let us compute now the restriction of the map $\beta \odot \widehat{\alpha}$ to the algebra $1 \otimes 1 \otimes A_{\sigma}$, to the algebra $\beta(D)_{13}$, and to the algebra $\alpha(E)_{23}$. This can be done as follows:

(1) The restriction of $\beta \odot \widehat{\alpha}$ to the algebra $1 \otimes 1 \otimes A_{\sigma}$ is $1 \otimes 1 \otimes \sigma\Delta$. In particular the map $\beta \odot \widehat{\alpha}$ has no fixed points in this algebra $1 \otimes 1 \otimes A_{\sigma}$.

(2) The algebra $\alpha(E)_{23}$ is by definition fixed by $\beta \odot \widehat{\alpha}$.

(3) We prove now that the algebra $\beta(D)_{13}$ is also fixed by $\beta \odot \widehat{\alpha}$. For this purpose, let $\{u_{ij}\}$ be an orthonormal basis of $l^2(A_{\sigma})$ consisting of coefficients of irreducible corepresentations of A_{σ} . Since we have $\beta(D) \subset D \otimes_{alg} A_{\sigma}$, for any $b \in D$ we can use the notation $\beta(b) = \sum_{uij} b_{ij}^u \otimes u_{ij}$. From the coassociativity of β we obtain:

$$\sum_{uij} \beta(b_{ij}^u) \otimes u_{ij} = \sum_{uijk} b_{ij}^u \otimes u_{kj} \otimes u_{ik}$$

Thus we have $\beta(b_{ik}^u) = \sum_j b_{ij}^u \otimes u_{kj}$ for any u, i, k , and so:

$$\begin{aligned} (id \otimes S)\beta(b_{ij}^u) &= (id \otimes S) \left(\sum_s b_{is}^u \otimes u_{js} \right) \\ &= \sum_s b_{is}^u \otimes u_{sj}^* \end{aligned}$$

Also, we have $\widehat{\alpha}(1 \otimes u_{ij}) = \sum_k 1 \otimes u_{ik} \otimes u_{kj}$, and we obtain from this that we have:

$$\begin{aligned} (\beta \odot \widehat{\alpha})(\beta(b)_{13}) &= \sum_{uij} \left(\sum_k 1 \otimes 1 \otimes u_{ik} \otimes u_{kj} \right) \left(\sum_s b_{is}^u \otimes 1 \otimes 1 \otimes u_{sj}^* \right) \\ &= \sum_{uijks} b_{is}^u \otimes 1 \otimes u_{ik} \otimes u_{kj} u_{sj}^* \end{aligned}$$

By summing over j the last term is replaced by $(uu^*)_{ks} = \delta_{k,s}1$. Thus we obtain, as desired, that our algebra consists indeed of fixed points:

$$\begin{aligned} (\beta \odot \widehat{\alpha})(\beta(b)_{13}) &= \sum_{uik} b_{ik}^u \otimes 1 \otimes u_{ik} \otimes 1 \\ &= (\beta(b)_{13}) \otimes 1 \end{aligned}$$

In order to finish now, observe that (1,2,3) above show that $(D \otimes (E \rtimes_{\alpha} \widehat{A}_{\sigma}))^{\beta \odot \widehat{\alpha}}$, which is the fixed point algebra of $\overline{sp}^w\{(1 \otimes 1 \otimes A_{\sigma}) \cdot \beta(D)_{13} \cdot \alpha(E)_{23}\}$ under the coaction $\beta \odot \widehat{\alpha}$, is equal to $\overline{sp}^w\{\beta(D)_{13} \cdot \alpha(E)_{23}\}$. This finishes the proof of the first assertion, and proves as well the non-degeneracy of the diagram in the statement.

Finally, observe that the diagram in the statement is the dual of the square on the left in the following diagram, where $P = E \rtimes_{\alpha} \widehat{A}_{\sigma}$ and $\pi = \widehat{\alpha}$:

$$\begin{array}{ccccc} D & \subset & (D \otimes P)^{\beta \odot \pi} & \subset & D \otimes P \\ \cup & & \cup & & \cup \\ \mathbb{C} & \subset & P^{\pi} & \subset & P \end{array}$$

Since π is dual, the square on the right is a non-degenerate commuting square. We also know that the rectangle is a non-degenerate commuting square. Thus if we denote by $E_X : D \otimes P \rightarrow D \otimes P$ the conditional expectation onto X , for any X , then for any $b \in D$ we have $E_{P^{\pi}}(b) = E_P(b) = E_{\mathbb{C}}(b)$, and this proves the commuting square condition. \square

Let us denote now by \mathcal{Alg} the category having as objects the finite dimensional C^* -algebras and having as arrows the inclusions of C^* -algebras which preserve the canonical traces. The above result suggests the following abstract definition:

DEFINITION 15.19. *Given objects $(D, \beta) \in A - \mathcal{Alg}$ and $(E, \alpha) \in \widehat{A}_{\sigma} - \mathcal{Alg}$, we let*

$$D \square_A E = (D \otimes (E \rtimes_{\alpha} \widehat{A}_{\sigma}))^{\beta \odot \widehat{\alpha}}$$

be the object in \mathcal{Alg} , constructed as in Proposition 15.18 above.

Observe that Proposition 15.18 above shows that the inclusion $\mathbb{C} \subset D \square_A E$ is Markov, so that the trace on $D \square_A E$ given by Proposition 15.18 is indeed its canonical trace.

If $(D', \beta') \subset (D, \beta)$ is an arrow in $A - \mathcal{Alg}$ and $(E', \alpha') \subset (E, \alpha)$ is an arrow in $\widehat{A}_{\sigma} - \mathcal{Alg}$, then we have a canonical embedding, as follows:

$$D' \square_A E' \subset D \square_A E$$

Now since both $D' \square_A E'$ and $D \square_A E$ are endowed with their canonical traces, this inclusion is Markov. Thus, we have constructed a bifunctor, as follows:

$$\square_A : A - \mathcal{Alg} \times \widehat{A}_{\sigma} - \mathcal{Alg} \rightarrow \mathcal{Alg}$$

With this convention, we have the following result, from [2]:

THEOREM 15.20. For any two arrows $D_0 \subset D_1$ in $A\text{-Alg}$ and $E_0 \subset E_1$ in $\widehat{A}_\sigma\text{-Alg}$,

$$\begin{array}{ccc} D_0 \square_A E_1 & \subset & D_1 \square_A E_1 \\ \cup & & \cup \\ D_0 \square_A E_0 & \subset & D_1 \square_A E_0 \end{array}$$

is a non-degenerate commuting square of finite dimensional von Neumann algebras.

PROOF. This can be proved in several steps, as follows:

Step I. In the simplest case, $D_0 = E_0 = \mathbb{C}$, this follows from the above.

Step II. We prove now the result in the general $E_0 = \mathbb{C}$ case. Indeed, let $E = E_1$, and consider the following diagram:

$$\begin{array}{ccccc} E & \subset & D_0 \square_A E & \subset & D_1 \square_A E \\ \cup & & \cup & & \cup \\ \mathbb{C} & \subset & D_0 & \subset & D_1 \end{array}$$

By the result of Step I, both the square on the left and the rectangle are non-degenerate commuting squares. We want to prove that the square on the right is a non-degenerate commuting square. But the non-degeneracy condition follows from:

$$D_1 \square_A E = sp\{E \cdot D_1\} \subset sp\{D_0 \square_A E \cdot D_1\}$$

Now let $x \in D_0 \square_A E$ and write $x = \sum_i b_i a_i$ with $b_i \in D_0$ and $a_i \in E$. Then:

$$\begin{aligned} E_{D_1}(x) &= \sum_i b_i E_{D_1}(a_i) \\ &= \sum_i b_i E_{\mathbb{C}}(a_i) \\ &= \sum_i b_i E_{D_0}(a_i) \\ &= E_{D_0}(x) \end{aligned}$$

But this proves the commuting square condition, and we are done.

Step III. A similar argument shows that the result holds in the case $D_0 = \mathbb{C}$.

Step IV. General case. We will use many times the following diagram, in which all the rectangles and all the squares, except possibly for the square in the statement, are non-degenerate commuting squares, cf. the conclusions of Steps I, II, III:

$$\begin{array}{ccccc} E_1 & \subset & D_0 \square_A E_1 & \subset & D_1 \square_A E_1 \\ \cup & & \cup & & \cup \\ E_0 & \subset & D_0 \square_A E_0 & \subset & D_1 \square_A E_0 \\ \cup & & \cup & & \cup \\ \mathbb{C} & \subset & D_0 & \subset & D_1 \end{array}$$

The non-degeneracy condition follows from:

$$D_1 \square_A E_1 = sp\{E_1 \cdot D_1\} \subset sp\{D_0 \square_A E_1 \cdot D_1 \square_A E_0\}$$

Now let $x \in D_0 \square_A E_1$ and write $x = \sum_i b_i a_i$ with $b_i \in D_0$ and $a_i \in E_1$. Then:

$$\begin{aligned} E_{D_1 \square_A E_0}(x) &= \sum_i b_i E_{D_1 \square_A E_0}(a_i) \\ &= \sum_i b_i E_{E_0}(a_i) \\ &= \sum_i b_i E_{D_0 \square_A E_0}(a_i) \\ &= E_{D_0 \square_A E_0}(x) \end{aligned}$$

But this proves the commuting square condition, and we are done. □

We show now that the bifunctor \square_A behaves well with respect to basic constructions. If A is a Woronowicz algebra, a sequence of two arrows $D_0 \subset D_1 \subset D_2$ in $A - Alg$ is called a basic construction if $D_0 \subset D_1 \subset D_2$ is a basic construction in Alg and if its Jones projection $e \in D_2$ is a fixed by the coaction $D_2 \rightarrow D_2 \otimes A$. An infinite sequence of basic constructions in $A - Alg$ is called a Jones tower in $A - Alg$. We have:

PROPOSITION 15.21. *If $D_0 \subset D_1 \subset D_2 \subset D_3 \subset \dots$ is a Jones tower in $A - Alg$ and $E_0 \subset E_1 \subset E_2 \subset E_3 \subset \dots$ is a Jones tower in $\widehat{A}_\sigma - Alg$ then*

$$\begin{array}{cccc} \vdots & \vdots & \vdots & \\ \cup & \cup & \cup & \\ D_0 \square_A E_2 & \subset & D_1 \square_A E_2 & \subset & D_2 \square_A E_2 & \subset & \dots \\ \cup & & \cup & & \cup & & \\ D_0 \square_A E_1 & \subset & D_1 \square_A E_1 & \subset & D_2 \square_A E_1 & \subset & \dots \\ \cup & & \cup & & \cup & & \\ D_0 \square_A E_0 & \subset & D_1 \square_A E_0 & \subset & D_2 \square_A E_0 & \subset & \dots \end{array}$$

is a lattice of basic constructions for non-degenerate commuting squares.

PROOF. We prove only that the rows are Jones towers, the proof for the columns being similar. By restricting the attention to a pair of consecutive inclusions, it is enough to prove that if $D_0 \subset D_1 \subset D_2$ is a basic construction in $A - Alg$ and E is an object of $\widehat{A}_\sigma - Alg$ then $D_0 \square_A E \subset D_1 \square_A E \subset D_2 \square_A E$ is a basic construction in Alg .

For this purpose, we will use many times the following diagram, in which all squares and rectangles are non-degenerate commuting squares:

$$\begin{array}{ccccccc} E & \subset & D_0 \square_A E & \subset & D_1 \square_A E & \subset & D_2 \square_A E \\ \cup & & \cup & & \cup & & \cup \\ \mathbb{C} & \subset & D_0 & \subset & D_1 & \subset & D_2 \end{array}$$

We will use the abstract characterization of the basic construction, stating that $N \subset M \subset P$ is a basic construction, with Jones projection $e \in P$, precisely when:

- (1) $P = sp\{M \cdot e \cdot M\}$.
- (2) $[e, N] = 0$.
- (3) $exe = E_N(x)e$ for any $x \in M$.
- (4) $tr(xe) = \lambda tr(x)$ for any $x \in M$, where λ is the inverse of the index of $N \subset M$.

Let $e \in D_2$ be the Jones projection for the basic construction $D_0 \subset D_1 \subset D_2$. With $N = D_0$, $M = D_1$ and $P = D_2$ the verification of (1-4) is as follows:

- (1) This follows from the following computation:

$$\begin{aligned} D_2 \square_A E &= sp\{D_2 \cdot E\} \\ &= sp\{D_1 \cdot e \cdot D_1 \cdot E\} \\ &= sp\{D_1 \cdot e \cdot D_1 \square_A E\} \end{aligned}$$

- (2) This follows from $D_0 \square_A E = sp\{D_0 \cdot E\}$, from $[e, E] = 0$ and from $[e, D_0] = 0$.

- (3) Let $x \in D_1 \square_A E$, and write $x = \sum_i b_i a_i$ with $b_i \in D_1$ and $a_i \in E$. Then:

$$\begin{aligned} exe &= \sum_i e b_i a_i e \\ &= \sum_i e b_i e a_i \\ &= \sum_i E_{D_0}(b_i) e a_i \\ &= \sum_i E_{D_0}(b_i) a_i e \end{aligned}$$

On the other hand, we have as well the following computation:

$$\begin{aligned} E_{D_0 \square_A E}(x)e &= \sum_i E_{D_0 \square_A E}(b_i a_i) e \\ &= \sum_i E_{D_0 \square_A E}(b_i) a_i e \\ &= \sum_i E_{D_0}(b_i) a_i e \end{aligned}$$

Thus, the third condition for a basic construction is verified, indeed.

(4) With the above notations, we have that:

$$\begin{aligned} E_{D_2}(xe) &= \sum_i E_{D_2}(b_i a_i e) \\ &= \sum_i b_i E_{D_2}(a_i) e \\ &= \sum_i b_i E_{\mathbb{C}}(a_i) e \end{aligned}$$

We also have $b_i E_{\mathbb{C}}(a_i) \in D_1$ for every i , and so:

$$\begin{aligned} tr_{D_2 \square_A E}(xe) &= tr_{D_2}(E_{D_2}(xe)) \\ &= \lambda \sum_i tr_{D_1}(b_i E_{\mathbb{C}}(a_i)) \end{aligned}$$

On the other hand, we have as well the following computation:

$$\begin{aligned} tr_{D_1 \square_A E}(x) &= tr_{D_1}(E_{D_1}(x)) \\ &= \sum_i tr_{D_1}(b_i E_{D_1}(a_i)) \\ &= \sum_i tr_{D_1}(b_i E_{\mathbb{C}}(a_i)) \end{aligned}$$

Thus, the fourth condition for a basic construction is verified, as desired. \square

With standard coaction conventions, from chapter 13, we have:

PROPOSITION 15.22. *Given a corepresentation and a representation, as follows,*

$$v \in M_n(\mathbb{C}) \otimes A \quad , \quad \pi : A_\sigma \rightarrow M_k(\mathbb{C})$$

consider, via some standard identifications, the associated objects

$$(M_n(\mathbb{C}), \iota_v) \in A - \mathcal{Alg} \quad , \quad (M_k(\mathbb{C}), \iota_{\tilde{\pi}}) \in \widehat{A}_\sigma - \mathcal{Alg}$$

and form the corresponding algebra $M_n(\mathbb{C}) \square_A M_k(\mathbb{C})$. Then there exists an isomorphism

$$\left(\begin{array}{ccc} M_k(\mathbb{C}) & \subset & M_n(\mathbb{C}) \square_A M_k(\mathbb{C}) \\ \cup & & \cup \\ \mathbb{C} & \subset & M_n(\mathbb{C}) \end{array} \right) \simeq \left(\begin{array}{ccc} \mathbb{C} \otimes M_k(\mathbb{C}) & \subset & M_n(\mathbb{C}) \otimes M_k(\mathbb{C}) \\ \cup & & \cup \\ \mathbb{C} & \subset & u(M_n(\mathbb{C}) \otimes \mathbb{C})u^* \end{array} \right)$$

sending $z \rightarrow 1 \otimes z$ for $z \in M_k(\mathbb{C})$ and $y \rightarrow \iota_u(y)$ for $y \in M_n(\mathbb{C})$, where $u = (id \otimes \pi)v$.

PROOF. Consider the following $*$ -morphism of algebras:

$$\begin{aligned} \Phi : M_n(\mathbb{C}) \otimes M_k(\mathbb{C}) &\rightarrow M_n(\mathbb{C}) \otimes M_k(\mathbb{C}) \otimes B(l^2(A_\sigma)) \\ x &\rightarrow ad(v_{13} \tilde{\pi}_{23} u_{12}^*)(x \otimes 1) \end{aligned}$$

Since both the squares in the statement are non-degenerate commuting squares, all the assertions are consequences of the following formulae, that we will prove now:

$$\Phi(1 \otimes z) = \iota_{\tilde{\pi}}(z)_{23} \quad , \quad \Phi(\iota_u(y)) = \iota_v(y)_{13}$$

The second formula follows from the following computation:

$$\begin{aligned} \Phi(u(y \otimes 1)u^*) &= v_{13}(y \otimes 1 \otimes 1)v_{13}^* \\ &= (v(y \otimes 1)v^*)_{13} \end{aligned}$$

For the first formula, what we have to prove is that:

$$v_{13}\tilde{\pi}_{23}u_{12}^*(1 \otimes z \otimes 1)u_{12}\tilde{\pi}_{23}^*v_{13}^* = (\tilde{\pi}(z \otimes 1)\tilde{\pi}^*)_{23}$$

By moving the unitaries to the left and to the right we have to prove that:

$$\tilde{\pi}_{23}^*v_{13}\tilde{\pi}_{23}u_{12}^* \in (\mathbb{C} \otimes M_k(\mathbb{C}) \otimes \mathbb{C})' = M_n(\mathbb{C}) \otimes \mathbb{C} \otimes B(l^2(H_\sigma))$$

Let us call this unitary U . Since $\tilde{\pi} = (\pi \otimes id)V'$ we have:

$$U = (id \otimes \pi \otimes id)(V_{23}^*v_{13}V_{23}v_{12}^*)$$

The comultiplication of H_σ is given by $\Delta(y) = V^*(1 \otimes y)V$. On the other hand since v^* is a corepresentation of H_σ , we have $(id \otimes \Delta)(v^*) = v_{12}^*v_{13}^*$. We get that:

$$\begin{aligned} V_{23}^*v_{13}V_{23} &= (V_{23}^*v_{13}^*V_{23})^* \\ &= ((id \otimes \Delta)(v^*))^* \\ &= (v_{12}^*v_{13}^*)^* \\ &= v_{13}v_{12} \end{aligned}$$

Thus we have $U = v_{13}$, and we are done. □

We are now ready to formulate our main result, as follows:

THEOREM 15.23. *Any commuting square having \mathbb{C} in the lower left corner,*

$$\begin{array}{ccc} E & \subset & X \\ \cup & & \cup \\ \mathbb{C} & \subset & D \end{array}$$

must appear as follows, for a suitable Woronowicz algebra A , with actions on D, E ,

$$\begin{array}{ccc} E & \subset & D \square_A E \\ \cup & & \cup \\ \mathbb{C} & \subset & D \end{array}$$

and the vertical subfactor associated to it is isomorphic to

$$R \subset (D \otimes (R \rtimes_{\gamma_\pi} \widehat{A}_\sigma))^{\beta_v \circ \widehat{\gamma}_\pi}$$

which is a fixed point subfactor, in the sense of chapter 13 above.

PROOF. This is something quite technical, which basically follows by combining the above results, and for full details on all this, we refer to [2], [3] and related papers. \square

Summarizing, all commuting squares having \mathbb{C} in the lower left corner are described by a quantum group. This is of course something quite special, and we will study more complicated commuting squares, not coming from quantum groups, in the next chapter.

15e. Exercises

Things have been quite technical here, and as an exercise on this, we have:

EXERCISE 15.24. *Given a commuting square of finite dimensional algebras*

$$\begin{array}{ccc} C_{01} & \longrightarrow & C_{11} \\ \uparrow & & \uparrow \\ C_{00} & \longrightarrow & C_{10} \end{array}$$

establish, with full details, the Ocneanu compactness formula

$$A'_0 \cap A_k = C'_{01} \cap C_{k0}$$

for the associated vertical subfactor $A_0 \subset A_1$.

This is something quite fundamental, that we discussed in the above, but with some details missing. Time to have this done, by working out all the linear algebra.

CHAPTER 16

Spectral measures

16a. Small index

We have seen so far the foundations of Jones' subfactor theory, along with results regarding the most basic classes of such subfactors, namely those coming from compact groups, discrete group duals, and more generally compact quantum groups. These subfactors all have integer index, $N \in \mathbb{N}$, and appear as subfactors of the Murray-von Neumann hyperfinite II_1 factor R , either by definition, or by theorem, or by conjecture.

This suggests looking into the classification of subfactors of integer index, or into the classification of the subfactors of R , or into the classification of the subfactors of R having integer index. These are all good questions, that we will discuss here.

Before starting, however, and in order to have an idea on what we want to do, we should discuss the following question: should the index $N \in [1, \infty)$ be small, or big? This is something quite philosophical, and non-trivial, the situation being as follows:

- (1) Mathematics and basic common sense suggest that subfactors should fall into two main classes, "series" and "exceptional". From this perspective, the series, corresponding to uniform values of the index, must be investigated first.
- (2) In practice now, passed a few simple cases, such as the FC or TL subfactors, we cannot hope for the index to take full uniform values. The more reasonable question here is that of looking at the case where $N \in \mathbb{N}$ is uniform.
- (3) The problem now is that, in the lack of theory here, this basically brings us back to groups, group duals, and more generally compact quantum groups, whose combinatorics is notoriously simpler than that of the arbitrary subfactors.
- (4) In short, naivity and pure mathematics tell us to investigate the "big index" case first, but with the remark however that we are missing something, and so that we must do in parallel some study in the "small index" case too.

All this does not look very clear, and so after this discussion, we are basically still in the dark. So, should the answer come then from physics, and applications?

Unfortunately, things here are quite complicated too, basically due to our current poor understanding of quantum mechanics, and of what precisely is to be done, in order to

have things in physics moving. And in fact, things here are in fact split too, a bit in the same way as above, the situation being basically as follows:

- (1) The very small index range, $N \in [1, 4]$, is subject to the remarkable “quantization” result of Jones, stating that we should have $N = 4 \cos^2(\frac{\pi}{n})$, and has strong ties with a number of theories in physics, such as conformal field theory.
- (2) In what concerns the other end, $N \gg 0$, this is in relation with statistical mechanics, once again following work of Jones on the subject, and with the index itself presumably corresponding to physicists’ famous “big N ” variable.

In short, no hope for an answer here. At least with our current knowledge of the subject. Probably most illustrating here is the fact that the main experts, starting with Jones himself, have always being split, working on both small and big index.

Getting away now from these philosophical difficulties, and back to our present book, which is rather elementary and mathematical, in this final chapter we will survey the main structure and classification results available, both in small and big index.

As already mentioned, we will focus on the subfactors of the Murray-von Neumann hyperfinite II_1 factor R , by taking for granted the fact that these subfactors are the most “important”, and related to physics. With the side remark, however, that this is actually subject to debate too, with many mathematicians opting for bigger factors like $L(F_\infty)$, and with some physicists joining them too. But let us not get into this here.

In order to get started now, in order to talk about classification, we need invariants for our subfactors. Which brings us into a third controversy, namely the choice between algebraic and analytic invariants. The situation here is as follows:

DEFINITION 16.1. *Associated to any finite index subfactor $A \subset B$, having planar algebra $P = (P_k)$, are the following invariants:*

- (1) *Its principal graph X , which describes the inclusions $P_0 \subset P_1 \subset P_2 \subset \dots$, with the reflections coming from basic constructions removed.*
- (2) *Its fusion algebra F , which describes the fusion rules for the various types of bimodules that can appear, namely $A - A$, $A - B$, $B - A$, $B - B$.*
- (3) *Its Poincaré series f , which is the generating series of the graded components of the planar algebra, $f(z) = \sum_k \dim(P_k)z^k$.*
- (4) *Its spectral measure μ , which is the probability measure having as moments the dimensions of the planar algebra components, $\int x^k d\mu(x) = \dim(P_k)$.*

This definition is of course something a bit informal, and there is certainly some work to be done, in order to fully define all the above invariants X, F, f, μ , and to work out the precise relation between them. We will be back to this later, but for the moment, let us keep in mind the fact that associated to a given subfactor $A \subset B$ are several combinatorial

invariants, which are not exactly equivalent, but are definitely versions of the same thing, the “combinatorics of the subfactor”, and which come in algebraic or analytic flavors.

So, what to use? As before, in relation with the previous controversies, the main experts, starting with Jones himself, have always being split themselves on this question, working with both algebraic and analytic invariants. Generally speaking, the algebraic invariants, which are (1) and (2) in the above list, tend to be more popular in small index, while the analytic invariants, (3) and (4), are definitely more popular in big index.

In order to get started now, let us first discuss the question of classifying the subfactors of the hyperfinite II_1 factor R , up to isomorphism, having index $N \leq 4$.

This is something quite tricky, and the main idea here will be the fact, coming from the proof of the Jones index restriction theorem, explained in chapter 13 above, that the index $N \in (1, 4]$ must be the squared norm of a certain graph:

$$N = \|X\|^2$$

Now with this observation in hand, the assumption $N \leq 4$ forces X to be one of the Coxeter-Dynkin graphs of type ADE, and then a lot of work, both of classification and exclusion, leads to an ADE classification for the subfactors of R having index $N \leq 4$.

This was for the idea. More in detail now, let us begin by explaining in detail how our subfactor invariant here, which will be the principal graph X , is constructed.

Consider first an arbitrary finite index irreducible subfactor $A_0 \subset A_1$, with associated planar algebra $P_k = A'_0 \cap A_k$, and let us look at the following system of inclusions:

$$P_0 \subset P_1 \subset P_2 \subset \dots$$

By taking the Bratelli diagram of this system of inclusions, and then deleting the reflections coming from basic constructions, which automatically appear at each step, according to the various results from chapter 13, we obtain a certain graph X , called principal graph of $A_0 \subset A_1$. The main properties of X can be summarized as follows:

THEOREM 16.2. *The principal graph X has the following properties:*

- (1) *The higher relative commutant $P_k = A'_0 \cap A_k$ is isomorphic to the abstract vector space spanned by the $2k$ -loops on X based at the root.*
- (2) *In the amenable case, where $A_1 = R$ and when the subfactor is “amenable”, the index of $A_0 \subset A_1$ is given by $N = \|X\|^2$.*

PROOF. This is something standard, the idea being as follows:

- (1) The statement here, which explains among others the relation between the principal graph X , and the other subfactor invariants, from Definition 16.1 above, comes from the definition of the principal graph, as a Bratelli diagram, with the reflections removed.

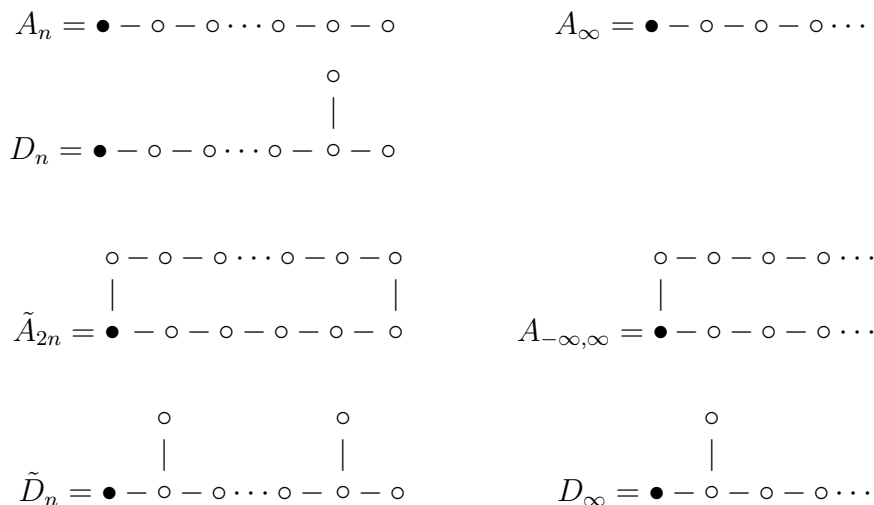
(2) This is actually a quite subtle statement, but for our purposes here, we can take the equality $N = \|X\|^2$, which reminds a bit the Kesten amenability condition for discrete groups, as a definition for the amenability of the subfactor. With the remark that for the Popa diagonal subfactors what we have here is precisely the Kesten amenability condition for the underlying discrete group Γ , and that, more generally, for the arbitrary generalized Popa or Wassermann subfactors, what we have here is precisely the Kesten type amenability condition for the underlying discrete quantum group Γ . \square

As a consequence of the above, in relation with classification questions, we have:

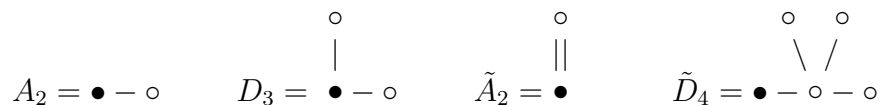
THEOREM 16.3. *The principal graph of a subfactor having index $N \leq 4$ must be one of the Coxeter-Dynkin graphs of type ADE.*

PROOF. This follows indeed from the formula $N = \|X\|^2$ from the above result, and from the considerations from the proof of the Jones index restriction theorem, explained in chapter 13 above. For full details on all this, we refer for instance to [35]. \square

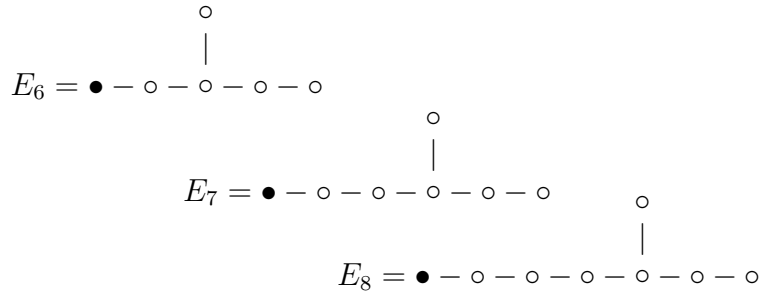
More in detail now, the usual Coxeter-Dynkin graphs are as follows:



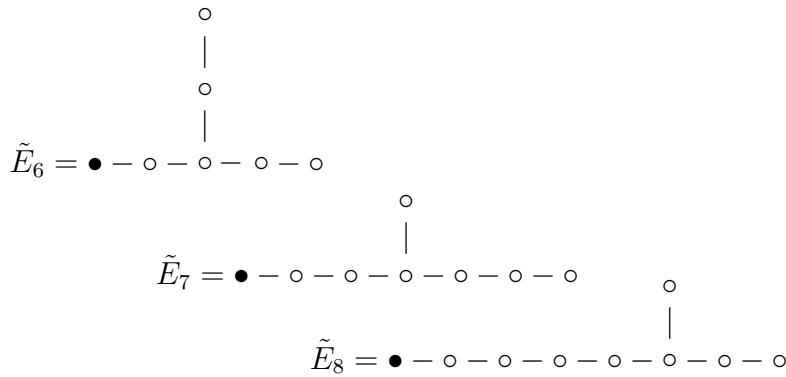
Here the graphs A_n with $n \geq 2$ and D_n with $n \geq 3$ have by definition n vertices each, \tilde{A}_{2n} with $n \geq 1$ has $2n$ vertices, and \tilde{D}_n with $n \geq 4$ has $n + 1$ vertices. Thus, the first graph in each series is by definition as follows:



There are also a number of exceptional Coxeter-Dynkin graphs. First we have:



Also, we have as well index 4 versions of the above exceptional graphs, as follows:



Getting back now to Theorem 16.3, with this list in hand, the story is not over here, because we still have to understand which of these graphs can really appear as principal graphs of subfactors. And, for those graphs which can appear, we must understand the structure and classification of the subfactors of R , having them as principal graphs.

In short, there is still a lot of work to be done, as a continuation of Theorem 16.3. The subfactors of index ≤ 4 were intensively studied in the 80s and early 90s, and about 10 years after Jones' foundational paper [43], a complete classification result was found, with contributions by many authors. A simplified form of this result is as follows:

THEOREM 16.4. *The principal graphs of subfactors of index ≤ 4 are:*

- (1) *Index < 4 graphs:* $A_n, D_{\text{even}}, E_6, E_8$.
- (2) *Index 4 finite graphs:* $\tilde{A}_{2n}, \tilde{D}_n, \tilde{E}_6, \tilde{E}_7, \tilde{E}_8$.
- (3) *Index 4 infinite graphs:* $A_\infty, A_{-\infty, \infty}, D_\infty$.

PROOF. As already mentioned, this is something quite heavy, with contributions by many authors, and among the main papers to be read here, let us mention [39], [41], [43], [64], [65], [70]. Observe that the graphs D_{odd} and E_7 don't appear in the above list. This is one of the subtle points of subfactor theory. For a discussion here, see [32]. \square

There are many other things that can be said about the subfactors of index $N \leq 4$, both at the theoretical level, of the finite depth and more generally of the amenable subfactors, and at the level of the ADE classification, which makes connections with other ADE classifications. We refer here to [8], [32], [35], [64], [65], [70], [71].

Regarding now the subfactors of index $N \in (4, 5]$, and also of small index above 5, these can be classified, but this is a long and complicated story. Let us just record here the result in index 5, which is something quite easy to formulate, as follows:

THEOREM 16.5. *The principal graphs of the irreducible index 5 subfactors are:*

- (1) A_∞ , and a non-extremal perturbation of $A_\infty^{(1)}$.
- (2) The McKay graphs of $\mathbb{Z}_5, D_5, GA_1(5), A_5, S_5$.
- (3) The twists of the McKay graphs of A_5, S_5 .

PROOF. This is a heavy result, coming from the work in [1], [35], [38], [46], and then in [40], [42], [58], [59], [67], and we refer to [51] for the whole story. The above formulation is the one from [51], with the subgroup subfactors there replaced by fixed point subfactors, and with the cyclic groups denoted as usual by \mathbb{Z}_N . \square

As a comment here, the above $N = 5$ result was much harder to obtain than the classification in index $N = 4$, obtained as a consequence of Theorem 16.4 above. However, at the level of the explicit construction of such subfactors, things are quite similar at $N = 4$ and $N = 5$, with the fixed point subfactors associated to quantum permutation groups $G \subset S_N^+$ providing most of the examples. We refer here to [10] and related papers.

In index $N = 6$ now, the subfactors cannot be classified, at least in general, due to several uncountable families, coming from groups, group duals, and more generally compact quantum groups. The exact assumption to be added is not known yet.

Summarizing, the current small index classification problem meets considerable difficulties in index $N = 6$, and right below. In small index $N > 6$ the situation is largely unexplored. We refer here to [18], [54], and the recent literature on the subject.

16b. Spectral measures

Before getting into the case where the index is big, $N \gg 0$, let us comment on one of the key ingredients for the above classification results, at $N < 6$. This is the Jones annular theory of subfactors, which is something very beautiful and useful, regarding the case where the index is arbitrary, $N \in [1, \infty)$. The main result is as follows:

THEOREM 16.6. *The theta series of a subfactor of index $N > 4$, which is given by*

$$\Theta(q) = q + \frac{1-q}{1+q} f\left(\frac{q}{(1+q)^2}\right)$$

with $f = \sum_k \dim(P_k)z^k$ being the Poincaré series, has positive coefficients.

PROOF. This is something quite advanced, the idea being that Θ is the generating series of a certain series of multiplicities associated to the subfactor, and more specifically associated to the canonical inclusion $TL_N \subset P$. We refer here to Jones' paper [49]. \square

In relation to this, and to some questions from physics as well, coming from conformal field theory, an interesting question is that of computing the “blowup” of the spectral measure of the subfactor, via the Jones change of variables, namely:

$$z \rightarrow \frac{q}{1 + q^2}$$

This question makes sense in any index, meaning both $N \in [1, 4]$, where Theorem 16.6 does not apply, and $N \in (4, \infty)$, where Theorem 16.6 does apply. We will discuss in what follows both these questions, by starting with the small index one, $N \in [1, 4]$.

Following [11] and related papers, it is convenient to stay, at least for the beginning, at a purely elementary level, and associate such series to any rooted bipartite graph. Let us start with the following definition, which is something straightforward, inspired by the definition of the Poincaré series of a subfactor, and by Theorem 16.2 above:

DEFINITION 16.7. *The Poincaré series of a rooted bipartite graph X is*

$$f(z) = \sum_{k=0}^{\infty} \text{loop}_X(2k)z^k$$

where $\text{loop}_X(2k)$ is the number of $2k$ -loops based at the root.

In the case where X is the principal graph of a subfactor $A_0 \subset A_1$, this series f is the Poincaré series of the subfactor, in the usual sense:

$$f(z) = \sum_{k=0}^{\infty} \dim(A'_0 \cap A_k)z^k$$

In general, the Poincaré series should be thought of as being a basic representation theory invariant of the underlying group-like object. For instance for the Wassermann type subfactor associated to a compact Lie group $G \subset U_N$, the Poincaré series is:

$$f(z) = \int_G \frac{1}{1 - \text{Tr}(g)z} dg$$

Regarding now the theta series, this can be introduced as a version of the Poincaré series, via the change of variables $z^{-1/2} = q^{1/2} + q^{-1/2}$, as follows:

DEFINITION 16.8. *The theta series of a rooted bipartite graph X is*

$$\Theta(q) = q + \frac{1 - q}{1 + q} f\left(\frac{q}{(1 + q)^2}\right)$$

where f is the Poincaré series.

The theta series can be written as $\Theta(q) = \sum a_r q^r$, and it follows from the above formula, via some simple manipulations, that its coefficients are integers:

$$a_r \in \mathbb{Z}$$

In fact, we have the following explicit formula from Jones' paper [49], relating the coefficients of $\Theta(q) = \sum a_r q^r$ to those of the Poincaré series $f(z) = \sum c_k z^k$:

$$a_r = \sum_{k=0}^r (-1)^{r-k} \frac{2r}{r+k} \binom{r+k}{r-k} c_k$$

In the case where X is the principal graph of a subfactor $A_0 \subset A_1$ of index $N > 4$, it is known from [49] that the numbers a_r are certain multiplicities associated to the planar algebra inclusion $TL_N \subset P$, as explained in Theorem 16.6 above and its proof. In particular, the coefficients of the theta series are in this case positive integers:

$$a_r \in \mathbb{N}$$

Before getting into computations, let us discuss as well the measure-theoretic versions of the above invariants. Once again, we start with an arbitrary rooted bipartite graph X . We can first introduce a measure μ , whose Stieltjes transform is f , as follows:

DEFINITION 16.9. *The real measure μ of a rooted bipartite graph X is given by*

$$f(z) = \int_0^\infty \frac{1}{1-xz} d\mu(x)$$

where f is the Poincaré series.

In the case where X is the principal graph of a subfactor $A_0 \subset A_1$, we recover in this way the spectral measure of the subfactor, as introduced in Definition 16.1 above, with the remark however that the existence of such a measure μ was not discussed there. In general, and so also in the particular subfactor case, clarifying the things here, the fact that μ as above exists indeed comes from the following simple fact:

PROPOSITION 16.10. *The real measure μ of a rooted bipartite graph X is given by the following formula, where $L = MM^t$, with M being the adjacency matrix of the graph,*

$$\mu = \text{law}(L)$$

and with the probabilistic computation being with respect to the expectation

$$A \rightarrow \langle A \rangle$$

with $\langle A \rangle$ being the $(*,*)$ -entry of a matrix A , where $*$ is the root.

PROOF. With the conventions in the statement, namely $L = MM^t$, with M being the adjacency matrix, and with $\langle A \rangle$ being the $(*, *)$ -entry of a matrix A , we have:

$$\begin{aligned} f(z) &= \sum_{k=0}^{\infty} \text{loop}_X(2k)z^k \\ &= \sum_{k=0}^{\infty} \langle L^k \rangle z^k \\ &= \left\langle \frac{1}{1 - Lz} \right\rangle \end{aligned}$$

But this shows that we have the formula $\mu = \text{law}(L)$, as desired. □

In the subfactor case some further interpretations are available as well. For instance in the case of the fixed point subfactors coming from of a compact group $G \subset U_N$, discussed after Definition 16.7 above, μ is the spectral measure of the main character:

$$\mu = \text{law}(\chi)$$

In relation now with the theta series, things are more tricky, in order to introduce its measure-theoretic version. Following [11], let us introduce the following notion:

DEFINITION 16.11. *The circular measure ε of a rooted bipartite graph X is given by*

$$d\varepsilon(q) = d\mu((q + q^{-1})^2)$$

where μ is the associated real measure.

In other words, the circular measure ε is by definition the pullback of the usual real measure μ via the following map, coming from the theory of the theta series in [49]:

$$\mathbb{R} \cup \mathbb{T} \rightarrow \mathbb{R}_+$$

$$q \rightarrow (q + q^{-1})^2$$

As we will see, all this best works in index $N \in [1, 4]$, with the circular measure ε being here the best-looking invariant, among all subfactor invariants. In index $N > 4$ things will turn to be quite complicated, but more on this later.

As a basic example for all this, assume that μ is a discrete measure, supported by n positive numbers $x_1 < \dots < x_n$, with corresponding densities p_1, \dots, p_n :

$$\mu = \sum_{i=1}^n p_i \delta_{x_i}$$

For each $i \in \{1, \dots, n\}$ the equation $(q + q^{-1})^2 = x_i$ has four solutions, that we can denote $q_i, q_i^{-1}, -q_i, -q_i^{-1}$. With this notation, we have:

$$\varepsilon = \frac{1}{4} \sum_{i=1}^n p_i \left(\delta_{q_i} + \delta_{q_i^{-1}} + \delta_{-q_i} + \delta_{-q_i^{-1}} \right)$$

In general, the basic properties of ε can be summarized as follows:

PROPOSITION 16.12. *The circular measure has the following properties:*

- (1) ε has equal density at $q, q^{-1}, -q, -q^{-1}$.
- (2) The odd moments of ε are 0.
- (3) The even moments of ε are half-integers.
- (4) When X has norm ≤ 2 , ε is supported by the unit circle.
- (5) When X is finite, ε is discrete.
- (6) If K is a solution of $L = (K + K^{-1})^2$, then $\varepsilon = \text{law}(K)$.

PROOF. These results can be deduced from definitions, the idea being that (1-5) are trivial, and that (6) follows from the formula of μ from Proposition 16.10 above. \square

In addition to the above, we have the following key formula, which gives the even moments of ε , and makes the connection with the Jones theta series:

THEOREM 16.13. *We have the Stieltjes transform type formula*

$$2 \int \frac{1}{1 - qu^2} d\varepsilon(u) = 1 + T(q)(1 - q)$$

where the T series of a rooted bipartite graph X is by definition given by

$$T(q) = \frac{\Theta(q) - q}{1 - q}$$

with Θ being the associated theta series.

PROOF. This follows by applying the change of variables $q \rightarrow (q + q^{-1})^2$ to the fact that f is the Stieltjes transform of μ . Indeed, we obtain in this way:

$$\begin{aligned} 2 \int \frac{1}{1 - qu^2} d\varepsilon(u) &= 1 + \frac{1 - q}{1 + q} f \left(\frac{q}{(1 + q)^2} \right) \\ &= 1 + \Theta(q) - q \\ &= 1 + T(q)(1 - q) \end{aligned}$$

Thus, we are led to the conclusion in the statement. \square

As a final theoretical result about all these invariants, which is this time something non-trivial, in the subfactor case, we have the following result, due to Jones [49]:

THEOREM 16.14. *In the case where X is the principal graph of an irreducible subfactor of index > 4 , the moments of ε are positive numbers.*

PROOF. This follows indeed from the result in [49] that the coefficients of Θ are positive numbers, as explained in Theorem 16.6, via the formula in Theorem 16.13. \square

Summarizing, we have a whole menagerie of subfactor, planar algebra and bipartite graph invariants, which come in several flavors, namely series and measures, and which can be linear or circular, and which all appear as versions of the Poincaré series.

Our claim now is that the circular measure ε is the “best” invariant. As a first justification for this claim, let us compute ε for the simplest possible graph in the index range $N \in [1, 4]$, namely the graph \tilde{A}_{2n} . We obtain here something nice, as follows:

THEOREM 16.15. *The circular measure of the basic index 4 graph, namely*

$$\tilde{A}_{2n} = \begin{array}{cccccccc} & \circ & - & \circ & - & \circ & \cdots & \circ & - & \circ & - & \circ \\ & | & & & & & & & & & & | \\ \bullet & - & \circ & - & \circ & - & \circ & - & \circ & - & \circ & \bullet \end{array}$$

is the uniform measure on the $2n$ -roots of unity.

PROOF. Let us identify the vertices of $X = \tilde{A}_{2n}$ with the group $\{w^k\}$ formed by the $2n$ -th roots of unity in the complex plane, where $w = e^{\pi i/n}$. The adjacency matrix of X acts then on the functions $f \in C(X)$ in the following way:

$$Mf(w^s) = f(w^{s-1}) + f(w^{s+1})$$

But this shows that we have $M = K + K^{-1}$, where K is given by:

$$Kf(w^s) = f(w^{s+1})$$

Thus we can use the last assertion in Proposition 16.12 above, and we get $\varepsilon = \text{law}(K)$, which is the uniform measure on the $2n$ -roots of unity. See [11] for details. \square

In order to discuss all this more systematically, and for all the ADE graphs, the idea will be that of looking at the combinatorics of the roots of unity. Let us introduce:

DEFINITION 16.16. *The series of the form*

$$\xi(n_1, \dots, n_s : m_1, \dots, m_t) = \frac{(1 - q^{n_1}) \dots (1 - q^{n_s})}{(1 - q^{m_1}) \dots (1 - q^{m_t})}$$

with $n_i, m_i \in \mathbb{N}$ are called *cyclotomic*.

It is technically convenient to allow as well $1 + q^n$ factors, to be designated by n^+ symbols in the above writing. For instance we have, by definition:

$$\xi(2^+ : 3) = \xi(4 : 2, 3)$$

Also, it is convenient in what follows to use the following notations:

$$\xi' = \frac{\xi}{1 - q} \quad , \quad \xi'' = \frac{\xi}{1 - q^2}$$

The Poincaré series of the ADE graphs are given by quite complicated formulae. However, the corresponding T series are all cyclotomic, as follows:

THEOREM 16.17. *The T series of the ADE graphs are as follows:*

- (1) For A_{n-1} we have $T = \xi(n-1 : n)$.
- (2) For D_{n+1} we have $T = \xi(n-1^+ : n^+)$.
- (3) For \tilde{A}_{2n} we have $T = \xi'(n^+ : n)$.
- (4) For \tilde{D}_{n+2} we have $T = \xi''(n+1^+ : n)$.
- (5) For E_6 we have $T = \xi(8 : 3, 6^+)$.
- (6) For E_7 we have $T = \xi(12 : 4, 9^+)$.
- (7) For E_8 we have $T = \xi(5^+, 9^+ : 15^+)$.
- (8) For \tilde{E}_6 we have $T = \xi(6^+ : 3, 4)$.
- (9) For \tilde{E}_7 we have $T = \xi(9^+ : 4, 6)$.
- (10) For \tilde{E}_8 we have $T = \xi(15^+ : 6, 10)$.

PROOF. These formulae were obtained in [11], by counting loops, then by making the change of variables $z^{-1/2} = q^{1/2} + q^{-1/2}$, and factorizing the resulting series. An alternative proof for these formulae can be obtained by using planar algebra methods. \square

Our purpose now will be that of converting the above technical results, regarding the T series, into some final results, regarding the corresponding circular measures ε . For this purpose, we will use the conversion formula in Theorem 16.13 above.

In order to formulate our results, we will need some more theory. First, we have:

DEFINITION 16.18. *A cyclotomic measure is a probability measure ε on the unit circle, having the following properties:*

- (1) ε is supported by the $2n$ -roots of unity, for some $n \in \mathbb{N}$.
- (2) ε has equal density at $q, q^{-1}, -q, -q^{-1}$.

It follows from Theorem 16.17 that the circular measures of the finite ADE graphs are supported by certain roots of unity, hence are cyclotomic. We will be back to this.

At the general level now, let us introduce as well the following notion:

DEFINITION 16.19. *The T series of a cyclotomic measure ε is given by:*

$$1 + T(q)(1 - q) = 2 \int \frac{1}{1 - qu^2} d\varepsilon(u)$$

Observe that this formula is nothing but the one in Theorem 16.13, written now in the other sense. In other words, if the cyclotomic measure ε happens to be the circular measure of a rooted bipartite graph, then the T series as defined above coincides with the T series as defined before. This is useful for explicit computations.

We are now ready to discuss the circular measures of the various ADE graphs. The idea is that these measures are all cyclotomic, of level ≤ 3 , and can be expressed in terms of the basic polynomial densities of degree ≤ 6 , namely:

$$\begin{aligned}\alpha &= \operatorname{Re}(1 - q^2) \\ \beta &= \operatorname{Re}(1 - q^4) \\ \gamma &= \operatorname{Re}(1 - q^6)\end{aligned}$$

To be more precise, we have the following result, with α, β, γ being as above, with d_n being the uniform measure on the $2n$ -th roots of unity, and with $d'_n = 2d_{2n} - d_n$ being the uniform measure on the odd $4n$ -roots of unity:

THEOREM 16.20. *The circular measures of the ADE graphs are given by:*

- (1) $A_{n-1} \rightarrow \alpha_n$.
- (2) $\tilde{A}_{2n} \rightarrow d_n$.
- (3) $D_{n+1} \rightarrow \alpha'_n$.
- (4) $\tilde{D}_{n+2} \rightarrow (d_n + d'_1)/2$.
- (5) $E_6 \rightarrow \alpha_{12} + (d_{12} - d_6 - d_4 + d_3)/2$.
- (6) $E_7 \rightarrow \beta'_9 + (d'_1 - d'_3)/2$.
- (7) $E_8 \rightarrow \alpha'_{15} + \gamma'_{15} - (d'_5 + d'_3)/2$.
- (8) $\tilde{E}_{n+3} \rightarrow (d_n + d_3 + d_2 - d_1)/2$.

PROOF. This follows from the T series formulae in Theorem 16.17, via some routine manipulations, based on the general conversion formulae given above. \square

It is possible to further build along the above lines, with a combinatorial refinement of the formulae in Theorem 16.20, making appear a certain connection with the Deligne work on the exceptional series of Lie groups, which is not understood yet. See [5].

16c. Measure blowup

All the above, which was quite nice, was about index $N \in [1, 4]$, where the Jones annular theory result from [49] does not apply. In higher index now, $N \in (4, \infty)$, where the Jones result does apply, the precise correct “blowup” manipulation on the spectral measure is not known yet. The known results here are as follows:

- (1) On one hand, there is as a computation for some basic Hadamard subfactors, with nice blowup, on a certain noncommutative manifold [9].
- (2) On the other hand, there are many computations by Evans-Pugh, with quite technical blowup results, on some suitable real algebraic manifolds [33].

We will discuss in what follows (1), and to be more precise the computation of the spectral measure, and then the blowup problem, for the subfactors coming from the deformed Fourier matrices. Let us start with the following definition:

DEFINITION 16.21. Given two finite abelian groups G, H , we consider the corresponding deformed Fourier matrix, given by the formula

$$(F_G \otimes_Q F_H)_{ia,jb} = Q_{ib}(F_G)_{ij}(F_H)_{ab}$$

and we factorize the associated representation π_Q of the algebra $C(S_{G \times H}^+)$,

$$\begin{array}{ccc} C(S_{G \times H}^+) & \xrightarrow{\pi_Q} & M_{G \times H}(\mathbb{C}) \\ & \searrow & \nearrow \pi \\ & C(G_Q) & \end{array}$$

with $C(G_Q)$ being the Hopf image of this representation π_Q .

Explicitly computing the above quantum permutation group $G_Q \subset S_{G \times H}^+$, as function of the parameter matrix $Q \in M_{G \times H}(\mathbb{T})$, will be our main purpose, in what follows. In order to do so, we first have the following elementary result:

PROPOSITION 16.22. We have a factorization as follows,

$$\begin{array}{ccc} C(S_{G \times H}^+) & \xrightarrow{\pi_Q} & M_{G \times H}(\mathbb{C}) \\ & \searrow & \nearrow \pi \\ & C(H \wr_* G) & \end{array}$$

given on the standard generators by the formulae

$$U_{ab}^{(i)} = \sum_j W_{ia,jb} \quad , \quad V_{ij} = \sum_a W_{ia,jb}$$

independently of b , where W is the magic matrix producing π_Q .

PROOF. With $K = F_G, L = F_H$ and $M = |G|, N = |H|$, the formula of the magic matrix $W \in M_{G \times H}(M_{G \times H}(\mathbb{C}))$ associated to $H = K \otimes_Q L$ is as follows:

$$\begin{aligned} (W_{ia,jb})_{kc,ld} &= \frac{1}{MN} \cdot \frac{Q_{ic}Q_{jd}}{Q_{id}Q_{jc}} \cdot \frac{K_{ik}K_{jl}}{K_{il}K_{jk}} \cdot \frac{L_{ac}L_{bd}}{L_{ad}L_{bc}} \\ &= \frac{1}{MN} \cdot \frac{Q_{ic}Q_{jd}}{Q_{id}Q_{jc}} \cdot K_{i-j,k-l} L_{a-b,c-d} \end{aligned}$$

Our claim now is that the representation π_Q constructed in Definition 16.21 can be factorized in three steps, up to the factorization in the statement, as follows:

$$\begin{array}{ccc}
 C(S_{G \times H}^+) & \xrightarrow{\pi_Q} & M_{G \times H}(\mathbb{C}) \\
 \downarrow & \nearrow & \uparrow \\
 C(S_H^+ \wr_* S_G^+) & \longrightarrow & C(S_H^+ \wr_* G) \longrightarrow C(H \wr_* G)
 \end{array}$$

Indeed, the construction of the map on the left is standard. Regarding the second factorization, this comes from the fact that since the elements V_{ij} depend on $i - j$, they satisfy the defining relations for the quotient algebra $C(S_G^+) \rightarrow C(G)$. Finally, regarding the third factorization, observe that $W_{ia,jb}$ depends only on i, j and on $a - b$. By summing over j we obtain that the elements $U_{ab}^{(i)}$ depend only on $a - b$, and we are done. \square

We have now all needed ingredients for refining Proposition 16.22, as follows:

PROPOSITION 16.23. *We have a factorization as follows,*

$$\begin{array}{ccc}
 C(S_{G \times H}^+) & \xrightarrow{\pi_Q} & M_{G \times H}(\mathbb{C}) \\
 \searrow & & \nearrow \rho \\
 & C^*(\Gamma_{G,H}) \rtimes C(G) &
 \end{array}$$

where the group on the bottom is given by:

$$\Gamma_{G,H} = H^{*G} / \left\langle [c_1^{(i_1)} \dots c_s^{(i_s)}, d_1^{(j_1)} \dots d_s^{(j_s)}] = 1 \mid \sum_r c_r = \sum_r d_r = 0 \right\rangle$$

PROOF. Assume that we have a representation, as follows:

$$\pi : C^*(\Gamma) \rtimes C(G) \rightarrow M_L(\mathbb{C})$$

Let Λ be a G -stable normal subgroup of Γ , so that G acts on Γ/Λ , and we can form the product $C^*(\Gamma/\Lambda) \rtimes C(G)$, and assume that π is trivial on Λ . Then π factorizes as:

$$\begin{array}{ccc}
 C^*(\Gamma) \rtimes C(G) & \xrightarrow{\pi} & M_L(\mathbb{C}) \\
 \searrow & & \nearrow \rho \\
 & C^*(\Gamma/\Lambda) \rtimes C(G) &
 \end{array}$$

With $\Gamma = H^{*G}$, this gives the result. \square

We have now all the needed ingredients for proving a main result, as follows:

THEOREM 16.24. *When Q is generic, the minimal factorization for π_Q is*

$$\begin{array}{ccc}
 C(S_{G \times H}^+) & \xrightarrow{\pi_Q} & M_{G \times H}(\mathbb{C}) \\
 & \searrow & \nearrow \pi \\
 & & C^*(\Gamma_{G,H}) \rtimes C(G)
 \end{array}$$

where on the bottom

$$\Gamma_{G,H} \simeq \mathbb{Z}^{(|G|-1)(|H|-1)} \rtimes H$$

is the discrete group constructed above.

PROOF. Consider the factorization in Proposition 16.23, which is as follows, where L denotes the Hopf image of π_Q :

$$\theta : C^*(\Gamma_{G,H}) \rtimes C(G) \rightarrow L$$

To be more precise, this morphism produces the following commutative diagram:

$$\begin{array}{ccc}
 C(S_{G \times H}^+) & \xrightarrow{\pi_Q} & M_{G \times H}(\mathbb{C}) \\
 & \searrow & \nearrow \pi \\
 & & L \\
 & \searrow & \uparrow \theta \\
 & & C^*(\Gamma_{G,H}) \rtimes C(G)
 \end{array}$$

The first observation is that the injectivity assumption on $C(G)$ holds by construction, and that for $f \in C(G)$, the matrix $\pi(f)$ is “block scalar”. Now for $r \in \Gamma_{G,H}$ with $\theta(r \otimes 1) = \theta(1 \otimes f)$ for some $f \in C(G)$, we see, using the commutative diagram, that $\pi(r \otimes 1)$ is block scalar. Thus, modulo some standard algebra, we are done. \square

Summarizing, we have computed the quantum permutation groups associated to the Diță deformations of the tensor products of Fourier matrices, in the case where the deformation matrix Q is generic. For some further computations, in the case where the deformation matrix Q is no longer generic, we refer to the literature.

Let us compute now the Kesten measure $\mu = \text{law}(\chi)$, in the case where the deformation matrix is generic, as before. Our results here will be a combinatorial moment formula, a geometric interpretation of it, and an asymptotic result. We first have:

THEOREM 16.25. *We have the moment formula*

$$\int \chi^p = \frac{1}{|G| \cdot |H|} \# \left\{ \begin{array}{l} i_1, \dots, i_p \in G \\ d_1, \dots, d_p \in H \end{array} \middle| \begin{array}{l} [(i_1, d_1), (i_2, d_2), \dots, (i_p, d_p)] \\ [(i_1, d_p), (i_2, d_1), \dots, (i_p, d_{p-1})] \end{array} \right\}$$

where the sets between square brackets are by definition sets with repetition.

PROOF. According to the various formulae above, the factorization found in Theorem 16.24 is, at the level of standard generators, as follows:

$$\begin{aligned} C(S_{G \times H}^+) &\rightarrow C^*(\Gamma_{G,H}) \otimes C(G) &\rightarrow M_{G \times H}(\mathbb{C}) \\ u_{ia,jb} &\rightarrow \frac{1}{|H|} \sum_c F_{b-a,c} c^{(i)} \otimes v_{ij} &\rightarrow W_{ia,jb} \end{aligned}$$

Thus, the main character of the quantum permutation group that we found in Theorem 16.24 is given by the following formula:

$$\begin{aligned} \chi &= \frac{1}{|H|} \sum_{iac} c^{(i)} \otimes v_{ii} \\ &= \sum_{ic} c^{(i)} \otimes v_{ii} \\ &= \left(\sum_{ic} c^{(i)} \right) \otimes \delta_1 \end{aligned}$$

Now since the Haar functional of $C^*(\Gamma) \rtimes C(H)$ is the tensor product of the Haar functionals of $C^*(\Gamma), C(H)$, this gives the following formula, valid for any $p \geq 1$:

$$\int \chi^p = \frac{1}{|G|} \int_{\widehat{\Gamma}_{G,H}} \left(\sum_{ic} c^{(i)} \right)^p$$

Consider the elements $S_i = \sum_c c^{(i)}$. With standard notations, we have:

$$S_i = \sum_c (b_{i0} - b_{ic}, c)$$

Now observe that these elements multiply as follows:

$$S_{i_1} \dots S_{i_p} = \sum_{c_1 \dots c_p} \left(\begin{array}{l} b_{i_1 0} - b_{i_1 c_1} + b_{i_2 c_1} - b_{i_2, c_1 + c_2} \\ + b_{i_3, c_1 + c_2} - b_{i_3, c_1 + c_2 + c_3} + \dots \\ \dots + b_{i_p, c_1 + \dots + c_{p-1}} - b_{i_p, c_1 + \dots + c_p} \end{array} , c_1 + \dots + c_p \right)$$

In terms of the new indices $d_r = c_1 + \dots + c_r$, this formula becomes:

$$S_{i_1} \dots S_{i_p} = \sum_{d_1 \dots d_p} \left(\begin{array}{l} b_{i_1 0} - b_{i_1 d_1} + b_{i_2 d_1} - b_{i_2 d_2} \\ + b_{i_3 d_2} - b_{i_3 d_3} + \dots \\ \dots + b_{i_p d_{p-1}} - b_{i_p d_p} \end{array} , d_p \right)$$

Now by integrating, we must have $d_p = 0$ on one hand, and on the other hand:

$$[(i_1, 0), (i_2, d_1), \dots, (i_p, d_{p-1})] = [(i_1, d_1), (i_2, d_2), \dots, (i_p, d_p)]$$

Equivalently, we must have $d_p = 0$ on one hand, and on the other hand:

$$[(i_1, d_p), (i_2, d_1), \dots, (i_p, d_{p-1})] = [(i_1, d_1), (i_2, d_2), \dots, (i_p, d_p)]$$

Thus, by translation invariance with respect to d_p , we obtain:

$$\int_{\widehat{\Gamma}_{G,H}} S_{i_1} \dots S_{i_p} = \frac{1}{|H|} \# \left\{ d_1, \dots, d_p \in H \mid \begin{matrix} [(i_1, d_1), (i_2, d_2), \dots, (i_p, d_p)] \\ = [(i_1, d_p), (i_2, d_1), \dots, (i_p, d_{p-1})] \end{matrix} \right\}$$

It follows that we have the following moment formula:

$$\int_{\widehat{\Gamma}_{G,H}} \left(\sum_i S_i \right)^p = \frac{1}{|H|} \# \left\{ \begin{matrix} i_1, \dots, i_p \in G \mid [(i_1, d_1), (i_2, d_2), \dots, (i_p, d_p)] \\ d_1, \dots, d_p \in H \mid = [(i_1, d_p), (i_2, d_1), \dots, (i_p, d_{p-1})] \end{matrix} \right\}$$

Now by dividing by $|G|$, we obtain the formula in the statement. □

The formula in Theorem 16.25 can be further interpreted as follows:

THEOREM 16.26. *With $M = |G|, N = |H|$ we have the formula*

$$law(\chi) = \left(1 - \frac{1}{N}\right) \delta_0 + \frac{1}{N} law(A)$$

where the matrix

$$A \in C(\mathbb{T}^{MN}, M_M(\mathbb{C}))$$

is given by $A(q) =$ Gram matrix of the rows of q .

PROOF. According to Theorem 16.25, we have the following formula:

$$\begin{aligned} \int \chi^p &= \frac{1}{MN} \sum_{i_1 \dots i_p} \sum_{d_1 \dots d_p} \delta_{[i_1 d_1, \dots, i_p d_p], [i_1 d_p, \dots, i_p d_{p-1}]} \\ &= \frac{1}{MN} \int_{\mathbb{T}^{MN}} \sum_{i_1 \dots i_p} \sum_{d_1 \dots d_p} \frac{q_{i_1 d_1} \dots q_{i_p d_p}}{q_{i_1 d_p} \dots q_{i_p d_{p-1}}} dq \\ &= \frac{1}{MN} \int_{\mathbb{T}^{MN}} \sum_{i_1 \dots i_p} \left(\sum_{d_1} \frac{q_{i_1 d_1}}{q_{i_2 d_1}} \right) \left(\sum_{d_2} \frac{q_{i_2 d_2}}{q_{i_3 d_2}} \right) \dots \left(\sum_{d_p} \frac{q_{i_p d_p}}{q_{i_1 d_p}} \right) dq \end{aligned}$$

Consider now the Gram matrix in the statement, namely:

$$A(q)_{ij} = \langle R_i, R_j \rangle$$

Here R_1, \dots, R_M are the rows of the following matrix:

$$q \in \mathbb{T}^{MN} \simeq M_{M \times N}(\mathbb{T})$$

We have then the following computation:

$$\begin{aligned} \int \chi^p &= \frac{1}{MN} \int_{\mathbb{T}^{MN}} \langle R_{i_1}, R_{i_2} \rangle \langle R_{i_2}, R_{i_3} \rangle \dots \langle R_{i_p}, R_{i_1} \rangle \\ &= \frac{1}{MN} \int_{\mathbb{T}^{MN}} A(q)_{i_1 i_2} A(q)_{i_2 i_3} \dots A(q)_{i_p i_1} \\ &= \frac{1}{MN} \int_{\mathbb{T}^{MN}} \text{Tr}(A(q)^p) dq \\ &= \frac{1}{N} \int_{\mathbb{T}^{MN}} \text{tr}(A(q)^p) dq \end{aligned}$$

But this gives the formula in the statement, and we are done. □

In general, the moments of the Gram matrix A are given by a quite complicated formula, and we cannot expect to have a refinement of Theorem 16.26, with A replaced by a plain, non-matricial random variable, say over a compact abelian group.

However, this kind of simplification does appear at $M = 2$, and since this phenomenon is quite interesting, we will explain this now. We first have:

PROPOSITION 16.27. *For $F_2 \otimes_Q F_H$, with $Q \in M_{2 \times N}(\mathbb{T})$ generic, we have*

$$N \int \left(\frac{\chi}{N}\right)^p = \int_{\mathbb{T}^N} \sum_{k \geq 0} \binom{p}{2k} \left| \frac{a_1 + \dots + a_N}{N} \right|^{2k} da$$

where the integral on the right is with respect to the uniform measure on \mathbb{T}^N .

PROOF. In order to prove the result, consider the following quantity, which appeared in the proof of Theorem 16.26 above:

$$\Phi(q) = \sum_{i_1 \dots i_p} \sum_{d_1 \dots d_p} \frac{q_{i_1 d_1} \dots q_{i_p d_p}}{q_{i_1 d_p} \dots q_{i_p d_{p-1}}}$$

We can “half-dephase” the matrix $q \in M_{2 \times N}(\mathbb{T})$ if we want to, as follows:

$$q = \begin{pmatrix} 1 & \dots & 1 \\ a_1 & \dots & a_N \end{pmatrix}$$

Let us compute now the above quantity $\Phi(q)$, in terms of the numbers a_1, \dots, a_N . Our claim is that we have the following formula:

$$\Phi(q) = 2 \sum_{k \geq 0} N^{p-2k} \binom{p}{2k} \left| \sum_i a_i \right|^{2k}$$

Indeed, the idea is that:

- (1) The $2N^k$ contribution will come from $i = (1 \dots 1)$ and $i = (2 \dots 2)$.

(2) Then we will have a $p(p-1)N^{k-2}|\sum_i a_i|^2$ contribution coming from indices of type $i = (2 \dots 21 \dots 1)$, up to cyclic permutations.

(3) Then we will have a $2\binom{p}{4}N^{p-4}|\sum_i a_i|^4$ contribution coming from indices of type $i = (2 \dots 21 \dots 12 \dots 21 \dots 1)$.

(4) And so on.

In practice now, in order to prove our claim, in order to find the $N^{p-2k}|\sum_i a_i|^{2k}$ contribution, we have to count the circular configurations consisting of p numbers 1, 2, such that the 1 values are arranged into k non-empty intervals, and the 2 values are arranged into k non-empty intervals as well. Now by looking at the endpoints of these $2k$ intervals, we have $2\binom{p}{2k}$ choices, and this gives the above formula.

Now by integrating, this gives the formula in the statement. □

Observe now that the integrals in Proposition 16.27 can be computed as follows:

$$\begin{aligned} \int_{\mathbb{T}^N} |a_1 + \dots + a_N|^{2k} da &= \int_{\mathbb{T}^N} \sum_{i_1 \dots i_k} \sum_{j_1 \dots j_k} \frac{a_{i_1} \dots a_{i_k}}{a_{j_1} \dots a_{j_k}} da \\ &= \# \left\{ i_1 \dots i_k, j_1 \dots j_k \mid [i_1, \dots, i_k] = [j_1, \dots, j_k] \right\} \\ &= \sum_{k=\sum r_i} \binom{k}{r_1, \dots, r_N}^2 \end{aligned}$$

We obtain in this way the following “blowup” result, for our measure:

PROPOSITION 16.28. *For $F_2 \otimes_Q F_H$, with $Q \in M_{2 \times N}(\mathbb{T})$ generic, we have*

$$\mu = \left(1 - \frac{1}{N}\right) \delta_0 + \frac{1}{2N} (\Psi_*^+ \varepsilon + \Psi_*^- \varepsilon)$$

where ε is the uniform measure on \mathbb{T}^N , and where the blowup function is:

$$\Psi^\pm(a) = N \pm \left| \sum_i a_i \right|$$

PROOF. We use the formula found in Proposition 16.27 above, along with the following standard identity, coming from the Taylor formula:

$$\sum_{k \geq 0} \binom{p}{2k} x^{2k} = \frac{(1+x)^p + (1-x)^p}{2}$$

By using this identity, Proposition 16.27 reformulates as follows:

$$N \int \left(\frac{\chi}{N}\right)^p = \frac{1}{2} \int_{\mathbb{T}^N} \left(1 + \left|\frac{\sum_i a_i}{N}\right|\right)^p + \left(1 - \left|\frac{\sum_i a_i}{N}\right|\right)^p da$$

Now by multiplying by N^{p-1} , we obtain the following formula:

$$\int \chi^k = \frac{1}{2N} \int_{\mathbb{T}^N} \left(N + \left| \sum_i a_i \right| \right)^p + \left(N - \left| \sum_i a_i \right| \right)^p da$$

But this gives the formula in the statement, and we are done. □

We can further improve the above result, by reducing the maps Ψ^\pm appearing there to a single one, and we are led to the following statement:

THEOREM 16.29. *For $F_2 \otimes_Q F_H$, with $Q \in M_{2 \times N}(\mathbb{T})$ generic, we have*

$$\mu = \left(1 - \frac{1}{N} \right) \delta_0 + \frac{1}{N} \Phi_* \varepsilon$$

where ε is the uniform measure on $\mathbb{Z}_2 \times \mathbb{T}^N$, and where the blowup map is:

$$\Phi(e, a) = N + e \left| \sum_i a_i \right|$$

PROOF. This is clear indeed from Proposition 16.28 above. □

As already mentioned, the above results at $M = 2$ are something quite special. In the general case, $M \in \mathbb{N}$, it is not clear how to construct a nice blowup of the measure.

Asymptotically, things are however quite simple. Let us go back indeed to the general case, where $M, N \in \mathbb{N}$ are both arbitrary. The problem that we would like to solve now is that of finding the good regime, of the following type, where the measure in Theorem 16.25 converges, after some suitable manipulations:

$$M = f(K) \quad , \quad N = g(K) \quad , \quad K \rightarrow \infty$$

In order to do so, we have to do some combinatorics. Let $NC(p)$ be the set of noncrossing partitions of $\{1, \dots, p\}$, and for $\pi \in P(p)$ we denote by $|\pi| \in \{1, \dots, p\}$ the number of blocks. With these conventions, we have the following result:

PROPOSITION 16.30. *With $M = \alpha K, N = \beta K, K \rightarrow \infty$ we have:*

$$\frac{c_p}{K^{p-1}} \simeq \sum_{r=1}^p \# \left\{ \pi \in NC(p) \mid |\pi| = r \right\} \alpha^{r-1} \beta^{p-r}$$

In particular, with $\alpha = \beta$ we have:

$$c_p \simeq \frac{1}{p+1} \binom{2p}{p} (\alpha K)^{p-1}$$

PROOF. We use the combinatorial formula in Theorem 16.25 above. Our claim is that, with $\pi = \ker(i_1, \dots, i_p)$, the corresponding contribution to c_p is:

$$C_\pi \simeq \begin{cases} \alpha^{|\pi|-1} \beta^{p-|\pi|} K^{p-1} & \text{if } \pi \in NC(p) \\ O(K^{p-2}) & \text{if } \pi \notin NC(p) \end{cases}$$

As a first observation, the number of choices for a multi-index $(i_1, \dots, i_p) \in X^p$ satisfying $\ker i = \pi$ is:

$$M(M-1) \dots (M-|\pi|+1) \simeq M^{|\pi|}$$

Thus, we have the following estimate:

$$C_\pi \simeq M^{|\pi|-1} N^{-1} \# \left\{ d_1, \dots, d_p \in Y \mid [d_\alpha | \alpha \in b] = [d_{\alpha-1} | \alpha \in b], \forall b \in \pi \right\}$$

Consider now the following partition:

$$\sigma = \ker d$$

The contribution of σ to the above quantity C_π is then given by:

$$\Delta(\pi, \sigma) N(N-1) \dots (N-|\sigma|+1) \simeq \Delta(\pi, \sigma) N^{|\sigma|}$$

Here the quantities on the right are as follows:

$$\Delta(\pi, \sigma) = \begin{cases} 1 & \text{if } |b \cap c| = |(b-1) \cap c|, \forall b \in \pi, \forall c \in \sigma \\ 0 & \text{otherwise} \end{cases}$$

We use now the standard fact that for $\pi, \sigma \in P(p)$ satisfying $\Delta(\pi, \sigma) = 1$ we have:

$$|\pi| + |\sigma| \leq p + 1$$

In addition, the equality case is well-known to happen when $\pi, \sigma \in NC(p)$ are inverse to each other, via Kreweras complementation. This shows that for $\pi \notin NC(p)$ we have:

$$C_\pi = O(K^{p-2})$$

Also, this shows that for $\pi \in NC(p)$ we have:

$$\begin{aligned} C_\pi &\simeq M^{|\pi|-1} N^{-1} N^{p-|\pi|-1} \\ &= \alpha^{|\pi|-1} \beta^{p-|\pi|} K^{p-1} \end{aligned}$$

Thus, we have obtained the result. □

We denote by D the dilation operation, given by:

$$D_r(\text{law}(X)) = \text{law}(rX)$$

With this convention, we have the following result:

THEOREM 16.31. *With $M = \alpha K, N = \beta K, K \rightarrow \infty$ we have:*

$$\mu = \left(1 - \frac{1}{\alpha\beta K^2}\right) \delta_0 + \frac{1}{\alpha\beta K^2} D_{\frac{1}{\beta K}}(\pi_{\alpha/\beta})$$

In particular with $\alpha = \beta$ we have:

$$\mu = \left(1 - \frac{1}{\alpha^2 K^2}\right) \delta_0 + \frac{1}{\alpha^2 K^2} D_{\frac{1}{\alpha K}}(\pi_1)$$

PROOF. At $\alpha = \beta$, this follows from Proposition 16.30. In general now, we have:

$$\begin{aligned} \frac{c_p}{K^{p-1}} &\simeq \sum_{\pi \in NC(p)} \alpha^{|\pi|-1} \beta^{p-|\pi|} \\ &= \frac{\beta^p}{\alpha} \sum_{\pi \in NC(p)} \left(\frac{\alpha}{\beta}\right)^{|\pi|} \\ &= \frac{\beta^p}{\alpha} \int x^p d\pi_{\alpha/\beta}(x) \end{aligned}$$

When $\alpha \geq \beta$, where $d\pi_{\alpha/\beta}(x) = \varphi_{\alpha/\beta}(x)dx$ is continuous, we obtain:

$$\begin{aligned} c_p &= \frac{1}{\alpha K} \int (\beta K x)^p \varphi_{\alpha/\beta}(x) dx \\ &= \frac{1}{\alpha\beta K^2} \int x^p \varphi_{\alpha/\beta}\left(\frac{x}{\beta K}\right) dx \end{aligned}$$

But this gives the formula in the statement. When $\alpha \leq \beta$ the computation is similar, with a Dirac mass as 0 disappearing and reappearing, and gives the same result. \square

16d. Big index

In big index now, the philosophy is that the index of subfactors $N \in [1, \infty)$ should be regarded as being the famous N variable from physics, which must be big:

$$N \rightarrow \infty$$

More precisely, the idea is that the constructions involving groups, group duals, or more generally compact quantum groups, producing subfactors of integer index, $N \in \mathbb{N}$, can be used with “uniform objects” as input, and so produce an asymptotic theory.

The problem however is how to axiomatize the uniformity notion which is needed, in order to have some control on the resulting planar algebra $P = (P_k)$. The answer here comes from the notion of easiness, that we already met in chapter 8 above, and its various technical extensions, which are in fact not currently unified, or even fully axiomatized.

The main technical questions here are the classification of the easy quantum groups on one hand, and the axiomatization of the super-quizzzy quantum groups on the other

hand. We also have the question of better understanding the relation between easiness, subfactors, planar algebras, noncommutative geometry and free probability, and we refer here to [12], [13], [14], [16], [23], [27], [28], [30], [54], [63], [74], [78], [79], [95], [97].

Summarizing, we have many interesting questions, both in small and big index. As a common ground here, both these happen inside R , although this is conjectural in big index. Thus, as a good problem to finish with, from [43], we have the question of classifying the finite index subfactors of the Murray-von Neumann hyperfinite II_1 factor R .

As already mentioned on several occasions, this longstanding question is in need of some new, brave functional analysis input, in relation with the notion of hyperfiniteness, which is probably of quite difficult type, beyond what the current experts can do.

16e. Exercises

Congratulations for having read this book, and no exercises here.

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