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# FACTORIZATION FOR NON-SYMMETRIC OPERATORS AND EXPONENTIAL $H$-THEOREM 

M.P. GUALDANI, S. MISCHLER, C. MOUHOT


#### Abstract

We present a factorization method for estimating resolvents of nonsymmetric operators in Banach or Hilbert spaces in terms of estimates in another (typically smaller) "reference" space. This applies to a class of operators writing as a "regularizing" part (in a broad sense) plus a dissipative part. Then in the Hilbert case we combine this factorization approach with an abstract Plancherel identity on the resolvent into a method for enlarging the functional space of decay estimates on semigroups. In the Banach case, we prove the same result however with some loss on the norm. We then apply these functional analysis approach to several PDEs: the Fokker-Planck and kinetic Fokker-Planck equations, the linear scattering Boltzmann equation in the torus, and, most importantly the linearized Boltzmann equation in the torus (at the price of extra specific work in the latter case). In addition to the abstract method in itself, the main outcome of the paper is indeed the first proof of exponential decay towards global equilibrium (e.g. in terms of the relative entropy) for the full Boltzmann equation for hard spheres, conditionnally to some smoothness and (polynomial) moment estimates. This improves on the result in [12] where the rate was "almost exponential", that is polynomial with exponent as high as wanted, and solves a long-standing conjecture about the rate of decay in the $H$-theorem for the nonlinear Boltzmann equation, see for instance [10, 27.


Mathematics Subject Classification (2000): 35P15 Estimation of eigenvalues, upper and lower bounds, 35Q84 Fokker-Planck equations, 76P05 Rarefied gas flows, Boltzmann equation [See also 82B40, 82C40, 82D05].

Keywords: spectral gap; semigroup; Plancherel theorem; coercivity; hypocoercivity; Fokker-Planck equation; Boltzmann equation; non cutoff; fractional derivative; exponential rate; thermalization; entropy.

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## 1. Introduction

This paper deals with the study of decay properties of linear semigroups and their link with spectral properties. Let us give a brief sketch of the question we address. Consider two Banach spaces $E \subset \mathcal{E}$, and two unbounded closed linear operator $L$ and $\mathcal{L}$ resp. on $E$ and $\mathcal{E}$ with spectrum $\Sigma(L), \Sigma(\mathcal{L}) \subsetneq \mathbb{C}$. They generate two $C_{0^{-}}$ semigroups $S(t)$ and $\mathcal{S}(t)$ resp. in $E$ and $\mathcal{E}$. Also assume that $\mathcal{L}_{\mid E}=L$, and $E$ is dense in $\mathcal{E}$.

The theoretical question we address in this work is the following: Can one deduce (quantitative) informations on $\Sigma(\mathcal{L})$ and $\mathcal{S}(t)$ in terms of informations on $\Sigma(L)$ and $S(t)$ ?

More precisely, we provide here an answer where (i) the spectral gap property of $L$ in $E$ can be shown to hold for $\mathcal{L}$ in the space $\mathcal{E}$ and consequently (ii) explicit estimates on the rate of decay of the semigroup $\mathcal{S}(t)$ can be computed from the ones on $S(t)$. Our results hold for a class of operators $\mathcal{L}$ which split as $\mathcal{L}=\mathcal{A}+\mathcal{B}$, where $\mathcal{A}$ is maps $\mathcal{E}$ into $E$ and $\mathcal{B}$ 's spectrum is well localized. We then show that linear Boltzmann-like and Fokker-Planck equations fall in that class and therefore their spectral-gap property can be extended from the linearization space (the weighted $L^{2}$ space, where the strong decay is prescribed by the equilibrium) to larger Hilbert or Banach spaces (for example with polynomial decay). It is worth mentioning that we provide results in the non spatially homogeneous situation, and that the method is optimal in terms of the rate of decay (no loss on the way from $E$ to $\mathcal{E}$, as would be the case in, say, some interpolation approach).

Enlarging the functional space for the spectral properties of linear operators is useful in nonlinear PDE analysis when the stability theory for the associated linearized equation is not compatible (that it, is established in a space too small) with the functional space in which the nonlinear PDE is well posed. The applications concern existence and uniqueness of stationary solution (thanks to perturbative or implicit function argument, see for instance [19, 2]) as well as the proof of long time convergence to the equilibrium (see for instance [22, 19]). In this paper we construct a general abstract approach whereas previous papers were dealing with particular cases. Within this general framework we give new results for the space homogeneous Fokker-Planck and linear Boltzmann equations, and more importantly we are able to tackle the corresponding space inhomogeneous (or kinetic) equations as well as the full (we mean again space inhomogeneous) Boltzmann equation. We also refer to the note [23] where some results of this paper were announced.

The main outcome of this paper is thus a proof of the exponential convergence to equilibrium for the full Boltzmann equation in the torus. This question goes back to the famous $H$-theorem of Boltzmann, and a functional inequality version of it (linear control of the relative entropy in terms of the entropy production functional) was conjectured by Cercignani [10] in the spatially homogeneous case (see also [11]). This conjecture has motivated important works from the early 1990's on: see [9, 8, 29, 6, 30] (in the spatially homogeneous case). While it has been shown to be false in general [6], it has given a formidable impulse to the works on the Boltzmann equation in the last two decades. It has been shown to be almost true in
many cases in the spatially homogeneous cases in [30] (in the sense that polynomial inequalities relating the relative entropy and the entropy production hold for a power as close to 1 as wanted), and was an important inspiration for the work [12] in the inhomogeneous case. However, due to the fact Cercignani's conjecture is false even in the homogeneous case [6], these important progresses in the far from equilibrium regime were unable to prove the natural conjecture about the exponential decay of the relative entropy in order to prove $H$-theorem with the correct time scale. This has motivated the work [22] which answers this question, but only in the spatially homogeneous case.

Here we finally answer this question for the full Boltzmann equation for hard spheres (or cutoff hard potentials), in the same setting as in [12], that is under some a priori regularity assumptions (Sobolev norms and polynomial moments bounds). This is based on the idea of connecting the nonlinear theory in [12] with the linearized theory developped in [24] and it makes crucial use of our new method for enlarging the space of decay estimates on semigroups. This hence answers the main question raised by Villani in [27, Subsection 1.8, page 62].

Let us mention that there is a huge gap between the spatially homogeneous case (where the equation is coercive and the linearized semigroup is self-adjoint or sectorial depending on the functional space) and the spatially inhomogeneous case (where the equation is hypocoercive and its linearized is neither sectorial, nor even hypoelliptic).

The plan of the paper is the following. In section 2 we state and prove the abstract factorization theorem for unbounded closed operators. Section 3 is devoted to the links between estimates on the resolvent and decay estimates on the semigroup and only in section 4 we can assemble the preceding elements into new estimates showing enlargement of the functional space of the decay of semigroups. In sections 5 and 6 we apply the abstract method to the linear Boltzmann and Fokker-Planck equations (either spatially homogeneous or in the torus). Finally in section 7 we consider the nonlinear Boltzmann equation. Using the abstract factorization approach for the linearized Boltzmann equation in the torus plus some additional specific work, we prove that one can "connect" the nonlinear theoy and the linearized theory in order to prove the exponential decay of the relative entropy.

If the reader is searching for a quick overview of our method, a possibility is to go directly to the statement of Corollary 4.2 which is the "concrete" version of the abstract method for most applications, and then to the section 6 for an easy and very short application of the method to the linear Boltzmann equation.

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## 2. The factorization theorem

Let us start with some notations. For some given Banach spaces $X$ and $Y$ we denote by $\mathscr{B}(Y \rightarrow X)$ the space of bounded linear operators from $Y$ to $X$, we denote by $\mathscr{B}(X)=\mathscr{B}(X \rightarrow X)$ in the case when $X=Y$, and we denote by $\mathscr{C}(X)$ the space of closed unbounded linear operators on $X$ with dense domain.

For $\Lambda \in \mathscr{C}(X)$ we denote by $D(\Lambda)$ its domain, by $N(\Lambda)$ its null space, by $M(\Lambda)=$ $\lim _{k \rightarrow \infty} N\left(\Lambda^{k}\right)$ its algebraic null space, by $R(\Lambda)$ its range and by $\Sigma(\Lambda)$ its spectrum, so that for any $\xi \in \mathbb{C} \backslash \Sigma(\Lambda)$ the operator $\Lambda-\xi$ is invertible and the resolvent operator $(\Lambda-\xi)^{-1}$ is well defined, belongs to $\mathcal{B}(X)$ and has range equal to $D(\Lambda)$.

For a given real number $a \in \mathbb{R}$, we define the half complex plane $\Delta_{a}:=\{z \in$ $\mathbb{C}, \Re e z>a\}$. We recall that for any isolated eigenvalue $\eta \in \Sigma(\Lambda)$ in the sense that $\Sigma(\Lambda) \cap\{\zeta \in \mathbb{C},|\zeta-\eta| \leq r\}=\{\eta\}$ for some $r>0$, we define $\Pi_{\Lambda, \eta}$ the projection on closure of the algebraic eigenspace associated to $\eta$ by (see [18, III-(6.19)])

$$
\begin{equation*}
\Pi_{\Lambda, \eta}:=\frac{i}{2 \pi} \int_{|\zeta-\eta|=r}(\Lambda-\zeta)^{-1} d \zeta \tag{2.1}
\end{equation*}
$$

Remind that $E_{\eta}:=R\left(\Pi_{\Lambda, \eta}\right)$ is finite dimensional iff there exists $\alpha_{0} \in \mathbb{N}^{*}$ such that $N\left((\Lambda-\eta)^{\alpha}\right)=N\left((\Lambda-\eta)^{\alpha_{0}}\right)$ for any $\alpha \geq \alpha_{0}$, so that $M(\Lambda-\eta)=N\left((\Lambda-\eta)^{\alpha_{0}}\right)$. In that case, we say that $\eta$ is a discrete eigenvalue.

In this section we consider $E \subset \mathcal{E}$ two Banach spaces such that $E$ is dense in $\mathcal{E}$, and two operators $L \in \mathscr{C}(E)$ and $\mathcal{L} \in \mathscr{C}(\mathcal{E})$ such that $\mathcal{L}_{\mid \mathcal{E}}=L$. The associated resolvent operators are denoted by $R(\xi):=(L-\xi)^{-1}$ for $\xi \notin \Sigma(L)$ and $\mathcal{R}(\xi):=(\mathcal{L}-\xi)^{-1}$ for $\xi \notin \Sigma(\mathcal{L})$.
Theorem 2.1. Assume there are some real numbers $a \in \mathbb{R}$, some complex numbers $\xi_{1}, \ldots, \xi_{k} \in \Delta_{a}, k \in \mathbb{N}$ (with the convention $\left\{\xi_{1}, \ldots, \xi_{k}\right\}=\emptyset$ if $k=0$ ) such that one has
(H1) Localization of the spectrum of $L: \Sigma(L) \cap \Delta_{a}=\left\{\xi_{1}, \ldots, \xi_{k}\right\}$;
(H2) Decomposition of $\mathcal{L}$ : there exist two closed unbounded operators $\mathcal{A}$ and $\mathcal{B}$ on $\mathcal{E}$ (with domains containing $\operatorname{Dom}(\mathcal{L}))$ such that $\mathcal{L}=\mathcal{A}+\mathcal{B}$, and
(i) $\Sigma(\mathcal{B}) \cap \Delta_{a}=\emptyset$;
(ii) $\mathcal{A} \in \mathscr{B}(\mathcal{E} \rightarrow E)$;
(iii) there is some $\xi_{0} \in \Delta_{a}$ such that $\mathcal{L}-\xi_{0}$ is invertible in $\mathcal{E}$.

Then $\mathcal{L}$ satisfies
(i) $\Sigma(\mathcal{L}) \cap \Delta_{a}=\left\{\xi_{1}, \ldots, \xi_{k}\right\}$;
(ii) for any $\xi \in \Delta_{a} \backslash\left\{\xi_{1}, \ldots, \xi_{k}\right\}$ the following bound for the resolvent operator holds:

$$
\begin{equation*}
\|\mathcal{R}(\xi)\|_{\mathscr{B}(\mathcal{E})} \leq\left\|(\mathcal{B}-\xi)^{-1}\right\|_{\mathscr{B}(\mathcal{E})}+\|R(\xi)\|_{\mathscr{B}(E)}\left\|\mathcal{A}(\mathcal{B}-\xi)^{-1}\right\|_{\mathscr{B}(\mathcal{E} \rightarrow E)} \tag{2.2}
\end{equation*}
$$

Remarks 2.2. (1) In words, assumption (H2) means that one may decompose $\mathcal{L}$ as $\mathcal{L}=\mathcal{A}+\mathcal{B}$ where $\mathcal{A}$ has a "good localization/regularization property" (in the generalized sense of the "change of space" (H2)-(ii)) and $\mathcal{B}$ has a "good spectral property" (in the sense that (H2)-(i) holds).
(2) One may relax (H2)-(i) into $\Sigma(\mathcal{B}) \cap \Delta_{a} \subset\left\{\xi_{1}, \ldots, \xi_{k}\right\}$.
(3) One may relax (H2)-(ii) into the fact that $\mathcal{A}(\mathcal{B}-\xi)^{-1}$ is bounded from $\mathcal{E}$ to $E$ for any $\xi \in \Delta_{a} \backslash\left\{\xi_{1}, \ldots, \xi_{k}\right\}$.
(4) In the theorem (and the previous remarks) one may replace $\Delta_{a} \backslash\left\{\xi_{1}, \ldots, \xi_{k}\right\}$ by any nonempty open connected set $\Omega \subset \mathbb{C}$.
(5) One may replace (H2)-(iii) by
(H2)-(iii') $\left\|(\mathcal{B}-\lambda)^{-1}\right\|_{\mathscr{B}(\mathcal{E})} \leq K /(\lambda-a)$ for some $K, a>0$ and for any $\lambda \in \mathbb{R}$, $\lambda>a$.
Indeed, assuming $\mathbf{( H 2 ) - ( i i i ' ) ~ o n e ~ m a y ~ f i n d ~} \xi_{0} \in \mathbb{R}, \xi_{0}>$ a such that $\| \mathcal{A}(\mathcal{B}-$ $\left.\xi_{0}\right)^{-1} \|_{\mathscr{B}(\mathcal{E})}<1$ and then $\mathcal{L}-\xi_{0}=\left(I d_{\mathcal{E}}+\mathcal{A}\left(\mathcal{B}-\xi_{0}\right)^{-1}\right)\left(\mathcal{B}-\xi_{0}\right)$ is invertible as the product of two invertible operators, which proves $(\mathbf{H} 2)$-(iii).

Proof of Theorem [2.1]. Define $\Omega:=\Delta_{a} \backslash\left\{\xi_{1}, \ldots, \xi_{k}\right\}$ and for $\xi \in \Omega$

$$
U(\xi):=\mathcal{B}(\xi)^{-1}-R(\xi) \mathcal{A} \mathcal{B}(\xi)^{-1}
$$

where $R(\xi)=(L-\xi)^{-1}$ and $\mathcal{B}(\xi):=\mathcal{B}-\xi$.
By assumptions (H1), (H2)-(i) and (H2)-(ii), $\mathcal{B}(\xi)^{-1}: \mathcal{E} \rightarrow \mathcal{E}, \mathcal{A B}(\xi)^{-1}: \mathcal{E} \rightarrow$ $E$ and $R(\xi): E \rightarrow E$ are bounded operators. Introduce now the canonical injection $J: E \rightarrow \mathcal{E}$. Taking into account that $R=J R, \mathcal{A}=J \mathcal{A}, \mathcal{L} J=J L$, the following identity holds:

$$
\begin{aligned}
(\mathcal{L}-\xi) R(\xi) \mathcal{A B}(\xi)^{-1} & =(\mathcal{L}-\xi) J R(\xi) \mathcal{A B}(\xi)^{-1}=J(L-\xi) R(\xi) \mathcal{A B}(\xi)^{-1} \\
& =J \operatorname{Id}_{E} \mathcal{A B}(\xi)^{-1}=J \mathcal{A B}(\xi)^{-1}=\mathcal{A B}(\xi)^{-1}
\end{aligned}
$$

The operator $U$ is a right-inverse of $\mathcal{L}-\xi$ since $U(\xi): \mathcal{E} \rightarrow \mathcal{E}$ is a well-defined bounded operator, and

$$
\begin{aligned}
(\mathcal{L}-\xi) U(\xi) & =(\mathcal{A}+\mathcal{B}(\xi)) \mathcal{B}(\xi)^{-1}-(\mathcal{L}-\xi) R(\xi) \mathcal{A B}(\xi)^{-1} \\
& =\mathcal{A B}(\xi)^{-1}+\operatorname{Id}_{\mathcal{E}}-(\mathcal{L}-\xi) R(\xi) \mathcal{A} \mathcal{B}(\xi)^{-1} \\
& =\mathcal{A B}(\xi)^{-1}+\operatorname{Id}_{\mathcal{E}}-\mathcal{A B}(\xi)^{-1}=\operatorname{Id}_{\mathcal{E}}
\end{aligned}
$$

As a consequence, $\mathcal{L}-\xi$ is onto. Moreover, if $\mathcal{L}-\xi$ is invertible we have $\mathcal{R}(\xi):=$ $(\mathcal{L}-\xi)^{-1}=U(\xi)$ and then $\mathcal{R}(\xi)$ satisfies the announced estimate (2.2). In order to prove that $\mathcal{L}-\xi$ is invertible we may argue as follows using assumption (H2)-(iii). It is well known that $\mathcal{L}-\eta$ invertible implies that $\mathcal{L}-\xi$ is invertible for any $\xi \in \mathbb{C}$ such that

$$
|\xi-\eta|<\|\mathcal{R}(\eta)\|_{\mathscr{B}(\mathcal{E})}^{-1}
$$

and

$$
\begin{equation*}
\mathcal{R}(\xi)=\mathcal{R}(\eta) \sum_{n=0}^{\infty}(\eta-\xi)^{n} \mathcal{R}(\eta)^{n} \tag{2.3}
\end{equation*}
$$

For a given $\xi \in \Omega$ we fix a path $\gamma$ from $\xi_{0}$ to $\xi$, that is a continuous function $\gamma:[0,1] \rightarrow \Omega$ such that $\gamma(0)=\xi_{0}, \gamma(1)=\xi$. Thanks to assumption (H2)-(ii) and the summation representation (2.3) for the operator $(\mathcal{B}-\xi)^{-1}$ we first deduce that

$$
\sup \left\{\left\|\mathcal{A}(\mathcal{B}-\gamma(t))^{-1}\right\|_{\mathscr{B}(\mathcal{E} \rightarrow E)} ; t \in[0,1]\right\}<\infty
$$

and then next that

$$
\inf \left\{\|U(\gamma(t))\|_{\mathscr{B}(\mathcal{E})}^{-1} ; t \in[0,1]\right\}>0
$$

We conclude that $\mathcal{L}-\xi$ is invertible, so that $\Sigma(\mathcal{L}) \cap \Delta_{a} \subset\left\{\xi_{1}, \ldots, \xi_{k}\right\}$.
Finally, since $N\left(L-\xi_{j}\right) \subset N\left(\mathcal{L}-\xi_{j}\right)$ for any $j=1, \ldots, k$, we also have $\left\{\xi_{1}, \ldots, \xi_{k}\right\} \subset$ $\Sigma(\mathcal{L}) \cap \Delta_{a}$.

Let us give quickly a sufficient condition in the spirit of this abstract section for answering the natural question about whether the geometric eigenspaces of $L$ and $\mathcal{L}$ associated with $\xi_{1}, \ldots, \xi_{k}$ are the same.

Proposition 2.3. Assume that $L$ satisfies (H1) with $k \geq 1$, (H2)-(i) and (H2)(ii), as well as

$$
\left(\mathcal{B}-\xi_{j}\right)[\operatorname{Dom}(L)] \supset \mathcal{A}\left[\mathcal{E}_{j}\right] \quad \text { for some } j \in\{1, \ldots, k\} .
$$

Then one has $N\left(\mathcal{L}-\xi_{j}\right)=N\left(L-\xi_{j}\right)$.
Proof of Proposition [2.3. The equation

$$
g \in \operatorname{Dom}(\mathcal{L}), \quad\left(\mathcal{L}-\xi_{j}\right) g=0
$$

implies $\left(\mathcal{B}-\xi_{j}\right) g=-\mathcal{A} g$ and from our assumption there exists $h \in \operatorname{Dom}(L)$ such that $\left(\mathcal{B}-\xi_{j}\right) h=-\mathcal{A} g$. As a consequence, we have $\left(\mathcal{B}-\xi_{j}\right)(g-h)=0$ and from assumption (H2)-(ii) we obtain $g=h \in \operatorname{Dom}(L)$ and $\left(L-\xi_{j}\right) g=0$, so that $g \in N\left(L-\xi_{j}\right)$.

## 3. SEmigroup decay versus Resolvent estimates

Consider $X$ a Banach space and $\Lambda \in \mathscr{C}(X)$. We shall make use of the following new assumptions:
(H1') Localization of the spectrum: $\Sigma(\Lambda) \cap \Delta_{a}=\left\{\xi_{1}, \ldots, \xi_{k}\right\}$ for some $a \in \mathbb{R}$, $k \in \mathbb{N}$, and some discrete eigenvalues $\xi_{1}, \ldots, \xi_{k} \in \Delta_{a}$ (isolated with finite dimension) with the convention $\left\{\xi_{1}, \ldots, \xi_{k}\right\}=\emptyset$ if $k=0$.
(H3) Line control on the resolvent: there exists some real numbers $a \in \mathbb{R}$, $K \in(0, \infty)$ such that

$$
\sup _{y \in \mathbb{R}}\|R(a+i y)\|_{\mathscr{B}(X)} \leq K
$$

(H3') Sectorial control on the resolvent: there exists some real numbers $a \in \mathbb{R}$, $K, \theta \in(0, \infty)$ such that

$$
\sup _{y= \pm \theta(x-a), x \leq a}\|R(x+i y)\|_{\mathscr{B}(X)} \leq K^{\prime}
$$

(H4) Mild control on the semigroup: $\Lambda$ generates a $C_{0}$-semigroup of bounded operators $e^{t \Lambda}$ on $X$, and more precisely there exists $b \in \mathbb{R}, C_{b} \geq 0$ such that

$$
\forall t \geq 0 \quad\left\|e^{t \Lambda}\right\|_{\mathscr{B}(X)} \leq C_{b} e^{b t}
$$

(H5) Accurate decomposition of the semigroup: there holds

$$
\begin{equation*}
\forall t \geq 0, \quad\left\|e^{t \Lambda}-\sum_{j=1}^{k} e^{t T_{j}} \Pi_{j}\right\|_{\mathscr{B}(X)} \leq C_{\tilde{a}} e^{\tilde{a} t} \tag{3.1}
\end{equation*}
$$

for some $\tilde{a} \in \mathbb{R}, k \in \mathbb{N}$, some constant $C_{\tilde{a}} \geq 1$, some finite dimensional projection operators $\Pi_{1}, \ldots, \Pi_{k}$ such that $\Pi_{i} \Pi_{j}=0$ for any $i \neq j$ and such that they all commute with the semigroup $e^{t \Lambda}$, some operators of the form $T_{j}=\xi_{j} I d_{X_{j}}+N_{j}$ with $\xi_{j} \in \Delta_{\tilde{a}}, X_{j}:=\Pi_{j} X, N_{j} \in \mathscr{B}\left(X_{j}\right)$ is a nilpotent operator.

Theorem 3.1. Consider a Banach space $X$ and $\Lambda \in \mathscr{C}(X)$.
(1) (General Banach case) Assume that (H1') and (H3) hold for the same real $a \in \mathbb{R}$ as well as $(\mathbf{H} 4)$. Then for any $\tilde{a}>a$ there exists a constant $C_{\tilde{a}} \geq 1$ (explicitly computable as a function of $\tilde{a}, a, b, C_{b}$ and $K$ ) such that

$$
\begin{equation*}
\forall t \geq 0, \quad\left\|e^{t \Lambda}-\sum_{j=1}^{k} e^{t \Lambda} \Pi_{\Lambda, \xi_{j}}\right\|_{\mathscr{B}(Y \rightarrow X)} \leq C_{\tilde{a}} e^{\tilde{a} t}, \tag{3.2}
\end{equation*}
$$

where $Y=D(\Lambda) \cap D\left(\Lambda^{2}\right)$ endowed with the norm $\|f\|_{Y}:=\|f\|_{X}+\|\Lambda f\|_{X}+$ $\left\|\Lambda^{2} f\right\|_{X}$.
(2) (General Hilbert case) Assume that (H1') and (H3) hold for the same real $a \in \mathbb{R}$ as well as $(\mathbf{H} 4)$. Assume furthermore that $X$ is a Hilbert space. Then (H5) holds for any $\tilde{a}>a$ and where $\Pi_{j}=\Pi_{\Lambda, \xi_{j}}$ is the spectral projector associated to $\xi_{j}$.
(3) (Sectorial case) Assume that ( $\mathbf{H} 1^{\prime}$ ) and ( $\mathbf{H} 3^{\prime}$ ) hold for the same real a. Then (H5) holds for any $\tilde{a}>a$ and where again $\Pi_{j}=\Pi_{\Lambda, \xi_{j}}$ is the spectral projector associated to $\xi_{j}$.
(4) (Partial reversed implication) Assume that (H5) holds. Then $\Lambda$ satisfies (H1'), (H3) for any $a \in\left(\tilde{a}, \min _{j} \Re e \xi_{j}\right)$, (H4) for any $b \geq \max _{j} \Re e \xi_{j}$ and $\Pi_{\Lambda, \xi_{j}}=\Pi_{j}$ for any $1 \leq j \leq k$.

Remarks 3.2. (1) In a Hilbert space, we see that the assumptions (H1), (H3), (H4) on the one hand, and the assumption (H5) on the other hand are "almost" equivalent, in the sense that we must chose $\tilde{a}>$ a for the direct implication, or $a>\tilde{a}$ for the reversed one.
(2) One could replace ( $\mathbf{H} 3)$ in this theorem by

$$
\begin{equation*}
\int_{0}^{\infty}\left\|e^{t \Lambda}\right\|_{\mathscr{B}(X)} e^{-b^{\prime} t} d t \leq C_{b^{\prime}} . \tag{3.3}
\end{equation*}
$$

Indeed, (H3) implies (3.3) for any $b^{\prime}>b$, and in the proof we only use (3.3).
(3) After finishing writing this paper, we were informed of the recent paper [17] which uses an idea similar to the proof below of point (2) of Theorem 3.1; use a Plancherel identity on the resolvent in Hilbert spaces in order to obtain explicit rates of decay on the semigroup in terms of bounds on the resolvent. Let us also add that this idea was more or less "well-known" in the spectral theory community (as is also acknowledged in [17]), even if it seems hard to
find a reference where it is done in a quantitative way as here and in [17]. In the present paper however, the focus is on the combination of this method with our factorization theorem in order to modify the space for the decay estimate on the semigroup.

Proof of Theorem 3.1.
Proof of point (1). Assume first that $\left\{\xi_{1}, \ldots \xi_{k}\right\}=\emptyset$. Starting from the identity

$$
\forall z \notin \Sigma(\Lambda), \quad R(z)=z^{-1}[-\operatorname{Id}+R(z) \Lambda]
$$

we deduce from (H2) that for any $f \in \operatorname{Dom}(\Lambda)$

$$
\begin{equation*}
\|R(a+i s) f\| \leq \frac{(1+K)}{|a+i s|}\left(\|f\|_{X}+\|\Lambda f\|_{X}\right) \rightarrow 0 \text { as }|s| \rightarrow \infty \tag{3.4}
\end{equation*}
$$

and it follows from the density of $\operatorname{Dom}(\Lambda)$ in $X$ that

$$
\forall f \in X, \quad\|R(z) f\|_{X} \rightarrow 0, \quad z=a+i s,|s| \rightarrow \infty
$$

Then consider for $M>0$

$$
I_{M}(f):=\int_{a-i M}^{a+i M} e^{z t} R(z) f d z
$$

Since $R$ is differentiable on the segment $[a-i M, a+i M]$ with $R^{\prime}(z)=R(z)^{2}$, we compute by integration by parts

$$
\begin{equation*}
I_{M}(f)=\frac{e^{(a+i M) t}}{t} R(a+i M) f-\frac{e^{(a-i M) t}}{t} R(a-i M) f-\int_{a-i M}^{a+i M} \frac{e^{z t}}{t} R(z)^{2} f d z \tag{3.5}
\end{equation*}
$$

We use (see for instance [26, Chapter 1, Section 1.7] and [31, Theorem 1.1]) the representation

$$
e^{t \Lambda} f=\frac{1}{2 i \pi} \int_{a-i \infty}^{a+i \infty} e^{z t} R(z) f d z:=\lim _{M \rightarrow \infty} \frac{1}{2 i \pi} I_{M}(f)
$$

From (3.4)-(3.5) this representation formula can be rewritten as

$$
\begin{equation*}
e^{t \Lambda} f=-\frac{1}{2 i \pi} \int_{a-i \infty}^{a+i \infty} \frac{e^{z t}}{t} R(z)^{2} f d z:=-\frac{1}{2 i \pi} \lim _{M \rightarrow \infty} \int_{a-i M}^{a+i M} \frac{e^{z t}}{t} R(z)^{2} f d z \tag{3.6}
\end{equation*}
$$

Then we use the identity

$$
\forall z \notin \Sigma(\Lambda), \quad R(z)=z^{-1}[-\operatorname{Id}+R(z) \Lambda]
$$

to deduce

$$
\left\|e^{t \Lambda} f\right\|_{X} \leq C \frac{e^{a t}}{t} \frac{1}{2 \pi}\left(\lim _{M \rightarrow \infty} \int_{a-i M}^{a+i M} \frac{d z}{|a+i z|^{2}}\right)\left(\|f\|_{X}+\|\Lambda f\|_{X}+\left\|\Lambda^{2} f\right\|_{X}\right)
$$

which concludes the proof.

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Proof of (2). As in the proof of point 1, we start from the representation formula (3.6). Let $f \in \operatorname{Dom}(\Lambda) \subset X$ and $\phi \in \operatorname{Dom}\left(\Lambda^{*}\right) \subset X^{*}$ (where $X^{*}$ denotes the dual space of $X$ and $\Lambda^{*}$ the adjoint operator of $\Lambda$ ). Let us estimate

$$
\begin{align*}
\left\langle\phi, e^{t \Lambda} f\right\rangle & =-\frac{1}{2 i \pi} \int_{a-i \infty}^{a+i \infty} \frac{e^{z t}}{t}\left\langle\phi, R(z)^{2} f\right\rangle d z \\
& :=-\frac{1}{2 i \pi} \lim _{M \rightarrow \infty} \int_{a-i M}^{a+i M} \frac{e^{z t}}{t}\left\langle\phi, R(z)^{2} f\right\rangle d z \tag{3.7}
\end{align*}
$$

Applying Cauchy-Schwartz inequality, assumption (H3) and the identity

$$
R(a+i s)=R(b+1+i s)+(a-b-1) R(a+i s) R(b+1+i s)
$$

we get

$$
\begin{aligned}
& \left|\int_{a-i M}^{a+i M} \frac{e^{z t}}{t}\left(\phi, R(z)^{2} f\right) d z\right| \leq \frac{e^{a t}}{t} \int_{-M}^{M}\left\|R(a+i s)^{*} \phi\right\|\|R(a+i s) f\| d s \\
& \leq(1+|b+1-a| K)^{2} \frac{e^{a t}}{t} \int_{-M}^{M}\left\|R(b+1+i s)^{*} \phi\right\|\|R(b+1+i s) f\| d s \\
& \leq(1+|b+1-a| K)^{2} \frac{e^{a t}}{t}\left(\int_{-M}^{M}\left\|R(b+1+i s)^{*} \phi\right\|^{2} d s\right)^{1 / 2} \\
& \quad \times\left(\int_{-M}^{M}\|R(b+1+i s) f\|^{2} d s\right)^{1 / 2}
\end{aligned}
$$

Plancherel's identity in $X$ (this is where we need that $X$ is a Hilbert space) and assumption (H2) imply:

$$
\begin{aligned}
\int_{\mathbb{R}}\|R(b+1+i s) f\|^{2} d s & =2 \pi \int_{0}^{+\infty}\left\|e^{-(b+1) t} e^{t \Lambda} f\right\|^{2} d t \\
& \leq\left(\int_{0}^{+\infty}\left\|e^{-(b+1) t} e^{t \Lambda}\right\|^{2} d t\right)\|f\|^{2} \\
& \leq\left(\int_{0}^{+\infty} C_{b} e^{-t} d t\right)\|f\|^{2}=C_{b}\|f\|_{X}^{2}
\end{aligned}
$$

Since (Hilbert space) $X^{*} \sim X$ and $\left\|e^{t \Lambda}\right\|=\left\|\left(e^{t \Lambda}\right)^{*}\right\|$, we have similarly

$$
\int_{-M}^{M}\left\|R(b+1+i s)^{*} \phi\right\|^{2} d s \leq \int_{0}^{+\infty}\left\|e^{-(b+1) t}\left(e^{t \Lambda}\right)^{*} \phi\right\|^{2} d t \leq C_{b}\|\phi\|_{X^{*}}^{2}
$$

Gathering the estimates above, we get

$$
\left|\left\langle\phi, e^{t \Lambda} f\right\rangle\right| \leq \underbrace{(1+|b+1-a| K)^{2} C_{b}}_{:=C} \frac{e^{a t}}{t}\|\phi\|_{X^{*}}\|f\|_{X}
$$

which implies

$$
\left\|e^{t \Lambda}\right\| \leq \frac{C e^{a t}}{t}
$$

and concludes the proof (by combining this last estimate with ( $\mathbf{H} 4$ ) for $t \leq 1$ ).

Suppose now that the set of isolated eigenvalues $\left\{\xi_{1}, \ldots, \xi_{k}\right\}$ is not empty. In this case, the representation is

$$
e^{t \Lambda}=\sum_{j=1}^{k} e^{\xi_{j} t} \Pi_{j}+\frac{1}{2 i \pi} \int_{a-i \infty}^{a+i \infty} \frac{e^{z t}}{t} R(z)^{2} d z
$$

and we proceed as above for the second term in the right-hand side.
Proof of point (3). The proof is classical and can be found in [26, Theorem 7.7 Chap 1] for instance.
Proof of point (4). On the one hand, define

$$
\Pi_{0}:=\operatorname{Id}-\Pi_{1}-\cdots-\Pi_{k} \text { and } S_{0}(t):=S(t) \Pi_{0}
$$

Its generator is $\Lambda \Pi_{0}$ and assumption (H5) together with the Hille-Yosida theorem (see [26, Theorem 5.3]) implies then $\Delta_{a} \cap \Sigma\left(\Lambda \Pi_{0}\right)=\emptyset$ and that the resolvent operator associate to $\Lambda \Pi_{0}$ satisfies (H3). On the other hand, since $\Lambda \Pi_{j}$ acts on a finite dimensional space, by the Jordan decomposition, one can find $\eta_{j, 1}, \ldots, \eta_{j, k_{j}}$ and projectors $\left(\Pi_{j, i}\right)$ such that

$$
\Pi_{j}=\Pi_{j, 1}+\ldots+\Pi_{j, k_{j}}, \quad \Pi_{j, i} \Pi_{j, \ell}=0 \forall i \neq \ell
$$

where $\Lambda \Pi_{j, i}$ is a matrix with only eigenvalue $\eta_{j, i}$ (writing as a diagonal part plus some nilpotent part) for any $1 \leq i \leq k_{j}$ and $1 \leq j \leq k$. Using assumption (H5) again, we have by composing by $\Pi_{j}$

$$
\begin{equation*}
\left\|e^{t \Lambda \Pi_{j}} \Pi_{j}-e^{t T_{j}} \Pi_{j}\right\|_{\mathscr{B}(X)} \leq C_{\tilde{a}} e^{\tilde{a} t} \tag{3.8}
\end{equation*}
$$

and next by composing by $\Pi_{j, i}$

$$
\begin{equation*}
\left\|e^{t \Lambda \Pi_{j, i}} \Pi_{j, i}-e^{t T_{j}} \Pi_{j, i}\right\|_{\mathscr{B}(X)} \leq C_{\tilde{a}} e^{\tilde{a} t} . \tag{3.9}
\end{equation*}
$$

By comparing the large time behavior $t \rightarrow+\infty$ of the two semigroups we obtain that necessarily $\eta_{j, i}=\xi_{j}$ for any $1 \leq i \leq k_{j}$, so that we can take $k_{j}=1$. In other words $\left\{\xi_{1}, \ldots, \xi_{k}\right\}$ are discrete eigenvalues of $\Lambda$ with associated eigenspaces $E_{j}:=\Pi_{j} X$ which is nothing but (H1').

Moreover by writing

$$
e^{t T_{j}}=e^{t \xi_{j}}\left(P_{j, 0}+t P_{j, 1}+\cdots+t^{k_{j}} P_{j, k_{j}}\right)
$$

for some new projectors $\left(P_{j, i}\right)$ such that

$$
P_{j}=P_{j, 1}+\ldots+P_{j, k_{j}}, \quad P_{j, i} P_{j, \ell}=0 \quad \forall i \neq \ell
$$

and writing a similar decomposition of the semigroup generated by the matrix $\Lambda \Pi_{j}$, we easily deduce $\Lambda \Pi_{j}=T_{j}$. We conclude that (H3) and (H4) hold by gathering (through a direct sum) the informations obtained on each subspace $\Pi_{j} X, 0 \leq j \leq k$.

Finally, define $\Pi:=\Pi_{1}-\cdots-\Pi_{k}$ so that $\Pi_{0}=\operatorname{Id}-\Pi$, and $\tilde{\Pi}:=\Pi_{\Lambda, \xi_{1}}+\cdots+\Pi_{\Lambda, \xi_{k}}$ so that $\tilde{\Pi}_{0}=\operatorname{Id}-\tilde{\Pi}$. By applying point (1) to $\Lambda$, we have

$$
\forall t \geq 0, \quad\left\|e^{t \Lambda}(\operatorname{Id}-\tilde{\Pi})\right\|_{\mathscr{B}(Y \rightarrow X)} \leq C_{\alpha} e^{\alpha t}
$$

for any $\alpha>\tilde{a}$. Composing by $\Pi_{0}$ we deduce with the help of (H5)
$\forall t \geq 0, \quad\left\|e^{t \Lambda} \tilde{\Pi} \Pi_{0}\right\|_{\mathscr{B}(Y \rightarrow X)} \leq\left\|e^{t \Lambda} \Pi_{0}-e^{t \Lambda} \tilde{\Pi} \Pi_{0}\right\|_{\mathscr{B}(Y \rightarrow X)}+\left\|e^{t \Lambda} \Pi_{0}\right\|_{\mathscr{B}(Y \rightarrow X)} \leq C_{\alpha} e^{\alpha t}$,
which implies (since the norms $\|\cdot\|_{Y}$ and $\|\cdot\|_{X}$ are equivalent on the finite dimensional subspace $\tilde{\Pi} X$, and $\Pi_{0}$ and $\tilde{\Pi}$ commute)

$$
\forall t \geq 0, \quad\left\|e^{t \Lambda \tilde{\Pi} \Pi_{0}}\right\|_{\mathscr{B}(X)} \leq C_{\alpha} e^{\alpha t}
$$

That last estimate together with the fact that $\Lambda \tilde{\Pi} \Pi_{0}$ is the restriction to the subspace $\tilde{\Pi} \Pi_{0} X$ of the matrix $\Lambda \tilde{\Pi}$ with spectrum $\Sigma(\Lambda \tilde{\Pi})=\left\{\xi_{1}, \ldots, \xi_{k}\right\}$ yields $\tilde{\Pi} \Pi_{0}=0$. Similarly, we obtain $\Pi \tilde{\Pi}_{0}=0$ and then

$$
\Pi=\Pi \tilde{\Pi}_{0}+\Pi \tilde{\Pi}=\tilde{\Pi} \Pi_{0}+\tilde{\Pi} \Pi=\tilde{\Pi}
$$

That last identity implies

$$
\sum_{j=1}^{k} e^{\Lambda t} \Pi_{\Lambda, \xi_{j}}=e^{\Lambda t} \tilde{\Pi}=e^{\Lambda t} \Pi=\sum_{j=1}^{k} e^{T_{j} t} \Pi_{j}
$$

Coming back to the complex plane integral definition of $\Pi_{\Lambda, \xi_{j}}$ we finally get that $e^{\Lambda t} \Pi_{\Lambda, \xi_{j}}=e^{T_{j} t} \Pi_{j}$ and therefore $\Pi_{\Lambda, \xi_{j}}=\Pi_{j}$.

We end this section by some recalls on the classical notion of dissipative operators together with the statement of a variant of the Lumer-Phillips theorem. The latter provides a very popular alternative method in order to prove assumption (H5) for a given operator. Its conclusion is stronger (the constant in front of the exponential is $C_{\tilde{a}}=1$, that is the operator is a contraction up to a translation) but its assumptions are also more restrictive. Nevertheless, in all the applications we have in mind, it is the starting point from which we can develop our space enlargement method of the spectral gap decay estimate.

Definition 3.3. Consider $X$ a Banach space, $\Lambda \in \mathscr{C}(X)$ and $a \in \mathbb{R}$. We say that $\Lambda-a$ is dissipative on a closed subspace $X_{0} \subset X$ if

$$
\begin{equation*}
\forall f \in D(\Lambda) \cap X_{0}, \exists \varphi \in F(f) \quad \Re e\langle\varphi,(\Lambda-a) f\rangle \leq 0 \tag{3.10}
\end{equation*}
$$

where $F(f) \subset X^{\prime}$ is the dual set of $f$ defined by

$$
F(f):=\left\{\varphi \in X^{\prime} ;\langle\varphi, f\rangle=\|f\|_{X}^{2}=\|\varphi\|_{X^{\prime}}^{2}\right\} .
$$

In the case when $X$ is a Hibert space $F(f)=\{f\}$ and (3.10) writes

$$
\begin{equation*}
\forall f \in D(\Lambda) \cap X_{0}, \quad \Re e(\Lambda f, f) \leq a\|f\|^{2} \tag{3.11}
\end{equation*}
$$

We say that $\Lambda-a$ is $m$-dissipative is furthermore

$$
R(\Lambda-a)=X
$$

Remark 3.4. Notice that $\Lambda$ satisfies (3.10) if an only if

$$
\forall f \in D(\Lambda), \forall \lambda>a \quad\|(\lambda-\Lambda) f\| \geq(\lambda-a)\|f\|
$$

We refer to [26, Chap 1, Theorem 4.2] for a proof of this claim.

Theorem 3.5. Consider $X$ a Banach space, $\Lambda \in \mathscr{C}(X)$. Assume that there exists $a, b \in \mathbb{R}$ and $\xi_{1}, \ldots, \xi_{k} \in \mathbb{C}$ such that $a<\min \Re e \xi_{j} \leq \max \Re e \xi_{j}<b$. Assume that there exits a closed subspace $X_{0}$ and some finite dimension subspaces $X_{1}, \ldots, X_{k}$ such that $X_{0} \oplus X_{1} \oplus \cdots \oplus X_{k}=X$ and
(i) $X_{j} \subset M\left(\Lambda-\xi_{j}\right)$ for any $j=1, \ldots, k$;
(ii) $X_{0}$ is invariant under the action of $\Lambda: \Lambda f \in X_{0}$ for any $f \in X_{0} \cap D(\Lambda)$;
(iii) $\Lambda-a$ is dissipative on $X_{0}$;
(iv) $\Lambda-b$ is m-dissipative on $X_{0}$ or on $X$.

Then $\Lambda$ generates a $C_{0}$-semigroup of bounded operators and (H5) with $\tilde{a}=a$ and $C_{\tilde{a}}=1$.
Sketch of the proof of Theorem 3.5. The operator $\Lambda_{\mid X_{j}}, j=1, \ldots, k$, generates a semigroup which writes $S_{j}(t):=e^{T_{j} t}$ where $T_{j}=\xi_{j} \operatorname{Id}_{X_{j}}+N_{j}, N_{j}$ nilpotent operator on $X_{j}$. We remark that $\Lambda-b$ is $m$-dissipative on $X_{0}$ due to (iv) and $X_{0} \oplus X_{1} \oplus \cdots \oplus X_{k}=X$. Then the classical Lumer-Phillips theorem (see [26, Chap 1, Theorem 4.3]) implies that $\Lambda_{\mid X_{0}}$ generates a semigroup $S_{0}(t)$. To conclude, we define

$$
S(t):=S_{0}(t) P_{0}+\cdots+S_{k}(t) P_{k}
$$

where $P_{j}$ is the projector on $X_{j}$. Using that each $P_{j}$ commutes with $\Lambda$, it is easily checked that $S(t)=e^{t \Lambda}$ by decomposing $f \in D(\Lambda)$ as $f=f_{0}+\cdots+f_{k}$ with $f_{j} \in X_{j} \cap D(\Lambda)$ so that $f_{0} \in D(\Lambda)$ and computing the time derivative of $S(t) f$.

## 4. Spectral gap estimates

In this section we combine the results obtained so far in order to prove that one can enlarge the functional space of decay estimates on semigroups.

Theorem 4.1. Assume that $L$ satisfies (H1') and

$$
\forall t \geq 0, \quad\left\|e^{t L}\left(I d-\sum_{j=1}^{k} \Pi_{j}\right)\right\|_{\mathscr{B}(E)} \leq C_{a} e^{a t}
$$

where $\Pi_{j}=\Pi_{L, \xi_{j}}, 1 \leq j \leq k$ denotes the (finite dimensional) spectral projectors of the eigenvalue $\xi_{j} \in \Delta_{a}$, and $\mathcal{L}$ satisfies $(\mathbf{H} 4)$ as well as the decomposition assumption (H2)-(ii), that is there exist two closed unbounded operators $\mathcal{A}$ and $\mathcal{B}$ (with domains containing $\operatorname{Dom}(\mathcal{L}))$ such that $\mathcal{L}=\mathcal{A}+\mathcal{B}$ and $\mathcal{A} \in \mathscr{B}(\mathcal{E} \rightarrow E)$. Then we distinguish the following cases:
(i) (Sectorial framework) $\mathcal{B}$ satisfies an assumption of type ( $\mathbf{H} 3^{\prime}$ ): more precisely there exists some real numbers $a \in \mathbb{R}$ and $K^{\prime}, \theta \in(0, \infty)$ such that

$$
\sup _{y= \pm \theta(x-a), x \leq a}\left\|(\mathcal{B}-(x+i y))^{-1}\right\|_{\mathscr{B}(X)} \leq K^{\prime}
$$

Then $\mathcal{L}$ is sectorial (with explicit estimates) and for any $\tilde{a}>a$ there exists an explicit constant $C_{\tilde{a}} \geq 1$ such that

$$
\begin{equation*}
\forall t \geq 0, \quad\left\|e^{t \mathcal{L}}\left(I d-\sum_{j=1}^{k} \Pi_{j}\right)\right\|_{\mathscr{B}(\mathcal{E})} \leq C_{\tilde{a}} e^{\tilde{a} t} \tag{4.1}
\end{equation*}
$$

(ii) (Hilbert space framework) $\mathcal{E}$ is an Hilbert space and $\mathcal{B}$ satisfies an assumption of type (H3): more precisely there exists some real numbers $a \in \mathbb{R}$ and $K \in(0, \infty)$ such that

$$
\begin{equation*}
\sup _{y \in \mathbb{R}}\left\|(\mathcal{B}-(a+i y))^{-1}\right\|_{\mathscr{B}(X)} \leq K \tag{4.2}
\end{equation*}
$$

Then $\mathcal{L}$ satisfies $(\mathbf{H} 1)$ and for any $\tilde{a}>a$ there exists an explicit constant $C_{\tilde{a}} \geq 1$ such that

$$
\forall t \geq 0, \quad\left\|e^{t \mathcal{L}}\left(I d-\sum_{j=1}^{k} \Pi_{j}\right)\right\|_{\mathscr{B}(\mathcal{E})} \leq C_{\tilde{a}} e^{\tilde{a} t}
$$

(iii) (General Banach space framework) $\mathcal{E}$ is a Banach space and $\mathcal{B}$ satisfies again an assumption of type (H3) in the form (4.2). Then $\mathcal{L}$ satisfies (H1) and for any $\tilde{a}>a$ there exists an explicit constant $C_{\tilde{a}} \geq 1$ such that

$$
\forall t \geq 0, \quad\left\|e^{t \mathcal{L}}\left(f-\sum_{j=1}^{k} \Pi_{j} f\right)\right\|_{\mathcal{E}} \leq C_{\tilde{a}} e^{\tilde{a} t}\left(\|f\|_{\mathcal{E}}+\|\mathcal{L} f\|_{\mathcal{E}}+\left\|\mathcal{L}^{2} f\right\|_{\mathcal{E}}\right)
$$

Proof of Theorem 4.1. The proof is immediate.
For the case (i) we combine uniform estimates on the resolvent of $L$ for $\xi \in\{y=$ $\pm \theta(x-a), x \leq a\}$ together with the sectorial estimate (H3') assumed on $\mathcal{B}$ to deduce by Theorem (2.1) sectorial estimates on $\mathcal{L}$ for $\xi \in\{y= \pm \theta(x-a), x \leq a\}$. We conclude by applying Theorem 3.1, point (3).

For the case (ii) we combine uniforme estimates on the resolvent of $L$ for $\{\Re e \xi=$ $a\}$ together with the estimate (H3) assumed on $\mathcal{B}$ to deduce by Theorem (2.1) uniform estimates of type ( $\mathbf{H} 3$ ) on the resolvent of $\mathcal{L}$ for $\{\Re e \xi=a\}$. We conclude by applying Theorem 3.1, point (2).

For the case (iii) we combine uniforme estimates on the resolvent of $L$ for $\{\Re e \xi=$ $a\}$ together with the estimate (H3) assumed on $\mathcal{B}$ to deduce by Theorem (2.1) uniform estimates of type (H3) on the resolvent of $\mathcal{L}$ for $\{\Re e \xi=a\}$. We conclude by applying Theorem 3.1, point (1).

In the following corollary we give a kind of "practical" case of application for PDEs, where the small space $E$ is a Hilbert space where the operator is self-adjoint and dissipative (typically $L^{2}\left(\mu^{-1}\right)$ where $\mu$ is the stationary distribution), and where one has some conservation law in the larger Banach space $\mathcal{E}$ in which we want to apply our space enlargement method.

Corollary 4.2 (Application to PDEs).
Consider a Hilbert space $E$ and a Banach space $\mathcal{E}$ such that $E \subset \mathcal{E}$ and $E$ is dense in $\mathcal{E}$. Consider two unbounded closed operators with dense domain $L$ on $E$ and $\mathcal{L}$ on $\mathcal{E}$ such that $\mathcal{L}_{\mid E}=L$.

- Assume in $E$ :
(i) There are $G_{1}, \ldots, G_{n} \in E$ linearly independent, and $\lambda_{1}, \ldots, \lambda_{n} \in \Delta_{\alpha}$ for some $\alpha \in \mathbb{R}$, such that $G_{j} \in M\left(L-\lambda_{j}\right)$ and $\left\|G_{j}\right\|_{E}=1$ for any $1 \leq j \leq n$ (note that the $\lambda_{j}$ are not necessarily distinct).
(ii) Defining $\psi_{j}(f):=\left\langle f, G_{j}\right\rangle_{E} G_{j}$ for any $1 \leq j \leq n$, the space

$$
E_{0}:=\left\{f \in E ; \forall j=1, \ldots, n, \psi_{j}(f)=0\right\}
$$

is invariant under the action of $L$, and $E=E_{0} \oplus \operatorname{Span}\left\{G_{1}, \ldots, G_{n}\right\}$.
(iii) $L-\alpha$ is dissipative on $E_{0}$ for some scalar product $\langle\langle\cdot, \cdot\rangle\rangle_{E}$ on $E$ equivalent to $\langle\cdot, \cdot\rangle_{E}$ :

$$
\forall f \in D(L) \cap E_{0}, \quad\langle\langle(L-\alpha) f, f\rangle\rangle_{E} \leq 0
$$

(iv) $L-b$ is $m$-disssipative in $E_{0}$ or $E$ for some $b \geq \alpha$.

- Assume in $\mathcal{E}$ :
(v) $\mathcal{L}$ decomposes as $\mathcal{L}=\mathcal{A}+\mathcal{B}, \mathcal{A}, \mathcal{B} \in \mathscr{C}(\mathcal{E})$, with $\mathcal{A}$ bounded from $\mathcal{E}$ to $E$ and $\mathcal{B}-\alpha$ is dissipative.
(vi) There exists $\Psi_{j} \in \mathscr{B}(\mathcal{E} \rightarrow E)$ such that $\Psi_{j \mid E}=\psi_{j}$.

Then, $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}=\Sigma(\mathcal{L}) \cap \Delta_{a}, G_{1}, \ldots, G_{n}$ is a base of the algebraic eigenspaces of $\mathcal{L}$ associated to the eigenvalues $\lambda_{1}, \ldots, \lambda_{n}$, and, introducing the notation $\xi_{1}, \ldots, \xi_{k}$ for the distinct eigenvalues, that is $\left\{\xi_{1}, \ldots, \xi_{k}\right\}=\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$ and $\xi_{j} \neq \xi_{i}$ for $i \neq j$, we have $\Pi_{\mathcal{L}, \xi_{j} \mid E}=\Pi_{L, \xi_{j}}$ for any $1 \leq j \leq k$. Moreover, the space

$$
\mathcal{E}_{0}:=\left\{f \in \mathcal{E} ; \forall j=1, \ldots, n, \Psi_{j}(f)=0\right\}
$$

is invariant under the action of $\mathcal{L}$ and for any $a>\alpha$ there exists $C_{a} \geq 1$ such that for any $f \in X \cap \mathcal{E}_{0}$ there holds:

$$
\forall t \geq 0 \quad\left\|e^{t \mathcal{L}} f\right\|_{\mathcal{E}} \leq C_{a} e^{a t}\|f\|_{X}
$$

where $X=\mathcal{E}$ if $\mathcal{E}$ is a Hilbert space and $X=D\left(\mathcal{L}^{2}\right) \cap D(\mathcal{L})$ endowed with the norm of the graphs of $\mathcal{L}$ and $\mathcal{L}^{2}$ if $\mathcal{E}$ is a Banach space.
Remark 4.3. Note that in the case where $L$ is self-adjoint in the small space $E$ (which however is not always satisfied in the following applications, in particular for inhomogeneous kinetic models), then there can be no nilpotent part in the action of the operator in the eigenspaces.

Proof of Corollary 4.2. First from (i), (ii), (iii), (iv) and Theorem 3.5, the operator $L$ satisfies (H1'), generates a semigroup $e^{t L}$ on $E$ and more precisely satisfies the bound

$$
\forall t \geq 0, \quad\left\|e^{t L}\left(\operatorname{Id}-\sum_{j=1}^{k} \Pi_{j}\right)\right\|_{\mathscr{B}(E)} \leq C_{a} e^{a t}
$$

Next, because $\left.A\right|_{E} \in \mathscr{B}(E)$, the operator $\left.\mathcal{B}\right|_{E}=L-\left.A\right|_{E}$ also generates a semigroup on $E$. We claim that

$$
\begin{equation*}
\forall f \in E, \forall t \geq 0 \quad\left\|e^{t \mathcal{B}} f\right\|_{\mathcal{E}} \leq e^{C t}\|f\|_{\mathcal{E}} \tag{4.3}
\end{equation*}
$$

for some constant $C \in \mathbb{R}$. Indeed, for any $f \in D(L)$ we may compute (in the case where $\mathcal{E}$ is a Hilbert space)

$$
\frac{d}{d t}\left\|e^{t \mathcal{B}} f\right\|_{\mathcal{E}}^{2}=2 \Re e\left\langle\mathcal{B} e^{t \mathcal{B}} f, e^{t \mathcal{B}} f\right\rangle \leq 2 \alpha\left\|e^{t \mathcal{B}} f\right\|_{\mathcal{E}}^{2}
$$

and we deduce (4.3) with $C=\alpha$. In the general Banach case, we can argue as follows. For any $\varepsilon>0$ we introduce the norm $\|f\|_{\epsilon}:=\|f\|_{\mathcal{E}}+\epsilon\|f\|_{E}$ so that $\|\cdot\|_{\epsilon}$ is
equivalent to $\|\cdot\|_{E}$ (for any $\epsilon>0$ ). Define $C:=\alpha+\|\mathcal{A}\|_{\mathscr{B}(\mathcal{E} \rightarrow E)}$. Using the second definition recalled in Remark 3.4 and the fact that $L-\alpha$ is dissipative in $E$, it is clear that $\mathcal{B}-C$ is $m$-dissipative in the Banach space $\left(E,\|\cdot\|_{\epsilon}\right.$ ). The Lumer-Phillips theorem says that $\mathcal{B}-C$ generates a semigroups of contractions on $\left(E,\|\cdot\|_{\epsilon}\right)$, in particular for any $f \in E, t \geq 0$,

$$
\left\|e^{t(\mathcal{B}-C)} f\right\|_{\mathcal{E}}+\epsilon\left\|e^{t(\mathcal{B}-C)} f\right\|_{E} \leq\|f\|_{\mathcal{E}}+\epsilon\|f\|_{E} .
$$

Letting $\epsilon$ goes to 0 , we obtain (4.3). Together with a density argument, this implies that $\mathcal{B}$ generates a $C_{0}$-semigroup on $\mathcal{E}$ and more precisely $\mathcal{B}-C$ is $m$-dissipative on $\mathcal{E}$. Using again that $\mathcal{A} \in \mathscr{B}(\mathcal{E})$ we deduce then that $\mathcal{L}=\mathcal{A}+\mathcal{B}$ generates a $C_{0}$-semigroup and more precisely that $\mathcal{L}$ satisfies (H4) with a growth rate smaller than $b=C+\|\mathcal{A}\|_{\mathscr{B}(\mathcal{E})}$.

Because of Remark 3.4, the line control (4.2) on the resolvent of $\mathcal{B}$ on any line $\{\xi=a+i y, y \in \mathbb{R}\}, a>\alpha$, is consequence of the fact $\mathcal{B}-\alpha$ is dissipative. The assumptions of Theorem 4.1 being fulfilled, we deduce that for any $f \in X$

$$
\forall t \geq 0 \quad\left\|e^{t \mathcal{L}} f-\sum_{j=1}^{k} e^{\mathcal{L} t} \Pi_{\mathcal{L}, \xi_{j}} f\right\|_{\mathcal{E}} \leq C_{a} e^{a t}\|f\|_{X}
$$

where by definition $\Pi_{\mathcal{L}, \xi_{j}}$ is the spectral projection on the eigenspace associated to the eigenvalue $\xi_{j}$ in $\mathcal{E}$ through the integral formula (2.1).

In order to conclude we have to prove that $\Pi_{\mathcal{L}, \xi_{j}} f=\Pi_{j}$ for any $f \in \mathcal{E}$, where we define

$$
\Pi_{j}:=\sum_{\ell} \Pi_{j, \ell} \quad \Pi_{j, \ell}:=\sum_{i, \lambda_{i}=\xi_{j}, d\left(G_{i}\right)=\ell} \Psi_{i}
$$

and $d\left(G_{i}\right)$ stands for the smallest $q \in \mathbb{N}$ such that $\left(L-\lambda_{i}\right)^{q} G_{i}=0$. First, for $f \in E \cap Y$, we write for some polynomial function $P_{j, \ell}(t)$

$$
\begin{aligned}
& \left\|\sum_{j=1}^{k} e^{\mathcal{L} t} \Pi_{\mathcal{L}, \xi_{j}} f-\sum_{j=1}^{k} \sum_{\ell} e^{\xi_{j} t} P_{j, \ell}(t) \Pi_{j, \ell} f\right\|_{\mathcal{E}} \leq \\
& \leq\left\|\sum_{j=1}^{k} e^{\mathcal{L} t} \Pi_{\mathcal{L}, \xi_{j}} f-e^{t \mathcal{L}} f\right\|_{\mathcal{E}}+\left\|e^{t L} f-\sum_{j=1}^{k} \sum_{\ell} e^{\xi_{j} t} P_{j, \ell}(t) \Pi_{j, \ell} f\right\|_{\mathcal{E}} \leq C_{a} e^{a t},
\end{aligned}
$$

because $\Pi_{L, \xi_{j}}=\Pi_{j}$ as a consequence of the first step and point (4) in Theorem 3.1, When the $\xi_{j}$ are real numbers, the fact that $P_{j, \ell}$ has exactly degree $\ell$ implies that all the the functions $e^{\xi_{j} t} P_{j, \ell}(t)$ have different growth when $t \rightarrow \infty$. By induction, we deduce that $\Pi_{\mathcal{L}, \xi_{j}} f=\Pi_{j} f$ for any $1 \leq j \leq k$. With similar arguments (and taking into account the frequency of oscillation) we have the same conclusion in the general case where the eigenvalues are not purely real. Because both semigroups are continuous on $\mathcal{E}$ the same identity holds on the whole space $\mathcal{E}$.

Remarks 4.4. (1) The assumptions made on $E$ in Corollary 4.2 are nothing but those made on $E$ in Theorem 3.5 and are then equivalent (up to a slight translation of value of $a \in \mathbb{R}$ ) to know an accurate decomposition (H5) of the semigroup $e^{t L}$.
(2) In section 5 and 6 below, we apply our space enlargement of the decay of semigroups from $E$ to $\mathcal{E} \supset E$. Since the conditions on $E$ are already known results and condition (ii)-(vi) is a straightforward consequence of the conservation laws of the equation in $\mathcal{E}$, the only thing we have to do in the following application is to exhibit a suitable decomposition $\mathcal{L}=\mathcal{A}+\mathcal{B}$ for which (v) holds.

## 5. Application to Fokker-Planck equations

5.1. The space homogeneous Fokker-Planck equation. Consider the equation

$$
\begin{equation*}
\partial_{t} f=L f:=\nabla_{v} \cdot\left(\nabla_{v} f+F f\right), \quad f(\cdot, 0)=f_{i n}(\cdot), \tag{5.1}
\end{equation*}
$$

for $f=f(t, v), t \geq 0, v \in \mathbb{R}^{d}$ and $F=F(v) \in \mathbb{R}^{d}$ which takes the form

$$
\begin{equation*}
F=\nabla_{v} \phi+U \tag{5.2}
\end{equation*}
$$

Assume
(FP1) The potential $\phi: \mathbb{R}^{d} \rightarrow \mathbb{R}$ satisfies: $\mu(d v):=e^{-\phi(v)} d v$ is a probability measure and (strong version of the Poincaré inequality) for some $\nu>0$ and for any $f$ such that $\int f d v=0$ one has

$$
\begin{equation*}
\int\left|\nabla\left(\frac{f}{\mu}\right)\right|^{2} \mu(d v) \geq \nu \int f^{2}\left(1+|\nabla \phi|^{2}\right) \mu^{-1}(d v) \tag{5.3}
\end{equation*}
$$

(FP2) The additional force field $U$ satisfies

$$
\forall v \in \mathbb{R}^{d}, \quad \nabla \cdot U=0, \quad \nabla_{v} \phi \cdot U=0, \quad|U| \leq C(1+|\nabla \phi|)
$$

It is immediate to check that $L(\mu)=0$ and

$$
\langle L f, f\rangle_{L^{2}\left(\mu^{-1}\right)}:=\int L f f \mu^{-1}(d v)=-\int|\nabla(f / \mu)|^{2} \mu(d v) \leq 0
$$

Let us first recall more or less well-known results about the spectrum and spectral gap properties of the Fokker-Planck equation in the space $L^{2}\left(\mu^{-1}\right)$.
Theorem 5.1. Assume that $F$ satisfies (FP1)-(FP2). Then in the space $E=$ $L^{2}\left(\mu^{-1}\right)$, the operator $L$ satisfies: there exists $\lambda>0$ such that $\Sigma(L) \subset\{z \in$ $\mathbb{C} \mid \operatorname{Re}(z) \leq-\lambda\} \cup\{0\}$, the null space of $L$ is exactly $\mathbb{R} \mu$, and any solution to the initial value problem (5.1) 5.2) for any initial datum $f_{i n} \in E$ satisfies

$$
\forall t \geq 0, \quad\left\|f(t)-\mu\left\langle f_{i n}\right\rangle\right\|_{E} \leq e^{-\lambda t}\left\|f_{i n}-\left\langle f_{i n}\right\rangle\right\|_{E}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d v
$$

Remark 5.2. Equation (5.3) is a strenghtened version of the classical Poincaré inequality

$$
\int\left|\nabla\left(\frac{f}{\mu}\right)\right|^{2} \mu(d v) \geq \nu \int f^{2} \mu^{-1}(d v)
$$

The latter is satisfied for instance under the so-called Bakry-Emery condition Hess $(\phi) \geq$ $\nu$ Id (see [4, 28, 3]).

A (more or less well-known) sufficient condition for the strenghtened form (5.3) is for instance $\alpha|\nabla \phi|^{2}-2 \Delta \phi \geq c$ for $\alpha \in(0,1)$ and $c \in \mathbb{R}$ and for any $|v| \geq R$ for some $R>0$ (see [25] for a quantitative proof).

Let us now consider a weight function $m(v):=e^{-\theta(\phi(v))}$ with $\theta \in C^{2}$ and the associated Hilbert space $\mathcal{E}=L^{2}\left(m^{-1}\right)$ such that $E \subset \mathcal{E}$. The new result is a spectral gap estimate in this larger space under some specific assumptions on $m$. Let us denote by $\mathcal{L}$ the operator (5.1)5.2) in the space $\mathcal{E}$.

Assume that $m$ satisfies:
(FP3) The function

$$
\psi(v):=\frac{1}{2 m} \nabla_{v} \cdot\left(\mu \nabla_{v}\left(\frac{m}{\mu}\right)\right)
$$

satisfies

$$
\psi(v) \xrightarrow[|v| \rightarrow \infty]{\longrightarrow}-\infty
$$

Remarks 5.3. (1) Other technical assumptions could have been chosen for the function $m$ in the proof below, however the formulation (FP3) seems to us the most natural one since it is based on the comparison of the FokkerPlanck operators for two different force field. In any case, the core idea in the decomposition in the proof below is that a coercive $\mathcal{B}$ in $\mathcal{E}$ is obtained by a negative local perturbation of the whole operator.
(2) Remark that by mollification the smoothness of $m$ could be relaxed: if $m$ is not smooth but $\tilde{m}$ is smooth, satisfies (FP3) and is such that $c_{1} m<\tilde{m} \leq c_{2} m$, then
$\left\|f_{t}-\mu\right\|_{L^{2}(m)} \leq c_{2}\left\|f_{t}-\mu\right\|_{L^{2}(\tilde{m})} \leq C e^{-\lambda_{1} t}\left\|f_{i n}-\mu\right\|_{L^{2}(\tilde{m})} \leq C^{\prime} e^{-\lambda_{1} t}\left\|f_{i n}-\mu\right\|_{L^{2}(m)}$.
Many interesting $\phi$ and $m$ sastisfy our assumptions (we omit the proof which is straightforward calculations):

Proposition 5.4. One has

$$
\psi(v)=\frac{1}{2}\left[\left|\nabla_{v} \phi\right|^{2}\left(\theta^{\prime}(\phi)^{2}-\theta^{\prime}(\phi)-\theta^{\prime \prime}(\phi)\right)+\Delta_{v} \phi\left(1-\theta^{\prime}(\phi)\right)\right]
$$

and admissible potentials and weights for (FP1)-(FP3) are for instance
(i) $\phi(v)=C\langle v\rangle^{s}, s \geq 1$ and $\theta(z)=z^{\alpha}, \alpha \in(0,1)$ with $s(1+\alpha)>2$;
(ii) $\phi(v)=C\langle v\rangle^{s}, s \geq 1$ and $\theta(z)=k \log z$ with $s>2$ and $k s>d+s-2$.

Theorem 5.5. Assume that $F$ satisfies (FP1)-(FP2) and m satisfies (FP3). Then for any $\lambda_{1}$ with $0<\lambda_{1}<\lambda$ (where $\lambda$ is given in Theorem 5.1) such $\Sigma(\mathcal{L}) \subset\{z \in$ $\left.\mathbb{C} \mid \Re e(z) \leq-\lambda_{1}\right\} \cup\{0\}$ and there exists $C_{\lambda_{1}} \in[1, \infty)$ such that, for any initial data $f_{\text {in }} \in \mathcal{E}$, equation (5.1 5.2) has a solution $f \in \mathcal{E}$ such that

$$
\forall t \geq 0 \quad\left\|f(t)-\mu\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}} \leq C_{\lambda_{1}} e^{-\lambda_{1} t}\left\|f_{i n}-\mu\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d v
$$

Remarks 5.6. (1) The case of equality $\lambda_{1}=\lambda$ could easily been obtained by considering a more detailed expansion of the semigroup for the first eigenvalues in the use of Theorem 5.1, and showing that the eigenspaces are the same in $E$ and $\mathcal{E}$.
(2) One could prove variant of this theorem for some Banach large space $\mathcal{E}$ such as $L^{1}\left(m^{-1}\right)$ (under some different conditions on $m$ ).
(3) Let us make a last important remark. One could have thought to another natural method of proof by interpolation: using the exponential relaxation in $E$ together with a uniform bound in $L^{1}$ (by mass conservation and preservation of nonnegativity) one could (hopefully) try to obtain some rate of decay in some intermediate spaces. This method would yield probably quite complicated rates. Moreover it would not be satisfactory in at least two aspects: first it would not recover optimal rate of decay, and second, most importantly, it would not apply to semigroups which do not preserve sign and therefore the $L^{1}$ norm, such as those obtained by linearization (e.g. the linearized Boltmann equation in section 7).

Proof of Theorem 5.5. We prove that one can apply Corollary 4.2 with $n=1$ (1-dimensional null space spanned by $\mu$ ), $E=L^{2}\left(\mu^{-1}\right), \mathcal{E}=L^{2}\left(m^{-1}\right)$ and $\psi(f)=$ $\langle f, \mu\rangle_{E} \mu$. Because theorem 5.1] implies that assumptions (i), (ii), (iii) and (iv) hold and the mass conservation along the flow implies that assumption (vi) holds, the only thing to do is to verify the operator spiting assumption (v).

Let us write the following splitting:

$$
L=L^{s}+L^{a s}, \quad L^{s}=\nabla_{v} \cdot\left(\nabla_{v} f+f \nabla_{v} \phi\right), \quad L^{a s}=\nabla_{v}(U f)
$$

on $E$ and similarly $\mathcal{L}=\mathcal{L}^{s}+\mathcal{L}^{a s}$ on the space $\mathcal{E}$. The operator $L^{s}$ is symmetric in $E$, since

$$
\left\langle L^{s} f, g\right\rangle_{E}=-\int \nabla\left(\frac{f}{\mu}\right) \cdot \nabla\left(\frac{g}{\mu}\right) \mu d v
$$

and the operators $L^{a s}$ and $\mathcal{L}^{a s}$ are both anti-symmetric in resp. $E$ and $\mathcal{E}$ by immediate computation (using (FP2) and that $m$ is a function of $\phi$ ).

One can compute

$$
\left\langle\mathcal{L}^{s} f, f\right\rangle_{\mathcal{E}}=-\int\left|\nabla_{v}\left(\frac{f}{m}\right)\right|^{2} m d v+\int f^{2} \psi(v) m^{-1}(d v) .
$$

Consider $\chi \in C_{c}^{\infty}\left(\mathbb{R}^{d}\right), 0 \leq \chi \leq 1, \chi \equiv 1$ on $B(0,1),\|\chi\|_{\infty} \leq 1$ and define $\chi_{R}(v)=\chi\left(R^{-1} v\right), R \geq 1$. Consider $R$ such that $\psi(v) \leq 0$ for $|v| \geq R$ and $M$ such that $|\psi(v)| \leq M / 2$ for $|v| \leq M$. We then define the decompositions $\mathcal{L}=\mathcal{A}+\mathcal{B}$ with

$$
\mathcal{A} f:=\chi_{R} M f \text { and } \mathcal{B} f:=\mathcal{L} f-\chi_{R} M f .
$$

The operator $\mathcal{A}$ is clearly bounded from $\mathcal{E}$ into $E$ thanks to the cutoff function, and for $\xi$ with real part $a$ :

$$
\begin{aligned}
\langle(\mathcal{B}-\xi) f, f\rangle_{E}= & \langle(\mathcal{L}-\xi) f, f\rangle_{\mathcal{E}}-\int \chi_{R} M f^{2} m^{-1}(d v) \\
& =\left\langle\left(\mathcal{L}^{s}-\xi\right) f, f\right\rangle_{\mathcal{E}}-\int \chi_{R} M f^{2} m^{-1}(d v) \\
=- & \int\left|\nabla_{v}\left(\frac{f}{m}\right)\right|^{2} m d v+\int f^{2}\left(\psi(v)-M \chi_{R}-\Re e \xi\right) m^{-1}(d v) .
\end{aligned}
$$

We then use that $\psi(v)-M \chi_{R} \leq 0$ and $\psi(v)-M \chi_{R} \rightarrow-\infty$ as $|v| \rightarrow \infty$ to deduce that (taking maybe larger $M$ and $R$ )

$$
\left(\psi(v)-M \chi_{R}-\Re e \xi\right) \leq-C<0
$$

which proves that $\mathcal{B}-a$ is dissipative.
5.2. The kinetic Fokker-Planck equation. Consider the equation

$$
\begin{equation*}
\partial_{t} f=L f:=\nabla_{v} \cdot\left(\nabla_{v} f+v f\right)-v \cdot \nabla_{x} f+\nabla_{x} \Phi \cdot \nabla_{v} f, \quad f(\cdot, 0)=f_{i n}(\cdot), \tag{5.4}
\end{equation*}
$$

for $f=f(t, x, v), t \geq 0, x, v \in \mathbb{R}^{d}$ and $\Phi=\Phi(x) \in \mathbb{R}^{d}$.
Assume
(FP1') The potential $\Phi: \mathbb{R}^{d} \rightarrow \mathbb{R}$ satisfies: $\pi(d x):=e^{-\Phi(x)} d x$ is a probability measure and (strong version of the Poincaré inequality) for some $\zeta>0$ and for any $f=f(x) \in L^{2}\left(\pi^{-1}\right)$ such that $\in f d x=0$

$$
\begin{equation*}
\int\left|\nabla\left(\frac{f}{\pi}\right)\right|^{2} \pi(d x) \geq \zeta \int f^{2}\left(1+|\nabla \Phi|^{2}\right) \pi^{-1}(d x) \tag{5.5}
\end{equation*}
$$

Let us denote

$$
\mu(v):=\frac{1}{(2 \pi)^{-d / 2}} e^{-|v|^{2} / 2}
$$

and

$$
\gamma(x, v):=\pi \mu=\frac{1}{(2 \pi)^{-d / 2}} e^{-|v|^{2} / 2-\Phi(x)}
$$

which is a probability measure on $\mathbb{R}^{d} \times \mathbb{R}^{d}$.
Let us first recall a recent result proved in [14, 13] for the kinetic Fokker-Planck equation in the space $L^{2}\left(\gamma^{-1}\right)$.

Theorem 5.7. Assume that $\Phi$ satisfies ( $\mathbf{F P 1}^{\prime}$ ). Then in the space $E=L^{2}\left(\gamma^{-1}\right)$, the operator $L$ satisfies: there exists $C>0$ and $\lambda>0$ such that $\Sigma(L) \subset\{z \in$ $\mathbb{C} \mid \operatorname{Re}(z) \leq-\lambda\} \cup\{0\}$, the null space of $L$ is exactly $\operatorname{Span}(\{\gamma\})$, and any solution to the initial value problem (5.4) for any initial datum $f_{\text {in }} \in E$ satisfies

$$
\forall t \geq 0, \quad\left\|f(t)-\gamma\left\langle f_{i n}\right\rangle\right\|_{E} \leq C e^{-\lambda t}\left\|f_{i n}-\gamma\left\langle f_{i n}\right\rangle\right\|_{E}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d x d v
$$

Let us denote $e(x, v)=|v|^{2} / 2+\Phi(x)$ and consider some weight $m=e^{-\theta(e)}$ with $\theta \in C^{2}$ and the associated Hilbert space $\mathcal{E}=L^{2}\left(m^{-1}\right)$ such that $E \subset \mathcal{E}$. The new result is a spectral gap estimate in this larger space under some specific assumptions on $m$. Let us denote by $\mathcal{L}$ the operator (5.4) in the space $\mathcal{E}$.

Assume that $m$ satisfies:
(FP3') The function

$$
\Psi(x, v):=\frac{1}{2 m} \nabla_{v} \cdot\left(\gamma \nabla_{v}\left(\frac{m}{\gamma}\right)\right)
$$

satisfies

$$
\Psi(x, v) \xrightarrow[|e(x, v)| \rightarrow \infty]{ }-\infty
$$

Theorem 5.8. Assume that $\Phi$ satisfies ( $\mathbf{F P} \mathbf{1}^{\prime}$ ) and $m$ satisfies ( $\mathbf{F P} \mathbf{3}^{\prime}$ ). Then for any $\lambda_{1}$ with $0<\lambda_{1}<\lambda$ (where $\lambda$ is given in Theorem 5.7) such $\Sigma(\mathcal{L}) \subset\{z \in$ $\left.\mathbb{C} \mid \Re e(z) \leq-\lambda_{1}\right\} \cup\{0\}$ and there exists $C_{\lambda_{1}} \in[1, \infty)$ such that, for any initial data $f_{\text {in }} \in \mathcal{E}$, equation (5.4) has a solution $f \in \mathcal{E}$ such that

$$
\forall t \geq 0 \quad\left\|f(t)-\gamma\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}} \leq C_{\lambda_{1}} e^{-\lambda_{1} t}\left\|f_{i n}-\gamma\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d x d v
$$

Remarks 5.9. (1) Again one could obtain the case of equality $\lambda_{1}=\lambda$ for the rate of convergence by considering a more detailed expansion of the semigroup for the first eigenvalue.
(2) It would be straightforward (and simpler) to adapt the previous theorem in the case of the kinetic Fokker-Planck equation in the periodic torus $x \in T^{d}$.
(3) Another possible modification could be to extend the decay estimate to larger Banach $L^{1}$-type weighted spaces but it would require specific additional work, since the operator $\mathcal{A}$ in the decomposition below does not map $L^{1}$ into $L^{2}$ whatever the weights. We refer to section 7 where it is done for the linearized Boltzmann equation.
Proof of Theorem 5.8. Again we prove that one can apply Corollary 4.2 with $n=1$ (1-dimensional null space spanned by $\gamma$ ), $E=L^{2}\left(\gamma^{-1}\right), \mathcal{E}=L^{2}\left(m^{-1}\right)$ and $\psi(f)=$ $\langle f, \gamma\rangle_{E} \gamma$.

Let us write the following splitting:

$$
L=L^{s}+L^{a s}, \quad L^{s}=\nabla_{v} \cdot\left(\nabla_{v} f+v f\right), \quad L^{a s}=-v \cdot \nabla_{x} f+\nabla_{x} \Phi \cdot \nabla_{v} f
$$

on $E$ and similarly $\mathcal{L}=\mathcal{L}^{s}+\mathcal{L}^{a s}$ on the space $\mathcal{E}$. The operator $L^{s}$ is symmetric in $E$, since

$$
\left\langle L^{s} f, g\right\rangle_{E}=-\int \nabla\left(\frac{f}{\gamma}\right) \cdot \nabla\left(\frac{g}{\gamma}\right) \gamma d x d v
$$

and the operators $L^{a s}$ and $\mathcal{L}^{a s}$ are both anti-symmetric in resp. $E$ and $\mathcal{E}$ by immediate computation (using that $-v \cdot \nabla_{x} e+\nabla_{x} \Phi \cdot \nabla_{v} e=0$ and that $m$ is a function of $e$ ).

One can compute

$$
\left\langle\mathcal{L}^{s} f, f\right\rangle_{\mathcal{E}}=-\int\left|\nabla_{v}\left(\frac{f}{m}\right)\right|^{2} m d x d v+\int f^{2} \psi(v) m^{-1}(d x d v)
$$

Consider $\chi \in C_{c}^{\infty}\left(\mathbb{R}^{d}\right), 0 \leq \chi \leq 1, \chi \equiv 1$ on $B(0,1),\|\chi\|_{\infty} \leq 1$ and define $\chi_{R}(v)=\chi\left(R^{-1} v\right), R \geq 1$. Consider $R$ such that $\Psi(x, v) \leq 0$ for $|e(x, v)| \geq R$ and $M$ such that $|\Psi| \leq M / 2$ for $|e(x, v)| \leq M$. We then define the decomposition $\mathcal{L}=\mathcal{A}+\mathcal{B}$ with

$$
\mathcal{A} f:=\chi_{R}(e) M f \text { and } \mathcal{B} f:=\mathcal{L} f-\chi_{R}(e) M f
$$

The operator $\mathcal{A}$ is clearly bounded from $E$ into $\mathcal{E}$ thanks to the cutoff function. For the invertibility of $\mathcal{B}-\xi$ we compute

$$
\begin{aligned}
\langle(\mathcal{B}-\xi) f, f\rangle_{E}= & \langle(\mathcal{L}-\xi) f, f\rangle_{\mathcal{E}}-\int \chi_{R} M f^{2} m^{-1}(d x d v) \\
& =\left\langle\left(\mathcal{L}^{s}-\xi\right) f, f\right\rangle_{\mathcal{E}}-\int \chi_{R}(e) M f^{2} m^{-1}(d x d v) \\
=- & \int\left|\nabla_{v}\left(\frac{f}{m}\right)\right|^{2} m d v+\int f^{2}\left(\Psi(e)-M \chi_{R}(e)-\Re e \xi\right) m^{-1}(d v) .
\end{aligned}
$$

We then use that $\Psi(e)-M \chi_{R}(e) \leq 0$ and $\psi(e)-M \chi_{R}(e) \rightarrow-\infty$ as $|e| \rightarrow \infty$ to deduce that (taking maybe larger $M$ and $R$ )

$$
\left(\Psi(e)-M \chi_{R}(e)-\Re e \xi\right) \leq-C<0
$$

which proves that $\mathcal{B}-a$ is dissipative.

## 6. Application to the linear Boltzmann equation

6.1. The space homogeneous linear Boltzmann equation. Let us consider as a first example the spatially homogeneous linear Boltzmann equation

$$
\begin{equation*}
\partial_{t} f=L f:=\int_{\mathbb{R}^{d}}\left(k\left(v, v_{*}\right) f\left(v_{*}\right)-k\left(v_{*}, v\right) f(v)\right) d v_{*} \tag{6.1}
\end{equation*}
$$

for $f=f(t, v), t \geq 0, v \in \mathbb{R}^{d}$ and $k\left(v, v_{*}\right) \geq 0$.
Assume
(LBE1) There is $\mu=\mu(v)$ probability measure such that

$$
k\left(v, v_{*}\right)=b\left(v, v_{*}\right) \mu(v) \text { and } 0<b_{*} \leq b\left(v, v_{*}\right)=b\left(v_{*}, v\right) \leq b^{*}<\infty
$$

Remark 6.1. More general setting without detailed balance could be considered within framework. We do not try to optimize the assumptions here but rather show how our method apply to Boltzmann equations.

It is well-known that

$$
\langle L f, f\rangle_{E}=-\frac{1}{2} \int b\left(v, v_{*}\right)\left(\frac{f(v)}{\mu(v)}-\frac{f\left(v_{*}\right)}{\mu\left(v_{*}\right)}\right)^{2} d v d v_{*}
$$

and that the homogeneous linear Boltzmann equation with such a kernel has a spectral gap in the Hilbert space $E=L^{2}\left(\mu^{-1}\right)$ :

Theorem 6.2. Assume that $k$ satisfies (LB1). Then in the space $E=L^{2}\left(\mu^{-1}\right)$, the operator $L$ satisfies: there exists $\lambda>0$ such that $\Sigma(L) \subset\{z \in \mathbb{C} \mid \operatorname{Re}(z) \leq$ $-\lambda\} \cup\{0\}$, the null space of $L$ is exactly $\mathbb{R} \mu$, and any solution to the initial value problem (6.1) for any initial datum $f_{\text {in }} \in E$ satisfies

$$
\forall t \geq 0, \quad\left\|f(t)-\mu\left\langle f_{i n}\right\rangle\right\|_{E} \leq e^{-\lambda t}\left\|f_{i n}-\mu\left\langle f_{i n}\right\rangle\right\|_{E}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d v
$$

Let us now consider a measurable weight function $m(v)>0$ and the associated Hilbert space $\mathcal{E}=L^{2}\left(m^{-1}\right)$ such that $E \subset \mathcal{E}$. The new result is a spectral gap estimate in this larger space under some specific assumptions on $m$. Let us denote by $\mathcal{L}$ the operator (6.1) in the space $\mathcal{E}$.

Assume that $m$ satisfies:
(LB2) The $L^{1}$ norm is bounded from above by the norm $\mathcal{E}$ times some constant and the norm $\mathcal{E}$ is controlled from above by the norm $E$ times some constant.

Theorem 6.3. Assume that $k$ satisfies (LB1) and $m$ satisfies (LB2). Then there exists a positive constant $\lambda_{1}$ with $0<\lambda_{1} \leq \lambda$ (where $\lambda$ is given in Theorem 6.2. such $\Sigma(\mathcal{L}) \subset\left\{z \in \mathbb{C} \mid \Re e(z) \leq-\lambda_{1}\right\} \cup\{0\}$ and there exists $C_{\lambda_{1}} \in[1, \infty)$ such that, for any initial data $f_{\text {in }} \in \mathcal{E}$, equation (6.1) has a solution $f \in \mathcal{E}$ such that

$$
\forall t \geq 0 \quad\left\|f(t)-\mu\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}} \leq C_{\lambda_{1}} e^{-\lambda_{1} t}\left\|f_{i n}-\mu\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d v
$$

Proof of Theorem 6.3. We only sketch the proof which is similar to the one of Theorem 5.5 for the spatially homogeneous Fokker-Planck equation.

We use the decomposition $\mathcal{L}=\mathcal{A}+\mathcal{B}$ with

$$
\mathcal{A} f:=\int k\left(v, v_{*}\right) f\left(v_{*}\right) d v_{*} \text { and } \mathcal{B} f:=\int k\left(v, v_{*}\right) f(v) d v_{*} .
$$

The fact that $\mathcal{A}$ is bounded from $\mathcal{E}$ to $E$ is clear:

$$
\begin{aligned}
\|\mathcal{A} f\|_{E}^{2} & =\int(\mathcal{A} f)^{2} \mu^{-1} \leq \int\left(\int b^{*}\left|f\left(v_{*}\right)\right| d v_{*}\right)^{2} \mu(v) d v \\
& \leq\left(b^{*}\right)^{2}\|f\|_{L^{1}}^{2} \leq C\|f\|_{\mathcal{E}}^{2}
\end{aligned}
$$

Finally dissipativity of $\mathcal{B}$ is clear from the fact that $\mathcal{B} f=\nu(v) f$ with $\nu \geq \nu_{0}>0$.

### 6.2. The space inhomogeneous linear Boltzmann equation in the torus.

 Consider$$
\begin{equation*}
\partial_{t} f=L f:=\int_{\mathbb{R}^{d}}\left(k\left(v, v_{*}\right) f\left(v_{*}\right)-k\left(v_{*}, v\right) f(v)\right) d v_{*}-v \cdot \nabla_{x} f \tag{6.2}
\end{equation*}
$$

for $f=f(t, x, v), t \geq 0, x \in \mathbb{T}^{d}$ (d-dimensional torus), $v \in \mathbb{R}^{d}$ and $k\left(v, v_{*}\right) \geq 0$.
Assume (LB1) on $k$ again.

Remark 6.4. Again more general settings could be considered: most importantly the case of a confining potential in $x \in \mathbb{R}^{d}$. We leave this interesting question for $a$ follow-up paper.

Let us denote

$$
\mu(v):=\frac{1}{(2 \pi)^{-d / 2}} e^{-|v|^{2} / 2},
$$

and assume without loss of generality that the torus has unit measure. Then $\mu$ is a probability measure on $\mathbb{R}^{d} \times \mathbb{R}^{d}$.

Let us first recall a recent result proved in [14, 13] for the kinetic Fokker-Planck equation in the space $L^{2}\left(\mu^{-1}\right)$.

Theorem 6.5. Assume that $k$ satisfies (LB1) and $\Phi$ satisfies (LB3). Then in the space $E=L^{2}\left(\gamma^{-1}\right)$, the operator $L$ satisfies: there exists $\lambda>0$ and $C>0$ such that $\Sigma(L) \subset\{z \in \mathbb{C} \mid \operatorname{Re}(z) \leq-\lambda\} \cup\{0\}$, the null space of $L$ is exactly $\mathbb{R} \mu$, and any solution to the initial value problem (6.2) for any initial datum $f_{i n} \in E$ satisfies

$$
\begin{equation*}
\forall t \geq 0, \quad\left\|f(t)-\mu\left\langle f_{i n}\right\rangle\right\|_{E} \leq C e^{-\lambda t}\left\|f_{i n}-\mu\left\langle f_{i n}\right\rangle\right\|_{E} \tag{6.3}
\end{equation*}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d x d v
$$

Consider some weight $m=m(v)$ and the associated Hilbert space $\mathcal{E}=L^{2}\left(m^{-1}\right)$ such that $E \subset \mathcal{E}$. The new result is a spectral gap estimate in this larger space under some specific assumptions on $m$. Let us denote by $\mathcal{L}$ the operator (6.2) in the space $\mathcal{E}$.

Assume that $m$ satisfies:
(LB2') The $L^{1}$ norm (in $x$ and $v$ ) is bounded from above by the norm $\mathcal{E}$ times some constant and the norm $\mathcal{E}$ (in $x$ and $v$ ) is controlled from above by the norm $E$ (in $x$ and $v$ ) times some constant.

Theorem 6.6. Assume that $k$ satisfies (LB1), $\Phi$ satisfies (LB3) and $m$ satisfies (LB2'). Then for any $\lambda_{1}$ with $0<\lambda_{1}<\lambda$ (where $\lambda$ is given in Theorem 6.5) such $\Sigma(\mathcal{L}) \subset\left\{z \in \mathbb{C} \mid \Re e(z) \leq-\lambda_{1}\right\} \cup\{0\}$ and there exists $C_{\lambda_{1}} \in[1, \infty)$ such that, for any initial data $f_{\text {in }} \in \mathcal{E}$, equation (5.4) has a solution $f \in \mathcal{E}$ such that

$$
\forall t \geq 0 \quad\left\|f(t)-\mu\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}} \leq C_{\lambda_{1}} e^{-\lambda_{1} t}\left\|f_{i n}-\mu\left\langle f_{i n}\right\rangle\right\|_{\mathcal{E}}
$$

where we denote

$$
\left\langle f_{i n}\right\rangle:=\int f_{i n} d x d v
$$

Remarks 6.7. (1) As before the case of equality $\lambda_{1}=\lambda$ for the rate of convergence could be obtained by refining the expansions on the semigroup.
(2) This would be again possible to extend the decay estimate to larger Banach $L^{1}$-type weighted spaces, see section 7 where it is done for the linearized Boltzmann equation.

Proof of Theorem 6.3. We only sketch the proof which is exactly similar to the one of Theorem 5.8 for the spatially inhomogeneous Fokker-Planck equation (it is in fact
even simpler since the confinement is ensured by the periodic domain and not by some external potential).

Define the decomposition $\mathcal{L}=\mathcal{A}+\mathcal{B}$ with

$$
\mathcal{A} f:=\int k\left(v, v_{*}\right) f\left(v_{*}\right) d v_{*} \text { and } \mathcal{B} f:=\int k\left(v, v_{*}\right) f(v) d v_{*}-v \cdot \nabla_{x} f
$$

Then the operator $\mathcal{A}$ is again clearly bounded from $\mathcal{E}$ into $E$, and the dissipativity of $\mathcal{B}$ is clear from

$$
\langle\mathcal{B} f, f\rangle_{\mathcal{E}}=\nu(v)\|f\|_{\mathcal{E}}^{2}
$$

since $v \cdot \nabla_{x}$ is anti-symmetric in $\mathcal{E}$ as $m$ is a function of $v$.

## 7. Application to the nonlinear Boltzmann equation

7.1. The Boltzmann equation. Consider the Boltzmann equation for hard potentials with cutoff (including hard spheres), which writes in the spatially homogeneous case

$$
\begin{equation*}
\partial_{t} f=Q(f, f)-v \cdot \nabla_{x} f \tag{7.4}
\end{equation*}
$$

for $f=f(t, x, v), x \in \mathbb{T}^{d}$ (d-dimensional torus), $v \in \mathbb{R}^{d}$ and

$$
\begin{equation*}
Q(f, f)=\int_{\mathbb{R}^{d}} \int_{\mathbb{S}^{d-1}}\left[f\left(v_{*}^{\prime}\right) f\left(v^{\prime}\right)-f(v) f\left(v_{*}\right)\right] b(\cos \theta)\left|v-v_{*}\right|^{\zeta} d v_{*} d \sigma \tag{7.5}
\end{equation*}
$$

with

$$
v^{\prime}=\frac{v+v_{*}}{2}+\sigma \frac{\left|v-v_{*}\right|}{2}, \quad v_{*}^{\prime}=\frac{v+v_{*}}{2}-\sigma \frac{\left|v-v_{*}\right|}{2}
$$

and where $b$ is a positive integrable function on the sphere $\mathbb{S}^{d-1}$ with $\cos \theta=\sigma \cdot(v-$ $\left.v_{*}\right) /\left|v-v_{*}\right|$, and $\zeta \in(0,1]$.

Assume without loss of generality that the torus has volume one. Then the unique global equilibrium (see [12] for instance) is

$$
\mu(v):=\frac{1}{(2 \pi)^{-d / 2}} e^{-|v|^{2} / 2}
$$

The linearized equation around this equilibrium is

$$
\begin{equation*}
\partial_{t} f=L(f)=C(f)-v \cdot \nabla_{x} f \tag{7.6}
\end{equation*}
$$

for $f=f(t, x, v), x \in \mathbb{T}^{d}, v \in \mathbb{R}^{d}$ and

$$
\begin{aligned}
C(f):=\int_{\mathbb{R}^{d}} \int_{\mathbb{S}^{d-1}}\left[\mu\left(v_{*}^{\prime}\right) f\left(v^{\prime}\right)\right. & +\mu\left(v^{\prime}\right) f\left(v_{*}^{\prime}\right) \\
& \left.-\mu\left(v_{*}\right) f(v)-\mu(v) f\left(v_{*}\right)\right] b(\cos \theta)\left|v-v_{*}\right|^{\zeta} d v_{*} d \sigma
\end{aligned}
$$

It is well-known that $L$ is self-adjoint non-positive and has a spectral gap in the Hilbert space $\bar{E}=L^{2}\left(\mu^{-1} d v\right)$ (see [7, 15, 16] and then [5] for explicit estimates). This result would be easily extended to $H^{1}\left(\mu^{-1} d v\right)$.

In [22], this spectral gap estimate on the semigroup was extended to the space $\overline{\mathcal{E}}:=L^{1}\left(e^{a|v|^{s}} d v\right)$ for $a>0$ and $s \in(0,2)$. Let us denote $\mathcal{C}$ and $\mathcal{L}$ the operators $C$ and $L$ when considered resp. in $L^{1}\left(e^{a|v|^{s}} d v\right)$ and $\mathcal{E}_{0}:=L^{1}\left(e^{a|v|^{s}} d x d v\right)$. Let us denote $E:=H^{1}\left(\mu^{-1} d x d v\right)$.

We shall extend the estimate of [22] to the full inhomogeneous case in the torus. First we shall recall the decomposition of $L$ devised in [22], then recall an hypocoercivity property in the torus proved in [24], then introduce a diffusive approximation and finally combine these preliminary steps with Corollary 4.2 (in the Banach space version) in order to extend decay estimates in $L^{1}\left(e^{a|v|^{s}} d x d v\right)$ in the torus, first for the diffusive approximation, and second for the orginal equation by relaxing the diffusion parameter (a key point is that the estimates are made uniform in terms of the diffusion with the help of the hypocoercivity theory). In the last subsection we shall turn to the nonlinear Boltzmann equation.
7.2. The decomposition of [22]. Let us state the key decomposition result which was proved in [22, Propositions 2.1-2.3-2.4-2.5-2.6]:

Proposition 7.8. There exists a decomposition

$$
\left\{\begin{array}{l}
C f=\bar{A} f+\bar{B} f  \tag{7.7}\\
\mathcal{C} f=\overline{\mathcal{A}} f+\overline{\mathcal{B}} f
\end{array}\right.
$$

such that $\bar{A}$ and $\overline{\mathcal{A}}$ (resp. $\bar{B}$ and $\overline{\mathcal{B}}$ ) coincide on $\bar{E} \cap \overline{\mathcal{E}}$ and such that $\overline{\mathcal{A}}$ is bounded from $\overline{\mathcal{E}}$ into $\bar{E}$ and also from $\overline{\mathcal{E}}$ into $H^{1}\left(\mu^{-1} d v\right)$ and $\bar{B}-\xi$ and $\overline{\mathcal{B}}-\xi$ are dissipative and invertible resp. in $\bar{E}\left(\right.$ or $\left.H^{1}\left(\mu^{-1} d v\right)\right)$ and $\overline{\mathcal{E}}$, for any $\Re e \xi=a$ for some $a<0$.
7.3. Hypocoercivity in the small Hilbert space. We have the following theorem from [24, Theorem 1.1]:

Theorem 7.9. In the space $E=H^{1}\left(\mu^{-1} d x d v\right)$, the operator $L$ satisfies: there exists $\lambda>0$ such that $\Sigma(L) \subset\{z \in \mathbb{C} \mid \operatorname{Re}(z) \leq-\lambda\} \cup\{0\}$, the null space of $L$ is spanned by $\mu, v_{i} \mu, 1 \leq i \leq d$ and $|v|^{2} \mu$, and any solution to the initial value problem (7.6) for any initial datum $f_{\text {in }} \in E$ satisfies

$$
\forall t \geq 0, \quad\left\|f(t)-\Pi f_{i n}\right\|_{E} \leq C e^{-\lambda t}\left\|f_{i n}-\Pi f_{i n}\right\|_{E}
$$

for some $C \geq 1$, where we denote $\Pi f_{\text {in }}$ the orthogonal projection in $L^{2}\left(\mu^{-1} d x d v\right)$ onto the null space of $L$.
7.4. A diffusive approximation. Let us introduce the following approximate equation:

$$
\begin{equation*}
\partial_{t} f=L_{\eta}(f)=C(f)-v \cdot \nabla_{x} f+\eta \Delta_{x} f \tag{7.8}
\end{equation*}
$$

for any $\eta>0$ and the corresponding approximate operator $\mathcal{L}_{\eta}$ in $\mathcal{E}$. Then Theorem 7.9 can be extended in the following way to this approximate equation:

Theorem 7.10. In the space $E=H^{1}\left(\mu^{-1} d x d v\right)$, there exists $\lambda>0$ such that for any $\eta \in(0,1]$, the operator $L_{\eta}$ satisfies: $\Sigma(L) \subset\{z \in \mathbb{C} \mid \operatorname{Re}(z) \leq-\lambda\} \cup\{0\}$, the null space of $L$ is spanned by $\mu, v_{i} \mu, 1 \leq i \leq d$ and $|v|^{2} \mu$, and any solution to the initial value problem (7.8) for any initial datum $f_{\text {in }} \in E$ satisfies

$$
\forall t \geq 0, \quad\left\|f(t)-\Pi f_{i n}\right\|_{E} \leq C e^{-\lambda t}\left\|f_{i n}-\Pi f_{i n}\right\|_{E}
$$

for some $C \geq 1$, where we denote $\Pi f_{\text {in }}$ the orthogonal projection in $L^{2}\left(\mu^{-1} d x d v\right)$ onto the null space of $L$.

Remark 7.11. Note the crucial point: the estimate from below $\lambda$ on the spectral gap is independent of the viscosity parameter $\eta$.

Proof of Theorem 7.10, Here we need to go back to the proof of [24, Theorem 1.1] in [24, Section 2]. The core of the proof there is the construction of a norm equivalent to $H^{1}$ in the form

$$
\begin{aligned}
\|f\|_{*}^{2}=A\|f\|_{L^{2}\left(\mu^{-1}\right)}^{2}+ & \alpha\left\|\nabla_{x} f\right\|_{L^{2}\left(\mu^{-1}\right)}^{2} \\
& +\beta\left\|\left(\nabla_{v}+v / 2\right) f\right\|_{L^{2}\left(\mu^{-1}\right)}^{2}+\gamma\left\langle\nabla_{x} f,\left(\nabla_{v}+v / 2\right) f\right\rangle_{L^{2}\left(\mu^{-1}\right)}
\end{aligned}
$$

by choosing $\beta>0$ big enough, then $A>0$ big enough, then $\gamma>0$ enough, then $\alpha>0$ big enough (in the paper [24] all computations are done for $h=\mu^{-1 / 2} f$ in $L^{2}$ without weight, and when translating these computations in $L^{2}\left(\mu^{-1}\right)$ one should be careful with $\left.\nabla_{v} h=\mu^{-1 / 2}\left(\nabla_{v} f+v / 2\right) f\right)$.

Then let us compute the action of the term $\eta \Delta_{x} f$ in the time-derivative of this norm:

$$
\begin{aligned}
\eta\left\langle\Delta_{x} f, f\right\rangle_{*} & =-A \eta\left\|\nabla_{x} f\right\|_{L^{2}\left(\mu^{-1}\right)}^{2}-\alpha \eta\left\|\nabla_{x}^{2} f\right\|_{L^{2}\left(\mu^{-1}\right)}^{2} \\
& -\beta \eta\left\|\nabla_{x}\left(\nabla_{v}+v / 2\right) f\right\|_{L^{2}\left(\mu^{-1}\right)}^{2}-\gamma \eta\left\langle\nabla_{x}^{2} f, \nabla_{x}\left(\nabla_{v}+v / 2\right) f\right\rangle_{L^{2}\left(\mu^{-1}\right)}^{2} .
\end{aligned}
$$

This only term of the right-hand side which is not non-positive is the fourth one, and by taking $\alpha$ big enough (possibly taking an $\alpha$ slightly larger than in the original proof in [24]), it can be controlled by the second and third terms of the right-hand side, uniformly in terms of $\eta$. Finally the choice of $\alpha$ big enough is also uniform in $\eta$. This concludes the proof.
7.5. Localization of the spectrum in a larger Banach space for the approximate operator. Let us consider for $\ell>1+d / 2$ the Banach space

$$
\mathcal{E}_{\ell}:=W_{x}^{\ell, 1} L_{v}^{1}\left(e^{a|v|^{s}} d x d v\right)
$$

defined by the norm

$$
\|f\|_{\mathcal{E}_{\ell}}:=\sum_{i=0}^{\ell} \int_{\mathbb{T}^{d} \times \mathbb{R}^{d}}\left|\nabla_{x}^{i} f\right| e^{a|v|^{s}} d x d v
$$

In this subsection we shall prove that the spectrum is unchanged from $E$ to $\mathcal{E}_{\ell}$. This will allow to justify the estimates on the resolvent with loss of derivatives which will be done in the next subsection.

Proposition 7.12. In the space $\mathcal{E}_{\ell}$, for any $\eta \in(0,1]$, the operator $\mathcal{L}_{\eta}$ satisfies: $\Sigma\left(\mathcal{L}_{\eta}\right) \subset\{\xi \in \mathbb{C} \mid \Re e(\xi) \leq-\lambda\} \cup\{0\}$ (where $\lambda$ is given by Theorem (7.10) and the null space of $L_{\eta}$ is spanned by $\left\{\mu, v_{1} \mu, \ldots, v_{d} \mu,|v|^{2} \mu\right\}$.
Proof of Theorem 7.12. We start from the decompositions

$$
L_{\eta}=A+B_{\eta}, \quad \mathcal{L}_{\eta}=\mathcal{A}+\mathcal{B}_{\eta}
$$

with

$$
\left\{\begin{array}{l}
A f:=\bar{A} f \text { and } B_{\eta} f:=\bar{B} f-v \cdot \nabla_{x} f+\eta \Delta_{x} f \\
\mathcal{A} f:=\overline{\mathcal{A}} f \text { and } \mathcal{B}_{\eta} f:=\overline{\mathcal{B}} f-v \cdot \nabla_{x} f+\eta \Delta_{x} f
\end{array}\right.
$$

Then first we prove that $\mathcal{A}$ is compact relatively to $\mathcal{B}_{\eta}$. This is a trivial consequence of the following facts: (1) the operator $\bar{A}$ regularizes in $v$ and cuts large velocities $v \in \mathbb{R}^{d}$, (2) the Dirichlet form of $B_{\eta}$ controls the $L^{1}$ norm of the gradient in $x \in \mathbb{T}^{d}$.

Proceeding as in the proof of [22, Proposition 3.4] (variant of Weyl's theorem [18, Chapter 4, Section 5] in Banach spaces based on the classification of the spectrum by the Fredholm theory), we deduce that the essential spectrum of $\mathcal{L}_{\eta}$ is the same as the one of $\mathcal{B}_{\eta}$. Moreover $\mathcal{B}_{\eta}$ is dissipative with constant $\lambda$ independent of $\eta$ since

$$
\left\langle\left(\mathcal{B}_{\eta}-\xi\right) f, f\right\rangle_{\mathcal{E}} \leq \sum_{i=0}^{\ell} \int\left(\int(\overline{\mathcal{B}}-\xi)\left(\nabla_{x}^{i} f\right) \operatorname{sign}\left(\nabla_{x}^{i} f\right) e^{a|v|^{s}} d v\right) d x
$$

and for each $i$ and any fixed $x \in \mathbb{T}^{d}$ we use the dissipativity of $\overline{\mathcal{B}}$ in $L^{1}\left(e^{a|v|^{s}} d v\right)$. Hence the essential spectrum of $\mathcal{L}_{\eta}$ is included in $\{z \in \mathbb{C} \mid \operatorname{Re}(z) \leq-\lambda\}$.

Then let us show that the discrete spectrum of $L_{\eta}$ and $\mathcal{L}_{\eta}$ are the same in the region $\{\xi \in \mathbb{C} \mid \Re e(\xi)>-\lambda\}$ : consider an isolated eigenvalue $\xi$ of $\mathcal{L}_{\eta}$ in $\mathcal{E}_{\ell}$ with multiplicity 1 (in case of higher multiplicity a similar argument can be performed on each Jordan block) and write the eigenvalue equation

$$
\mathcal{A} f=\xi f-\mathcal{B}_{\eta} f
$$

with $\Re e(\xi)>-\lambda$. Then by using the dissipativity of $\mathcal{B}_{\eta}$ in $\mathcal{E}_{\ell}$ and $E$ and the fact that $\mathcal{A}$ is bounded from $\mathcal{E}_{\ell}$ to $E$ (it is immediate since $\mathcal{A}_{0}$ is bounded from $L^{1}\left(e^{a|v|^{s}} d v\right)$ into $H_{v}^{1}\left(\mu^{-1}\right)$ and by Sobolev embedding $W_{x}^{\ell, 1} \rightarrow H_{x}^{1}$ in the torus as $\ell>1+d / 2$ ), we deduce that $f \in E$ and therefore $\xi$ is an eigenvalue of $L_{\eta}$ in $E$. The converse implication (that the eivenvalues of $L_{\eta}$ are eigenvalues of $\mathcal{L}_{\eta}$ ) is trivial from the fact that these eigenvalues are known in an explicit ways (note that a more general argument could be the use of the regularizing properties in $v$ of $A$, and in $x$ of $\eta \Delta_{x}$ ).

### 7.6. Hypocoercivity in a larger Banach space.

Theorem 7.13. In the space $\mathcal{E}_{\ell}$ with $\ell>d / 2+3$, the semigroup of the operator $\mathcal{L}$ satisfies: for any $\lambda_{1} \in(0, \lambda)$, any solution to the initial value problem (7.6) for any initial datum $f_{\text {in }} \in \mathcal{E}_{\ell}$ satisfies

$$
\forall t \geq 0, \quad\left\|f(t)-\Pi f_{i n}\right\|_{\mathcal{E}_{0}} \leq C_{1} e^{-\lambda_{1} t}\left\|\langle v\rangle^{2}\left(f_{i n}-\Pi f_{i n}\right)\right\|_{\mathcal{E}_{\ell}}
$$

for some $C_{1} \geq 1$, where we denote $\Pi$ the orthogonal projection in $L^{2}\left(\mu^{-1}\right)$ onto the null space of $L$.

Remark 7.14. As before the optimal rate $\lambda_{1}=\lambda$ is reachable by refining the expansion of the semigroup.

Proof of Theorem 7.13. We first work at the level of the approximate operator $\mathcal{L}_{\eta}$, and we start from the already used representation (see for instance [26, Chapter 1,

Section 1.7] and [31, Theorem 1.1])

$$
\begin{align*}
e^{t \mathcal{L}_{\eta}}(f-\tilde{\Pi} f)=\frac{1}{2 i \pi} & \int_{a-i \infty}^{a+i \infty} e^{z t} \mathcal{R}_{\eta}(z) f d z  \tag{7.9}\\
& =\lim _{M \rightarrow \infty} \frac{1}{2 i \pi} \int_{\lambda_{1}-i M}^{\lambda_{1}+i M} e^{z t} \mathcal{R}_{\eta}(z) f d z:=\lim _{M \rightarrow \infty} \frac{1}{2 i \pi} I_{M, \eta}(f)
\end{align*}
$$

for any $\lambda_{1}>\lambda$ and where $\mathcal{R}_{\eta}$ is the resolvent of $\mathcal{L}_{\eta}$ in $\mathcal{E}_{\ell}$ and $\tilde{\Pi}$ the spectral projector of $\mathcal{L}_{\eta}$ for the eigenvalue 0 in $\mathcal{E}_{\ell}$; this is justified in $\mathcal{E}_{\ell}$ by the localization of the sectrum of $\mathcal{L}_{\eta}$ in Proposition 7.12.

Consider again the decompositions $L_{\eta}=A+B_{\eta}$ and $\mathcal{L}_{\eta}=\mathcal{A}+\mathcal{B}_{\eta}$ with

$$
\left\{\begin{array}{l}
A f:=\bar{A} f \text { and } B_{\eta} f:=\bar{B} f-v \cdot \nabla_{x} f+\eta \Delta_{x} f \\
\mathcal{A} f:=\overline{\mathcal{A}} f \text { and } \mathcal{B}_{\eta} f:=\overline{\mathcal{B}} f-v \cdot \nabla_{x} f+\eta \Delta_{x} f
\end{array}\right.
$$

We know that for $\lambda_{1} \in(0, \lambda)$, the resolvent $\mathcal{R}_{\eta}(\xi)$ of $\mathcal{L}_{\eta}$ with $\Re e \xi=\lambda_{1}$ is welldefined in $\mathcal{E}_{\ell}$, with the representation (7.9). Moreover we have seen in the proof of Proposition 7.12 that $\mathcal{A}$ is bounded from $\mathcal{E}_{\ell}$ into $E$ and $\mathcal{B}_{\eta}-\xi$ is dissipative for $\Re e \xi=\lambda_{1}$ uniformly in terms of $\eta$. Therefore we can write the factorization formula as in Theorem 2.1:

$$
\mathcal{R}_{\eta}(\xi)=-\left(\mathcal{B}_{\eta}-\xi\right)^{-1}+R_{\eta}(\xi) \mathcal{A}\left(\mathcal{B}_{\eta}-\xi\right)^{-1}
$$

where $R_{\eta}(\xi)$ is the resolvent of $L_{\eta}$ in $E$.
We now estimate $I_{M, \eta}$ using that

$$
\begin{aligned}
\left\|\mathcal{R}_{\eta}(\xi)\right\|_{\mathcal{E}_{\ell} \rightarrow \mathcal{E}_{0}} \leq & \left\|\left(\mathcal{B}_{\eta}-\xi\right)^{-1}\right\|_{\mathcal{E}_{\ell} \rightarrow \mathcal{E}_{\ell}} \\
& +\left\|R_{\eta}(\xi)\right\|_{E \rightarrow \mathcal{E}_{0}}\|\mathcal{A}\|_{\mathcal{E}_{\ell-2} \rightarrow E}\left\|\left(\mathcal{B}_{\eta}-\xi\right)^{-1}\right\|_{\mathcal{E}_{\ell} \rightarrow \mathcal{E}_{\ell-2}}
\end{aligned}
$$

where we have used that the norm of $\mathcal{E}_{0}$ is controlled by the norm of $E$ and the fact that for because $g \in \mathcal{E}_{\ell-2}$ we have $\mathcal{A} g \in W_{x}^{\ell-2,1}\left(H_{v}^{1}\right) \subset H_{x}^{[\ell-d / 2]}\left(H_{v}^{1}\right) \subset E$ by $\ell-d / 2>1$.

Finally by using

$$
\left\|\left(\mathcal{B}_{\eta}-\xi\right) f\right\|_{\mathcal{E}_{\ell-2}} \leq \frac{C}{|\Im m \xi|^{2}}\left\|\langle v\rangle^{2} f\right\|_{\mathcal{E}_{\ell}}
$$

(loss of twice the graph norm) for some constant $C$ independent of $\eta$, we deduce that for $\ell>d / 2+3$ :

$$
\left\|\mathcal{R}_{\eta}(\xi) f\right\|_{\mathcal{E}_{0}} \leq \frac{C}{|\Im m \xi|^{2}}\left\|\langle v\rangle^{2} f\right\|_{\mathcal{E}_{\ell}}
$$

for some constant $C$ independent of $\eta$ and therefore we obtain the decay of the semigroup of $\mathcal{L}_{\eta}$ by the representation formula (7.9).

Finally it is clear (from a Gronwall estimate for instance) that the semigroup of $\mathcal{L}_{\eta}$ pointwise converges to the semigroup of $\mathcal{L}$ in $\mathcal{E}_{\ell}$ as $\eta \rightarrow 0$. Therefore we can let the parameter $\eta$ goes to 0 since the estimate does not depend on $\eta$, and we obtain the desired estimate for the semigroup of $L$, which concludes the proof.
7.7. Relaxation for the nonlinear Boltzmann equation. Let us first give a theorem combining existing results:

Theorem 7.15 ([12]). Let us $\left(f_{t}\right)_{t \geq 0}$ be a nonnegative nonzero smooth solution of (17.4) such that for $k, s$ big enough

$$
\sup _{t \geq 0}\left(\|f\|_{H^{k}\left(\mathbb{T}^{d} \times \mathbb{R}^{d}\right)}+\|f\|_{L^{1}\left(1+|v|^{s}\right)}\right) \leq C<+\infty
$$

Then ([21, Theorem 1.1]) for some $\tau, K, A, q>0$

$$
\inf _{t \geq \tau, x \in \mathbb{T}^{d}} f_{t}(x, v) \geq K e^{A|v|^{s}}
$$

and there exists an explicit function $\varphi=\varphi(t)$ which goes to 0 as $t$ goes to infinity such that

$$
\forall t \geq 0, \quad\left\|f_{t}-M\right\|_{W_{x}^{k, 1}\left(L^{1}\left(e^{a|v| s}\right)\right)} \leq \varphi(t)
$$

where $M=\Pi f$ is the Gaussian equilibrium associated with $f$.
Proof of Theorem [21]. This theorem is an immediate consequence of [12, Theorem 2] together with lower bounds proved in [21, Theorem 1.1] and the proof that the moments $L^{1}\left(e^{a|v|^{s}}\right), s \in(0,1)$ appears and are then uniformly bounded in [22, Lemma 4.7] and [20, Proposition 3.2] (in the latter paper the proof was done in the homogeneous case, however one can integrate in $x$ these moment estimates by using the uniform bounds in $x$ from above and below on $f$ ).

Then we write the nonlinear equation for the fluctuation $f=M+h$ as

$$
\partial_{t} h=\mathcal{L} h+Q(h, h)
$$

and we use that (for some $k$ big enough in terms of the dimension)

$$
\|Q(h, h)\|_{W_{x}^{k, 1}\left(L^{1}\left(e^{a|v|^{s}}\right)\right)} \leq C\|h\|_{W_{x}^{k, 1}\left(L^{1}\left(e^{a|v|^{s}}\right)\right)}\left\|h\left(1+|v|^{\gamma}\right)\right\|_{W_{x}^{k, 1}\left(L^{1}\left(e^{a|v|^{s}}\right)\right)}
$$

The latter inequality is a consequence of the $L^{1}$ theory:

$$
\|Q(h, h)\|_{L^{1}\left(e^{a|v|} \mid\right.} \leq C\|h\|_{L^{1}\left(e^{a|v|^{s}}\right)}\left\|h\left(1+|v|^{\gamma}\right)\right\|_{L^{1}\left(\left.e^{a|v|}\right|^{s}\right)}
$$

(see for instance [20]) together with the Leibniz formula for the derivatives in $x$, and Sobolev embeddings.

Finally by elementary Gronwall estimates we obtain the following theorem:
Theorem 7.16. Let us $\left(f_{t}\right)_{t \geq 0}$ be a nonnegative nonzero smooth solution of (7.4) such that for $k, s$ big enough

$$
\sup _{t \geq 0}\left(\|f\|_{H^{k}\left(\mathbb{T}^{d} \times \mathbb{R}^{d}\right)}+\|f\|_{L^{1}\left(1+|v|^{s}\right)}\right) \leq C<+\infty .
$$

Then for some $C>0$ and $\lambda_{1} \in(0, \lambda)$ (where $\lambda$ is given by the rate of decay of the linearized semigroup in $W_{x}^{k, 1}\left(L^{1}\left(e^{a|v|^{s}}\right)\right.$ ))

$$
\forall t \geq 0, \quad\left\|f_{t}-M\right\|_{W_{x}^{k, 1}\left(L^{1}\left(e^{a|v| s}\right)\right)} \leq C e^{-\lambda t}
$$

where $M=\Pi f$ is the Gaussian equilibrium associated with $f$.
In particular we deduce a proof of the exponential decay of the relative entropy:

$$
\forall t \geq 0, \quad \int f_{t} \log \frac{f_{t}}{M} d x d v \leq C e^{-\lambda t}
$$

7.8. Other applications. Here are some natural questions left open by this study:

- The extension of Theorem 7.16 to the Boltzmann equation for non-cutoff hard and moderately soft potentials (in the case of long-range interactions)
- The extension of Theorem 7.16 in the case of a confining potential as in Theorem 5.8 for the kinetic Fokker-Planck equation.
Moreover two other interesting applications of our new estimates should be looked in the futur:
- To develop a complete theory of nonlinear stability around the Maxwellian equilibrium in $W_{x}^{k, 1}\left(L^{1}\left(e^{a|v|^{s}}\right)\right)$.
- To develop a quantitative close-to-homogeneous theory for the Boltzmann equation, improving on the non-constructive results [1]. This is allowed by our estimates since the functional space of the nonlinear stability (1) is compatible with the nonlinear flow of the spatially homogeneous Boltzmann equation.


## References

[1] Arkeryd, L., Esposito, R., and Pulvirenti, M. The Boltzmann equation for weakly inhomogeneous data. Comm. Math. Phys. 111, 3 (1987), 393-407.
[2] Arnold, A., Gamba, I., Gualdani, M. P., Mischler, S., Mouthot, C., and Sparber, C. The wigner-fokker-planck equation: stationary state and large time behavior. In progress.
[3] Arnold, A., Markowich, P., Toscani, G., and Unterreiter, A. On convex Sobolev inequalities and the rate of convergence to equilibrium for Fokker-Planck type equations. Comm. Partial Differential Equations 26, 1-2 (2001), 43-100.
[4] Bakry, D., and Émery, M. Propaganda for $\Gamma_{2}$. In From local times to global geometry, control and physics (Coventry, 1984/85), vol. 150 of Pitman Res. Notes Math. Ser. Longman Sci. Tech., Harlow, 1986, pp. 39-46.
[5] Baranger, C., and Mouhot, C. Explicit spectral gap estimates for the linearized Boltzmann and Landau operators with hard potentials. Rev. Mat. Iberoamericana 21, 3 (2005), 819-841.
[6] Bobylev, A. V., and Cercignani, C. On the rate of entropy production for the Boltzmann equation. J. Statist. Phys. 94, 3-4 (1999), 603-618.
[7] Carleman, T. Problèmes mathématiques dans la théorie cinétique des gaz. Publ. Sci. Inst. Mittag-Leffler. 2. Almqvist \& Wiksells Boktryckeri Ab, Uppsala, 1957.
[8] Carlen, E. A., and Carvalho, M. C. Strict entropy production bounds and stability of the rate of convergence to equilibrium for the Boltzmann equation. J. Statist. Phys. 67, 3-4 (1992), 575-608.
[9] Carlen, E. A., and Carvalho, M. C. Entropy production estimates for Boltzmann equations with physically realistic collision kernels. J. Statist. Phys. 74, 3-4 (1994), 743-782.
[10] Cercignani, C. $H$-theorem and trend to equilibrium in the kinetic theory of gases. Arch. Mech. (Arch. Mech. Stos.) 34, 3 (1982), 231-241 (1983).
[11] Cercignani, C. The Boltzmann equation and its applications, vol. 67 of Applied Mathematical Sciences. Springer-Verlag, New York, 1988.
[12] Desvillettes, L., and Villani, C. On the trend to global equilibrium for spatially inhomogeneous kinetic systems: the Boltzmann equation. Invent. Math. 159, 2 (2005), 245-316.
[13] Dolbeault, J., Mouhot, C., and Schmeiser, C. Hypocoercivity for kinetic equations conserving mass. ArXiv eprint 1005.1495.
[14] Dolbeault, J., Mouhot, C., and Schmeiser, C. Hypocoercivity for kinetic equations with linear relaxation terms. C. R. Math. Acad. Sci. Paris 347, 9-10 (2009), 511-516.
[15] Grad, H. Principles of the kinetic theory of gases. In Handbuch der Physik (herausgegeben von S. Flügge), Bd. 12, Thermodynamik der Gase. Springer-Verlag, Berlin, 1958, pp. 205-294.
[16] Grad, H. Asymptotic theory of the Boltzmann equation. II. In Rarefied Gas Dynamics (Proc. 3rd Internat. Sympos., Palais de l'UNESCO, Paris, 1962), Vol. I. Academic Press, New York, 1963, pp. 26-59.
[17] Helffer, B., and Sjöstrand, J. From resolvent bounds to semigroup bounds. Tech. rep., ArXiv:1001.4171, 2010.
[18] Kato, T. Perturbation theory for linear operators. Classics in Mathematics. Springer-Verlag, Berlin, 1995. Reprint of the 1980 edition.
[19] Mischler, S., and Mouhot, C. Stability, convergence to self-similarity and elastic limit for the Boltzmann equation for inelastic hard spheres. Comm. Math. Phys. 288, 2 (2009), 431-502.
[20] Mischler, S., Mouhot, C., and Rodriguez Ricard, M. Cooling process for inelastic Boltzmann equations for hard spheres. I. The Cauchy problem. J. Stat. Phys. 124, 2-4 (2006), 655-702.
[21] Mouhot, C. Quantitative lower bounds for the full Boltzmann equation. I. Periodic boundary conditions. Comm. Partial Differential Equations 30, 4-6 (2005), 881-917.
[22] Mouhot, C. Rate of convergence to equilibrium for the spatially homogeneous Boltzmann equation with hard potentials. Comm. Math. Phys. 261, 3 (2006), 629-672.
[23] Mouhot, C. Enlarging the functional space of decay estimates on semigroups. ArXiv eprint 0912.1747.
[24] Mouhot, C., and Neumann, L. Quantitative perturbative study of convergence to equilibrium for collisional kinetic models in the torus. Nonlinearity 19, 4 (2006), 969-998.
[25] Mouhot, C., Russ, E., and Sire, Y. Fractional Poincaré inequality for general measures. To appear in J. Math. Pures Appl., arXiv eprint 0911.4563.
[26] Pazy, A. Semigroups of linear operators and applications to partial differential equations, vol. 44 of Applied Mathematical Sciences. Springer-Verlag, New York, 1983.
[27] Rezakhanlou, F., and Villani, C. Entropy methods for the Boltzmann equation, vol. 1916 of Lecture Notes in Mathematics. Springer, Berlin, 2008. Lectures from a Special Semester on Hydrodynamic Limits held at the Université de Paris VI, Paris, 2001, Edited by François Golse and Stefano Olla.
[28] Rothaus, O. Logarithmic Sobolev inequalities and the spectrum of Schrödinger operators. J. Funct. Anal. 42 (1981), 110-120.
[29] Toscani, G., and Villani, C. Sharp entropy dissipation bounds and explicit rate of trend to equilibrium for the spatially homogeneous Boltzmann equation. Comm. Math. Phys. 203, 3 (1999), 667-706.
[30] Villani, C. Cercignani's conjecture is sometimes true and always almost true. Comm. Math. Phys. 234, 3 (2003), 455-490.
[31] YaO, P. F. On the inversion of the Laplace transform of $C_{0}$ semigroups and its applications. SIAM J. Math. Anal. 26, 5 (1995), 1331-1341.

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