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# Lower exponential strong well-posedness of hyperbolic boundary value problems in a strip 

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#### Abstract

In this article we are interested in the lower exponential strong well-posedness of hyperbolic boundary value problems set in the strip $\mathbb{R}^{d-1} \times[0,1]$. We assume that each component of the boundary of the strip satisfies the so-called uniform Kreiss-Lopatinskii condition (which is the condition ensuring the lower exponential strong well-posedness of the boundary value problem in the half space) and we show that due to selfinteraction phenomenon, an extra condition has to be made to ensure the lower exponential strong well-posedness of the boundary value problem in the strip. This condition imposes that a selfinteracting wave in the strip can not be amplified by repetitive reflections against each component of the boundary. This condition is very analogous to the one imposed in Osher, 1973 to study the problem in the quarter space. However, due to the simplicity of the phases generation process for the strip geometry (less phases are generated) and this condition involves matrices and not Fourier integral operators.


## AMS subject classification :35L50

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## 1 Introduction

In this article we are interested in the study of hyperbolic boundary value problems set in the strip $\mathbb{R}^{d-1} \times$ $[0,1]$. Such problems read:

$$
\begin{cases}L(\partial) u:=\partial_{t} u+\sum_{j=1}^{d} A_{j} \partial_{j} u=f & \text { for } \left.\left(t, x^{\prime}, x_{d}\right) \in \mathbb{R}_{t} \times \mathbb{R}^{d-1} \times\right] 0,1[  \tag{1}\\ B_{0} u_{\mid x_{d}=0}=g_{0} & \text { for }\left(t, x^{\prime}\right) \in \mathbb{R}_{t} \times \mathbb{R}^{d-1} \\ B_{1} u_{\mid x_{d}=1}=g_{1} & \text { for }\left(t, x^{\prime}\right) \in \mathbb{R}_{t} \times \mathbb{R}^{d-1} \\ u_{\mid t \leq 0}=0 & \text { for }\left(x^{\prime}, x_{d}\right) \in \mathbb{R}^{d-1} \times[0,1]\end{cases}
$$

where the coefficients $A_{j}$ are given matrices in $\mathbf{M}_{N \times N}(\mathbb{R})$, the source terms $f, g_{0}$ and $g_{1}$ are given functions and where the solution $u$ has its values in $\mathbb{R}^{N}$.

The boundary matrices in (1) namely $B_{0}$ and $B_{1}$ are generic matrices encoding the good number of boundary conditions to make sure that the problem $\sqrt[11]{ }$ is not underdetermined or at the contrary overdetermined. More precisely $B_{0} \in \mathbf{M}_{p \times N}(\mathbb{R})$ and $B_{1} \in \mathbf{M}_{(N-p) \times N}(\mathbb{R})$, the value of $p$ will be made precise in Assumption 2.2 .

Our aim in this article is to establish a full characterization of the lower exponential strong well-posedness of (1) in terms of the boundary conditions $B_{0}$ and $B_{1}$. Firstly, let us give more details about what we mean by lower exponential strong well-posedness by considering the analogous of 11 in the half space geometry, namely

$$
\begin{cases}L(\partial) u=f & \text { for }\left(t, x^{\prime}, x_{d}\right) \in \mathbb{R}_{t} \times \mathbb{R}^{d-1} \times \mathbb{R}_{+}  \tag{2}\\ B_{0} u_{\mid x_{d}=0}=g & \text { for }\left(t, x^{\prime}\right) \in \mathbb{R}_{t} \times \mathbb{R}^{d-1} \\ u_{\mid t \leq 0}=0 & \text { for }\left(x^{\prime}, x_{d}\right) \in \mathbb{R}^{d-1} \times \mathbb{R}_{+}\end{cases}
$$

The boundary value problem (2) has been deeply studied since the seminal work of Kreiss, 1970 and in particular we know that the strong well-posedness of $(2)$ is equivalent to the fullfillment of the so called uniform Kreiss-Lopatinskii condition. This condition states that in the normal mode analysis no stable mode is solution of the homogeneous boundary value problem (we refer to Kreiss, 1970 for more details).

By strong well-posedness we mean that for all source terms $f \in L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times \mathbb{R}_{+}\right)$and $g \in L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)$ the boundary value problem (2) admits a unique solution $u \in L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times \mathbb{R}_{+}\right)$with trace $u_{\mid x_{d}=0} \in$ $L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)$ satisfying the following energy estimate: there exists $C>0$ and $\gamma_{0} \geq 0$ such that for all $\gamma>\gamma_{0}$ we have the inequality:

$$
\begin{equation*}
\gamma\|u\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times \mathbb{R}_{+}\right)}^{2}+\left\|u_{\mid x_{d}=0}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2} \leq C\left(\frac{1}{\gamma}\|f\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times \mathbb{R}_{+}\right)}^{2}+\|g\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2}\right) \tag{3}
\end{equation*}
$$

where for some Banach space $X$ we defined $L_{\gamma}^{2}(X):=\left\{u \in \mathscr{D}^{\prime}(X) \backslash u e^{-\gamma t} \in L^{2}(X)\right\}$ (equipped with the obvious norm).

The fact of interest here is that the estimate (3) permits a solution with exponential growth in time (bounded by $e^{\gamma_{0} t}$ ). However it has also been shown in Kreiss, 1970 that in fact one can take $\gamma_{0}$ to be zero and consequently the solution $u$ does not have exponential growth in time. It is what we mean by lower exponential strong well-posedness.

We now go back to the problem in the strip namely (1) and derive some consequences of the result of Kreiss, 1970. As the operator $L(\partial)$ is assumed to be hyperbolic we have finite speed of propagation so that it is clear (by localization) that impose the uniform Kreiss-Lopatinskii condition on each side of the boundary is a necessary condition for the strong well-posedness of (1) (the definition is the same as for the half space problem except that we now want to control the two traces, namely $u_{\mid x_{d}=0}$ and $\left.u_{\mid x_{d}=1}\right)$. Moreover the fact that $L(\partial)$ is stable by zero order perturbations directly implies that impose the uniform Kreiss-Lopatinskii condition on each side of the boundary define a strongly well-posed problem.

Let us briefly sketch the proof of the a priori energy estimate in such a situation. Let $\chi_{0}, \chi_{1} \in \mathscr{D}([0,1])$ be such that:

$$
\chi_{0}\left(x_{d}\right)=\left\{\begin{array}{ll}
1 & \text { for } x_{d} \in\left[0, \frac{1}{4}\right] \\
0 & \text { for } x_{d} \geq \frac{1}{2}
\end{array} \text { and } \chi_{1}\left(x_{d}\right)= \begin{cases}1 & \text { for } x_{d} \in\left[\frac{3}{4}, 1\right] \\
0 & \text { for } x_{d} \leq \frac{1}{2}\end{cases}\right.
$$

and define $\chi_{\frac{1}{2}}=1-\chi_{0}-\chi_{1}$. Let $u$ be a regular solution of (1) and define $u_{0}:=\chi_{0} u, u_{\frac{1}{2}}:=\chi_{\frac{1}{2}} u$ and $u_{1}:=\chi_{1} u$. Clearly, $u=u_{0}+u_{\frac{1}{2}}+u_{1}, u_{\mid x_{d}=j}=u_{j_{\mid x_{d}=j}}$ for $j=0,1$. By hyperbolicity the problem in the full space is strongly well-posed (and even lower exponentially strongly well-posed) so that there exists $C>0$ such that $\gamma>0$

$$
\begin{aligned}
\gamma\left\|u_{\frac{1}{2}}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2} & \leq \frac{C}{\gamma}\left\|L(\partial) u_{\frac{1}{2}}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{T} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2} \\
& \leq \frac{C}{\gamma}\|L(\partial) u\|_{L_{\gamma}^{2}\left(\mathbb{R}_{T} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2}+\frac{C_{A_{d}}}{\gamma}\|u\|_{L_{\gamma}^{2}\left(\mathbb{R}_{T} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2}
\end{aligned}
$$

by standard estimate on the commutator $\left[L(\partial), \varphi_{\frac{1}{2}}\right]$. Consequently, taking $\gamma$ large enough, we obtain

$$
\begin{equation*}
\gamma\left\|u_{\frac{1}{2}}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2} \leq \frac{C}{\gamma}\|f\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2} \tag{4}
\end{equation*}
$$

By the Kreiss-Lopatinskii condition on each side, the same reasoning gives (for $\gamma$ large enough)

$$
\begin{equation*}
\gamma\left\|u_{0}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2}+\left\|u_{0_{\mid x_{d}=0}}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2} \leq C\left(\frac{1}{\gamma}\|f\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2}+\left\|g_{0}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2}\right) \tag{5}
\end{equation*}
$$

and

$$
\begin{equation*}
\gamma\left\|u_{1}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2}+\left\|u_{1_{\mid x_{d}=1}}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2} \leq C\left(\frac{1}{\gamma}\|f\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2}+\left\|g_{1}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2}\right) \tag{6}
\end{equation*}
$$

Sum (4)-(5) and (6) gives the desired result that is to say: that there exist $C>0$ and $\gamma_{0} \geq 0$ such that for all $\gamma>\gamma_{0}$ we have the following a priori energy estimate for (1):

$$
\begin{align*}
\gamma\|u\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2} & +\left\|u_{\mid x_{d}=0}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2}+\left\|u_{\mid x_{d}=1}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2}  \tag{7}\\
& \leq C\left(\frac{1}{\gamma}\|f\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1} \times[0,1]\right)}^{2}+\left\|g_{0}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2}+\left\|g_{1}\right\|_{L_{\gamma}^{2}\left(\mathbb{R}_{t} \times \mathbb{R}^{d-1}\right)}^{2}\right)
\end{align*}
$$

so that the problem in the strip (1) is strongly well-posed. The rest of this paper adress the question of the lower exponential strong well-posedness of (1) that is, crudely speaking, "does (7) holds with $\gamma_{0}=0$ ?"

Before we turn to the description of our main result let us motivate a bit this study. Firstly hyperbolic boundary value problems in the strip are, in the author's knowledge, very little considered in the litterature. The reason of that is probably that, as we have just shown, they are easily solvable by localization. However the localization technique is far to be entirely satisfactory because the maximal time exponential growth given in (7) is a crude one (especially if there is no time exponential growth at all). So a full characterization of the lower exponential strong well-posedness is a first substancial refinement.

More generally, hyperbolic boundary value problems with more than one boundary are also not a lot considered in the litterature. For example, in the quarter space geometry, we refere to the seminal work of Osher, 1973-Osher, 1974b (see also Benoit, 2015 and the references therein), the strong well-posedness of the problem is far to be clearly characterized. As the reader will notice, many of the ideas described in this paper are adapted for the ones use in the quarter space geometry. So the author believes that a good understanding of the hyperbolic boundary value problem in the (simpler) geometry of the strip will improve the understanding of the one in the quarter space.

An other interest of this study is linked to the study of the stability of finite difference scheme approximations. Indeed when one implements a numerical scheme to approximate the solution of the Cauchy problem associated to (1) then due to the impossibility to modelized the full space $\mathbb{R}^{d}$ artificial boundary conditions have to be imposed numerically. Consequently the scheme will be defined on some (possibly big) rectangle.

Let us for simplicity only truncate the space $\mathbb{R}^{d}$ with respect to one variable, to fix the ideas $x_{d}$, then the associated scheme is defined in a strip and it is interesting to determine the boundary conditions leading to the stability of the scheme (without entering into details, let us say that some discretized version of (7) holds (we refer to Coulombel, 2011b or B. Gustafsson and Sundstrom., 1972 for precise estimates)).

The seminal result of B. Gustafsson and Sundstrom., 1972 ensures that if each boundary condition satisfies some discretized version of the uniform Kreiss-Lopatinskii condition, then the associated scheme is strongly stable (that is that the analogous of (7) holds with some treshold for $\gamma_{0}$ ). The question of an estimate like (7) with $\gamma_{0}=0$, refered as the $\bar{P}$-stability ( $P$ stands for pratical), is also considered in Trefethen, 1985 where the author shows that two stable boundary conditions are not sufficient to ensure the $P$-stability of the associated scheme defined in the strip. In the author's opinion a full characterization of the numerical boundary condition leading to $P$-stability is an interesting question. Indeed for numerical schemes, compared to the analogous continous problem, all the components of the traces have to be specified so that some extra (and arbitrary, in the sens that they are not prescribe by the physics of the problem) boundary conditions have to be imposed. Having a characterization allowing to determine if these extra boundary conditions induce an exponential growth of the solution, or not, is thus a natural question that we expect to clarify in a forthcoming contribution.

Unsurprisingly impose the uniform Kreiss-Lopatinskii condition on each side of the boundary will not be sufficient to ensure lower exponential strong well-posedness. Intuitively and in a formal way the reason is the following. Consider the strip problem (1) with $f \equiv 0$ and $g_{1} \equiv 0$ and assume that the source term $g_{0}$ induces some information traveling to the side $\left\{x_{d}=1\right\}$ (think, for example, to a transport phenomenon) then this information will hit the side $\left\{x_{d}=1\right\}$ after a finite time and be reflected back (by the uniform Kreiss-Lopatinskii condition) with some amplification reflection coefficient $\alpha_{1}$. The reflected information travels to the side $\left\{x_{d}=0\right\}$, it hits it in finite time, and is reflected again with an amplification reflection coefficient $\alpha_{0}$. Now imagine that this reflected information is (or contains) the initial one (this kind of phenomenon was refered as selfinteraction in Benoit, 2016), then this process will be repeated periodically in time and one expected that the solution $u$ behaves more or less like $\left(\alpha_{0} \alpha_{1}\right)^{t}$. Consequently if $\left|\alpha_{0} \alpha_{1}\right|>1$ (that is to say the information is amplified during the two reflections) then $u$ is expected to have a non trivial exponential growth in time and (1) can not be lower exponentially strongly well-posed. As the uniform Kreiss-Lopatinskii condition implies the existence of reflected rays but does not say anything about the associated reflection coefficient we deduce that impose this condition on each side of the boundary is not sufficient for lower exponential strong well-posedness.

Consequently two new inversibility conditions are necessary and (also) sufficient to ensure the lower exponential strong well-posedness. These conditions has to be understood as a non amplification condition for selfinteracting modes and thus has to be compared with the condition described in Osher, 1973 for the hyperbolic corner problem. The first condition ensures the existence of a weak solution while the second one is used to establish the a priori energy estimate so that it gives the uniqueness of the solution.

However the results of Proposition 3.6 and 3.13 indicate that these two conditions seem to be equivalent (we are not so far for the proof of such a result see Proposition 3.6 and 3.13 ) so that in the rest of the introduction we will only describe the first inversibility condition.

More precisely this new condition in this article involves two operators $T_{0 \rightarrow 1}(\zeta)$ (resp. $T_{1 \rightarrow 0}(\zeta)$ ), where $\zeta$ stands for a frequency parameter, that to the value of the trace of the solution on the side $\left\{x_{d}=0\right\}$ (resp. $\left\{x_{d}=1\right\}$ ) gives in output the value of the trace of the solution on the side $\left\{x_{d}=1\right\}$ (resp. $\left\{x_{d}=0\right\}$ ). The new necessary and sufficient condition for lower exponential strong well-posedness of the hyperbolic boundary value problem in the strip then means that the operator $\left(I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right)$ is uniformly invertible, on some subspace, in terms of the frequency parameter $\zeta$. The operator $T_{0 \rightarrow 1}(\zeta)$ (resp. $T_{1 \rightarrow 0}(\zeta)$ ) encodes the amplification factor during the reflection of an incoming wave for the side $\left\{x_{d}=1\right\}$ (resp. $\left\{x_{d}=0\right\}$ ) against this side. Consequently to ensure that $\left(I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right)$ it is sufficient to impose that $\left|T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right|<1$, which means the natural condition that to ensure the lower exponential strong well-posedness it is necessary, to prevent some kind of blow-up in time, to impose that the energy amplification of a wave descreases when it comes from an information on the side $\left\{x_{d}=0\right\}$, hits $\left\{x_{d}=1\right\}$ and finally goes back to $\left\{x_{d}=0\right\}$ as described before.

This phenomenon was already present in the geometry of the corner for which some phases can regenerate themselves after a suitable number of reflections against the sides of the corner. However in the geometry of the corner, this phenomenon is a bit hidden in some really specific and probably not so common geometries of the characteristic variety (see Benoit, 2016 for more details).

Whereas for the problem in the strip, this phenomenon becomes generic. Indeed as soon as the side $\left\{x_{d}=0\right\}$ generates incoming waves, they will travel to $\left\{x_{d}=1\right\}$ and be reflected back to $\left\{x_{d}=0\right\}$ and consequently the initial considered wave will be regenerated after two rebounds. The author excepts to give more details about this assertion in a forthcoming contribution dealing with WKB expansions for the strip problem.

The paper is organized as follows: in Section 2 we introduce some notations and definitions and state the main result, that is to say the full characterization of lower exponentially strong well-posed problems.

Section 3 is devoted to the proof of this result. Firstly we give a proof in the more favorable framework of strictly dissipatives boundary conditions (see Paragraph 3.1) and then we use this result to deal with the general framework in Paragraphs $3.2 \mid 3.3$.

The proof of the a priori energy estimate that we choose to follow here is in the spirit of the one given by Osher, 1973 to treat the case of the corner problem and uses symmetrizors as for the half space Kreiss, 1970. However, as the reader will notice the geometry of the characteristic variety induced by the strip problem will lead to really simple operators $T_{0 \rightarrow 1}(\zeta)$ (resp. $T_{1 \rightarrow 0}(\zeta)$ ). These operators involve matrices and not integral Fourier operators as the operators used in Osher, 1973. As a matter of fact, we are able to complete the discussion and obtain an a priori energy estimate without loss of derivatives which is not the case in Osher, 1973.

Existence of the solution can be obtained by two distinct ways, the first one is an explicit construction under a stronger version of the inversibility condition and the second one uses a standard duality argument. Finally uniqueness follows from a "weak=strong" lemma which is essentially trivial for the strip geometry. At the end of Section 3, more precisely in Paragraph 3.4, we discuss and illustrate the necessary and sufficient condition for lower exponential strong well-posedness on some examples.

At last Section 4 contains a conclusion, a summary and gives some conjectures that seem to be reasonable.

## 2 Notations and main results

### 2.1 Notations and assumptions

Througouht this paper, the strip will be denoted by $\Gamma:=\mathbb{R}^{d-1} \times[0,1]$, the two sides of its boundary will be denoted by $\partial \Gamma_{0}:=\mathbb{R}^{d-1} \times\{0\}$ and $\partial \Gamma_{1}:=\mathbb{R}^{d-1} \times\{1\}$.

For $\chi, s \in \mathbb{R}$ and for some Banach space $X$ including the time variable $t$, we introduce the weighted $H^{s}$-weighted (in time) Sobolev spaces spaces by:

$$
H_{\chi}^{s}(X):=\left\{u \in \mathscr{D}^{\prime}(X) \backslash u e^{-\chi t} \in H^{s}(X)\right\}
$$

equipped with the norm $\|\cdot\|_{H_{\chi}^{s}(X)}:=\left\|\cdot e^{-\chi t}\right\|_{H^{s}(X)}$. We have $\left(H_{-\chi}^{-s}(X)\right)^{*}=H_{\chi}^{s}(X)$.
We will make the following assumptions about the system (1). The first one is a hyperbolicity assumption while the second one restricts the analysis to non-characteristic boundaries. More precisely:

Assumption 2.1 (Constantly hyperbolic operator) The system (11) is constantly hyperbolic in the sense that there exists $q \geq 1$, real valued analytic functions $\lambda_{1}, \ldots, \lambda_{q}$ on $\mathbb{R}^{d} \backslash\{0\}$ and positive integers $\mu_{1}, \ldots, \mu_{q}$ such that:

$$
\forall \xi \in \mathbb{S}^{d-1}, \operatorname{det}\left(\tau+\sum_{j=1}^{d} \xi_{j} A_{j}\right)=\prod_{j=1}^{q}\left(\tau+\lambda_{j}(\xi)\right)^{\mu_{j}}
$$

with $\lambda_{1}(\xi)<\cdots<\lambda_{q}(\xi)$ and the eigenvalues $\lambda_{j}(\xi)$ of $\sum_{j=1}^{d} \xi_{j} A_{j}$ are semi-simple.
Assumption 2.2 (Non-characteristic boundaries) The matrix $A_{d}$ is invertible. We denote by $p$ its number of strictly positive eigenvalues.

The matrices $B_{0} \in \mathbf{M}_{p \times N}(\mathbb{R})$ and $B_{1} \in \mathbf{M}_{(N-p) \times N}(\mathbb{R})$ are of maximal rank.

For technical reasons that will be made precise in the core of the proofs we will in addition to the hyperbolicity Assumption 2.1 assume that the coefficients of (1) are symmetric

Assumption 2.3 (Symmetric operator) The matrices $A_{j}, j=1, \ldots, d$ are symmetric.
As well as the following assumption
Assumption 2.4 We assume that

$$
\operatorname{ker} B_{0} \cap \operatorname{ker} B_{1}=\{0\}
$$

Combined with Assumption 2.2 it immediately implies that we have the following decomposition:

$$
\operatorname{ker} B_{0} \oplus \operatorname{ker} B_{1}=\mathbb{R}^{N}
$$

We introduce the following partition of the frequency space:

$$
\Xi:=\left\{\zeta:=(\sigma=\gamma+i \tau, \eta) \in \mathbb{C} \times \mathbb{R}^{d-1} \backslash \gamma \geq 0\right\} \text { and } \Xi_{0}:=\Xi \cap\{\gamma=0\}
$$

After Laplace transform in time $(t \nsim \leadsto \sigma)$ and Fourier transform in the tangential space ( $\left.x^{\prime} \nrightarrow \eta\right)$, denoted by $\widehat{\cdot}$, in the hyperbolic boundary value problem (1) we obtain the following ordinary differential equation in the normal variable $x_{d}$ :

$$
\left\{\begin{array}{l}
\left.\frac{\mathrm{d}}{\mathrm{~d} x_{d}} \widehat{u}\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) \widehat{u}\left(\zeta, x_{d}\right)+A_{d}^{-1} \widehat{f}\left(\zeta, x_{d}\right), \quad \text { for } x_{d} \in\right] 0,1[  \tag{8}\\
B_{0} \widehat{u}_{\mid x_{d}=0}(\zeta)=\widehat{g}_{0}(\zeta) \\
B_{1} \widehat{u}_{\mid x_{d}=1}(\zeta)=\widehat{g}_{1}(\zeta)
\end{array}\right.
$$

where $\mathscr{A}(\zeta)$ is the so-called resolvent matrix defined by:

$$
\begin{equation*}
\mathscr{A}(\zeta):=-A_{d}^{-1}\left(\sigma I+i \sum_{j=1}^{d-1} \eta_{j} A_{j}\right) \tag{9}
\end{equation*}
$$

and we introduce

$$
\begin{equation*}
M(\zeta):=\left(\sigma I+i \sum_{j=1}^{d-1} \eta_{j} A_{j}\right) \tag{10}
\end{equation*}
$$

A direct consequence of Assumption 2.1 is the following lemma due to Hersh, 1963:
Lemma 2.1 Hersh, 1963]For all frequency parameter $\zeta \in \Xi \backslash \Xi_{0}$, the resolvent matrix $\mathscr{A}(\zeta)$ only admits eigenvalues with non-zero real part, and thus do not has purely imaginary eigenvalues. We denote by $E^{s}(\zeta)$ (resp. $\left.E^{u}(\zeta)\right)$, the stable (resp. unstable) space of $\mathscr{A}(\zeta)$ that is the eigenspace associated with the negative (resp. positive) real part eigenvalues. Then independently of $\zeta, \operatorname{dim} E^{s}(\zeta)=p$ and $\operatorname{dim} E^{s}(\zeta)=N-p$ and we have the following decomposition:

$$
\begin{equation*}
\mathbb{C}^{N}=E^{s}(\zeta) \oplus E^{u}(\zeta) \tag{11}
\end{equation*}
$$

We denote by $\Pi^{s}(\zeta)$ (resp. $\Pi^{u}(\zeta)$ ) the projection upon $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ ) with respect to the decomposition (11).

For $\zeta \in \Xi_{0}$, the decomposition (11) does not hold anymore. However it has been shown by Kreiss, 1970 for strictly hyperbolic operators and then generalized by Métivier, 2000 to constanly hyperbolic operators that the stable subspace $E^{s}(\zeta)$ and the unstable subspace $E^{u}(\zeta)$ admit continuous extension up to $\Xi_{0}$. It is a consequence of the following Theorem:

Theorem 2.1 (Block structure) Under Assumptions 2.1.2.2, for all $\underline{\zeta} \in \Xi$, there exists a neighborhood $\mathscr{V}$ of $\underline{\zeta}$ in $\Xi$, an integer $L \geq 1$, a partition $N=\mu_{1}+\cdots+\mu_{L}$, with $\mu_{1}, \ldots, \mu_{L} \geq 1$ and an invertible matrix $T$, regular on $\mathscr{V}$ such that:

$$
\forall \zeta \in \mathscr{V}, T^{-1}(\zeta) \mathscr{A}(\zeta) T(\zeta)=\operatorname{diag}\left(\mathscr{A}_{1}(\zeta), \cdots, \mathscr{A}_{L}(\zeta)\right)
$$

where the blocks $\mathscr{A}_{j}(\zeta) \in \mathbf{M}_{\mu_{j} \times \mu_{j}}(\mathbb{C})$ satisfy one of the following alternatives:
i) all the elements in the spectrum of $\mathscr{A}_{j}(\underline{\zeta})$ have positive real part.
ii) All the elements in the spectrum of $\mathscr{A}_{j}(\underline{\zeta})$ have negative real part.
iii) $\mu_{j}=1, \mathscr{A}_{j}(\underline{\zeta}) \in i \mathbb{R}, \partial_{\gamma} \mathscr{A}_{j}(\underline{\zeta}) \in \mathbb{R} \backslash\{0\}$ and $\mathscr{A}_{j}(\zeta) \in$ iR for all $\zeta \in \Xi_{0} \cap \mathscr{V}$.
iv) $\mu_{j}>1$ and there exists $k_{j} \in i \mathbb{R}$ such that

$$
\mathscr{A}_{j}(\underline{\zeta})=\left[\begin{array}{ccc}
k_{j} & i & 0 \\
& \ddots & i \\
0 & & k_{j}
\end{array}\right]
$$

the coefficient in the lower left corner of $\partial_{\gamma} \mathscr{A}_{j}(\underline{\zeta}) \in \mathbb{R} \backslash\{0\}$ and for all $\zeta \in \Xi_{0} \cap \mathscr{V}, \mathscr{A}_{j}(\zeta) \in i \mathbf{M}_{\mu_{j} \times \mu_{j}}(\mathbb{R})$.
An important remark for what follows is this nice property induced by the geometry of the strip. On the one hand, if one considers the boundary value problem in the half space:
then the subspaces $E_{0}^{s}(\zeta)$ and $E_{0}^{u}(\zeta)$ (extended by continuity to $\Xi_{0}$ ) associated to the problem (12) are given by $E_{0}^{s}(\zeta):=E^{s}(\zeta)$ and $E_{0}^{u}(\zeta):=E^{u}(\zeta)$. But on the other hand if one considers the hyperbolic problem in the half space:

$$
\begin{cases}\frac{\mathrm{d}}{\mathrm{~d} x_{d}} \widehat{u}\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) \widehat{u}\left(\zeta, x_{d}\right)+A_{d}^{-1} \widehat{f}\left(\zeta, x_{d}\right) & \text { for } \left.x_{d} \in\right]-\infty, 1[  \tag{13}\\ B_{1} \widehat{u}_{\mid x_{d}=1}(\zeta)=\widehat{g}_{1}(\zeta) & \end{cases}
$$

and makes the change of variable $\widetilde{x}_{d}=1-x_{d}$, then he recovers the problem:

$$
\left\{\begin{array}{l}
\left.\frac{\mathrm{d}}{\mathrm{~d} x_{d}} \widehat{u}\left(\zeta, x_{d}\right)=-\mathscr{A}(\zeta) \widehat{u}\left(\zeta, x_{d}\right)-A_{d}^{-1} \widehat{f}\left(\zeta, x_{d}\right) \quad \text { for } \widetilde{x}_{d} \in\right] 0,+\infty[  \tag{14}\\
B_{1} \widehat{u} \mid \widetilde{x}_{d}=0 \\
\widehat{g}_{1}(\zeta)=\widehat{g}_{1}(\zeta)
\end{array}\right.
$$

So from (14) Lemma 2.1 and Theorem 2.1 it is easy to see that the stable subspace $E_{1}^{s}(\zeta)$ and the unstable subspace $E_{1}^{u}(\zeta)$ associated for (13) (both extended by continuity to $\Xi_{0}$ ) are respectively given by $E_{1}^{s}(\zeta)=$ $E^{u}(\zeta)$ and $E_{1}^{u}(\zeta)=E^{s}(\zeta)$.

In other words, a stable or (outgoing) information for the side $\partial \Gamma_{0}$ is unstable or (incoming) for the side $\partial \Gamma_{1}$ and vice versa.

With these remarks in hand we can state the definition of the uniform Kreiss-Lopatinskii condition for each side of the strip:

Assumption-Definition 2.1 (Uniform Kreiss-Lopatinskii condition) For all frequency parameter, $\zeta \in$ $\Xi$, we still denote by $E^{s}(\zeta)$ and $E^{u}(\zeta)$ the extension up to $\Xi_{0}$ of the stable/unstable space defined in Lemma 2.1, we have:

$$
\operatorname{ker} B_{0} \cap E^{s}(\zeta)=\{0\} \text { and } \operatorname{ker} B_{1} \cap E^{u}(\zeta)=\{0\}
$$

so that we have the decompositions

$$
\begin{align*}
& \operatorname{ker} B_{0} \oplus E^{s}(\zeta)=\mathbb{C}^{N}  \tag{15}\\
& \operatorname{ker} B_{1} \oplus E^{u}(\zeta)=\mathbb{C}^{N} \tag{16}
\end{align*}
$$

Equivalently the restriction of $B_{0}$ (resp. $B_{1}$ ) to the stable subspace $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ ) is uniformly invertible in terms of $\zeta$. More precisely there exists $C_{0}, C_{1}>0$ such that for all $\zeta \in \Xi$ :

$$
\forall v \in E^{s}(\zeta),|v| \leq C_{0}\left|B_{0} v\right| \text { and } \forall w \in E^{u}(\zeta),|w| \leq C_{1}\left|B_{1} w\right|
$$

where we stress that the constants $C_{0}$ and $C_{1}$ do not depend on $\zeta \in \Xi$.
We introduce the notations

$$
\begin{equation*}
\phi_{0}(\zeta):=B_{0 \mid E^{s}(\zeta)}^{-1} \text { and } \phi_{1}(\zeta):=B_{1 \mid E^{u}(\zeta)}^{-1} \tag{17}
\end{equation*}
$$

The classical work of Kreiss, 1970 states that the boundary value problem in the half space 12) (resp. (13)) is strongly well-posed if and only if the uniform Kreiss-Lopatinskii condition (15) (resp. 16p) is satisfied.

Some parts of the proofs in the article are a bit technical, especially the ones based on several decomposition of the space $\mathbb{C}^{N}$. That is why, for the reader's convenience, we summarize some of the projectors used in this article in the following definition

Definition 2.1 (Projectors) From Lemma 2.1, for all $\zeta \in \Xi \backslash \Xi_{0}$, we have the decomposition:

$$
\mathbb{C}^{N}=E^{s}(\zeta) \oplus E^{u}(\zeta)
$$

and we recall that $\Pi^{s}(\zeta)$ (resp. $\left.\Pi^{u}(\zeta)\right)$ is the projector upon $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ ) with respect to the decomposition $\mathbb{C}^{N}=E^{s}(\zeta) \oplus E^{u}(\zeta)$.

From Assumption-Definition 2.1, for all $\zeta \in \Xi$, we have the decomposition

$$
\mathbb{C}^{N}=\operatorname{ker} B_{0} \oplus E^{s}(\zeta),
$$

and we define by $\bar{\Pi}^{s}(\zeta)$ the projector upon $E^{s}(\zeta), \bar{\Pi}_{\text {ker } B_{0}}(\zeta)$ the projector upon ker $B_{0}$ with respect to the decomposition $\mathbb{C}^{N}=\operatorname{ker} B_{0} \oplus E^{s}(\zeta)$. Similarly we have the decomposition

$$
\mathbb{C}^{N}=\operatorname{ker} B_{1} \oplus E^{u}(\zeta),
$$

so that we define by $\bar{\Pi}^{u}(\zeta)$ the projector upon $E^{u}(\zeta), \bar{\Pi}_{\mathrm{ker} B_{1}}(\zeta)$ the projector upon $\mathrm{ker} B_{1}$ with respect to the decomposition $\mathbb{C}^{N}=\operatorname{ker} B_{1} \oplus E^{s}(\zeta)$.

Finally from Assumption 2.4, we have the decomposition

$$
\mathbb{R}^{N}=\operatorname{ker} B_{0} \oplus \operatorname{ker} B_{1},
$$

and we define $\Pi_{\text {ker } B_{0}}$ (resp. $\Pi_{\mathrm{ker} B_{1}}$ ) the projector upon ker $B_{0}$ (resp. ker $B_{1}$ ) with respect to the decomposition $\mathbb{R}^{N}=\operatorname{ker} B_{0} \oplus \operatorname{ker} B_{1}$.

To conclude this paragraph we make clear the concepts of well-posedness that we will use in the following definition. Both of them are natural generalizations of the ones introduced in the half space geometry.

Definition 2.2 i) The hyperbolic boundary value problem in the strip (1) is said to be strongly well-posed if and only if, for all source terms $f \in L_{\gamma}^{2}(\Omega), g_{0} \in L_{\gamma}^{2}\left(\partial \Omega_{0}\right)$ and $g_{1} \in L_{\gamma}^{2}\left(\partial \Omega_{1}\right)$, the system (1) admits a unique solution $u \in L_{\gamma}^{2}(\Omega)$, with traces in $L_{\gamma}^{2}\left(\partial \Omega_{0}\right)$ and $L_{\gamma}^{2}\left(\partial \Omega_{1}\right)$ satisfying the energy estimate: there exists $C>0$ and $\gamma_{0} \geq 0$ such that for all $\gamma>\gamma_{0}$;

$$
\begin{equation*}
\gamma\|u\|_{L_{\gamma}^{2}(\Omega)}^{2}+\left\|u_{\mid x_{d}=0}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{0}\right)}^{2}+\left\|u_{\mid x_{d}=1}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{1}\right)}^{2} \leq C\left(\frac{1}{\gamma}\|f\|_{L_{\gamma}^{2}(\Omega)}^{2}+\left\|g_{0}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{0}\right)}^{2}+\left\|g_{1}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{1}\right)}^{2}\right) \tag{18}
\end{equation*}
$$

ii) If the hyperbolic boundary value problem in the strip (1) is strongly well-posed with $\gamma_{0}=0$ then it is said to be lower exponentially strongly well-posed.

We recall that if 18 holds for all $\gamma>\gamma_{0}$ and $\gamma_{0} \neq 0$ then the solution may admit a exponential growth in time whereas when the problem is lower exponentially strongly well-posed such a growth is impossible.

### 2.2 Main result

The main result of this article is the following. It gives a necessary and sufficient condition for lower exponential strong well-posedness.

Theorem 2.2 Under Assumptions 2.1 2.2 2.3 and 2.4, the hyperbolic boundary value problem in the strip (1) is lower exponentially strongly well-posed in the sense of Definition 2.2 if and only if each boundary condition satisfies the uniform Kreiss-Lopatinskii condition (see Assumption-Definition 2.1) and if moreover
the matrix $\left(I-T_{0 \rightarrow 0}(\zeta)\right):=\left(I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right)\left(\right.$ resp. $\left.\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right):=\left(I-\bar{T}_{1 \rightarrow 0}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right)\right)$ where the matrices $T_{0 \rightarrow 1}(\zeta)$ and $T_{1 \rightarrow 0}(\zeta)$ (resp. $\bar{T}_{0 \rightarrow 1}(\zeta)$ and $\bar{T}_{1 \rightarrow 0}(\zeta)$ ) are respectively given by:

$$
\begin{aligned}
T_{0 \rightarrow 1}(\zeta) & :=\phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \text { and } T_{1 \rightarrow 0}(\zeta)=\phi_{0}(\zeta) B_{0} e^{-\mathscr{A}(\zeta)}, \\
\left(\text { resp. } \bar{T}_{0 \rightarrow 1}(\zeta)\right. & \left.:=\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}(\zeta)} \text { and } \bar{T}_{1 \rightarrow 0}(\zeta)=\left(1-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)}\right)
\end{aligned}
$$

is uniformly invertible over $E^{s}(\zeta)$ (resp. ker $B_{0}$ ) compared to the frequency parameter $\zeta \in \Xi \backslash \Xi_{0}$. That is that there exist $C, \bar{C}>0$ such that for all $\zeta \in \Xi \backslash \Xi_{0}$ we have:

$$
\begin{gathered}
\forall v \in E^{s}(\zeta),|v| \leq C\left|\left(I-T_{0 \rightarrow 0}(\zeta)\right) v\right| \\
\forall w \in \operatorname{ker} B_{0},|w| \leq \overline{C \mid}\left|\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right) w\right| .
\end{gathered}
$$

This theorem has to be compared with the case of a hyperbolic system of equations in a corner domain studied in Osher, 1973-Osher, 1974b. Indeed, in this setting under a new inversibility condition involving Fourier integral operators and reading under the form $(I-T(\zeta))$, the author manages to construct symmetrizors leading to an a priori energy estimate. However, in this analysis the number of losses of derivatives in the energy estimate is non-explicit. This is due to the difficulty to study, with enough precision, the structure of the integral Fourier operator $T(\zeta)$.

In Benoit, 2015, the author precises the results of Osher, 1973 - Osher, 1974b by showing that if the operator $T(\zeta)$ is uniformly bounded in terms of the parameter $\zeta$ then the obtained a priori energy estimate is without loss. Moreover, the boundedness (unfortunaltely not the uniform part of it) of this operator $T(\zeta)$ is also establishes.

For the strip problem the operator $T(\zeta)$, namely $T_{0 \rightarrow 0}(\zeta)$, is not a Fourier integral operator anymore but is just a matrix (see Theorem 2.2). Consequently this boundedness question is much more easy to deal with and permits us to show the lower exponential strong well-posedness of (1) because the a priori energy estimate that we are able to show is lossless.

Another difference compared to the quarter space geometry is that for the strip problem we do not ask the inversibility of the operator $(I-T(\zeta))$ on the all space but only on the stable subspace $E^{s}(\zeta)$. This restriction will be crucial especially to construct a weak solution for (1).

## 3 Lower exponential strong well-posedness of the problem in a strip

The first step in the study of strong well-posedness for the strip problem (1) is to reduce the analysis to the question of the lower exponetial strong well-posedness of the time-space resolvent problem (8). In order to do so we introduce the following concepts of strong well-posedness for the problem (8):

Definition 3.1 i) Under Assumptions 2.1 and 2.2 we say that the resolvent strip problem (8) is strongly well-posed if for all sources terms $f \in L^{2}\left([0,1], \mathbb{C}^{N}\right), g_{0}, g_{1} \in \mathbb{C}^{N}$, the system (8) admits a unique solution $u \in L^{2}\left([0,1], \mathbb{C}^{N}\right)$, with well-defined traces on $\partial \Gamma_{0}$ and $\partial \Gamma_{1}$ satisfying the energy estimate: there exists $C, \gamma_{0}>0$, such that for all $\gamma>\gamma_{0}$ we have:

$$
\begin{equation*}
\gamma \int_{0}^{1}\left|u\left(x_{d}\right)\right|^{2} d x_{d}+\left|u_{\mid x_{d}=0}\right|^{2}+\left|u_{\mid x_{d}=1}\right|^{2} \leq C\left(\frac{1}{\gamma} \int_{0}^{1}\left|f\left(x_{d}\right)\right|^{2} d x_{d}+\left|g_{0}\right|^{2}+\left|g_{1}\right|^{2}\right) . \tag{19}
\end{equation*}
$$

ii) If the resolvent strip problem is (8) strongly well-posed in the sense of i) and if moreover one can choose $\gamma_{0}=0$ then we say that this problem is lower exponentially strongly well-posed.

The following proposition states the equivalence between the two definitions of strong well-posedness introduced so far. The proof of this proposition is a straightforward generalization of the ones in the half space geometry or in the quarter space geometry and will be ommited here.

Proposition 3.1 The hyperbolic strip boundary value problem (1) is strongly well-posed (resp. lower exponentially strongly well-posed) in the sense of Definition 2.2 if and only if the time-space resolvent problem is strongly well-posed (resp. lower exponentially strongly well-posed) in the sense of Definition 3.1.

Consequently (19) written for $f \equiv 0, g_{1} \equiv 0$ and $f \equiv 0, g_{0} \equiv 0$ imply that impose the uniform KreissLopatinskii condition (see Definition 2.1) on each component of the boundary is a necessary condition for the strong well-posedness of (1).

### 3.1 The case of strictly dissipative boundary conditions

There exists particular boundary conditions for which the proof of the lower exponential strong well-posedness of the strip problem (1) is much more simple. Namely these boundary conditions are the strictly dissipative ones, that is to say the boundary conditions that impose that the energy descreases during a reflection.

In this setting the proof of the a priori energy estimate is classical and relies on a standard integration by parts argument. Then the existence of a weak solution comes from a slight adaptation of the duality argument introduced in Lax and Phillips, 1960 to deal with the geometry of the half space. Finally the uniqueness of the solution is provided by the fact that from a "weak=strong" lemma, the weak solution satisfies the energy estimate 18 .

In this paragraph we describe this proof. This is firstly motivated by the fact that this proof in this simplified setting is a good preliminary work to study the proof in the general one. But also because the proof in the general setting is based upon the lower exponential strong well-posedness of the problem in a strip for strictly dissipative boundary conditions (see Osher, 1973 or Benoit, 2015 for more details in the quarter space geometry).

Moreover for a fixed symmetric hyperbolic operator $L(\partial)$ strictly dissipative boundary conditions always exist and are algebrically characterized. Consequently these conditions are rather easy to determine compared to the boundary conditions that only satisfy the uniform Kreiss-Lopatinskii condition. Thus the strictly dissipative framework is an interesting one because it gives a non empty and easily determinable set of suitable boundary conditions ensuring lower exponential strong well-posedness.

Before to state the lower exponential strong well-posedness result we give some more details about what we mean by strict dissipativity as well as some classical properties of these particular boundary conditions.

Definition-Proposition 3.1 The boundary condition on the side $\partial \Gamma_{0}$ (resp. $\partial \Gamma_{1}$ ) is said to be strictly dissipative is the following inequality holds:

$$
\begin{equation*}
\left.\forall v \in \operatorname{ker} B_{0} \backslash\{0\},\left\langle A_{d} v, v\right\rangle<0\left(\text { resp. } \operatorname{ker} B_{1} \backslash\{0\}\right),\left\langle A_{d} v, v\right\rangle>0\right) \tag{20}
\end{equation*}
$$

Or equivalently (see for example Benzoni-Gavage and Serre, 2007), the boundary condition on the side $\partial \Gamma_{0}$ (resp. $\partial \Gamma_{1}$ ) is strictly dissipative if and only if there exists $\varepsilon_{0}, C_{0}>0$ (resp. $\varepsilon_{1}, C_{1}>0$ ) such that:

$$
\begin{equation*}
\forall v \in \mathbb{R}^{N}, \varepsilon_{0}|v|^{2}+\left\langle A_{d} v, v\right\rangle-C_{0}\left|B_{0} v\right|^{2}<0,\left(\text { resp. } \varepsilon_{1}|v|^{2}+\left\langle A_{d} v, v\right\rangle-C_{1}\left|B_{1} v\right|^{2}>0\right) \tag{21}
\end{equation*}
$$

Under Assumption 2.2 for all fixed operator $L(\partial)$ with symmetrics coefficients there exists boundary matrices $B_{0}$ and $B_{1}$ such that the boundary conditions in (1) are strictly dissipative.

Finally under Assumptions 2.1 2.2 if the boundary condition on the side $\partial \Gamma_{0}$ (resp. $\partial \Gamma_{1}$ ) is strictly dissipative then it satisfies the uniform Kreiss-Lopatinskii condition on the side $\partial \Gamma_{0}$ (resp. $\partial \Gamma_{1}$ ) (we refer, for example, to Benzoni-Gavage and Serre, 2007-Paragraph 4.3.4 for a proof).

In particular, we remark that if the boundary conditions in (1) are strictly dissipative then the necessary Assumption 2.4 is trivially satisfied because of 20 .

The lower exponential strong well-posedness result for strictly dissipative boundary condition is the following:
Theorem 3.1 Under Assumptions 2.2 2.3, assume that each boundary condition in (1) is strictly dissipative in the sense of Definition-Proposition 3.1. Then the strip problem (1) is lower exponentially strongly well posed in the sense of Definition 2.2.

The proof of Theorem 3.1 is given in the following paragraphs.

### 3.1.1 The a priori energy estimate

The main point in the proof of Theorem (3.1) is the following a priori energy estimate, based upon an easy integration by parts argument.

Proposition 3.2 (A priori energy estimate) Under Assumption 2.3, assume that each boundary condition in (1) is strictly dissipative in the sense of Definition-Proposition 3.1, then let $u$ be a regular (for example $H^{1}(\Omega)$ with decay at infinity in space) solution of (11), then $u$ satisfies the a priori energy estimate that is there exists $C>0$ such that for all $\gamma>0$ we have

$$
\begin{equation*}
\gamma\|u\|_{L_{\gamma}^{2}(\Omega)}^{2}+\left\|u_{\mid x_{d}=0}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{0}\right)}^{2}+\left\|u_{\mid x_{d}=1}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{1}\right)}^{2} \leq C\left(\frac{1}{\gamma}\|f\|_{L_{\gamma}^{2}(\Omega)}^{2}+\left\|g_{0}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{0}\right)}^{2}+\left\|g_{1}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{1}\right)}^{2}\right) \tag{22}
\end{equation*}
$$

where we stress that $C$ does not depend on $\gamma$.
Proof : Let $\widehat{u}$ be the Laplace in time, Fourier in the $x^{\prime}$-space transform of $u$ the regular solution of (1). Then $\widehat{u}$ solves the resolvent strip problem (8).

Multiply the evolution equation of (8) by $\widehat{u}$ and integrate over $[0,1]$ with respect to the variable $x_{d}$ gives:

$$
\begin{equation*}
\int_{0}^{1}\left\langle A_{d} \frac{\mathrm{~d}}{\mathrm{~d} x_{d}} \widehat{u}, \widehat{u}\right\rangle \mathrm{d} x_{d}+\int_{0}^{1}\langle M(\zeta) \widehat{u}, \widehat{u}\rangle \mathrm{d} x_{d}=\int_{0}^{1}\langle\widehat{f}, \widehat{u}\rangle \mathrm{d} x_{d} \tag{23}
\end{equation*}
$$

where we recall that $M(\zeta)$ is defined in 10$)$. By symmetry of $A_{d}$ we derive the relation:

$$
\frac{\mathrm{d}}{\mathrm{~d} x_{d}}\left\langle A_{d} \widehat{u}, \widehat{u}\right\rangle=2 \operatorname{Re}\left\langle A_{d} \frac{\mathrm{~d}}{\mathrm{~d} x_{d}} \widehat{u}, \widehat{u}\right\rangle
$$

and from the symmetry of the $A_{j}, j=1, \ldots d-1$ and the special form of $M$ we obtain by taking the real part of 23):

$$
\left\langle A_{d} \widehat{u}_{\mid x_{d}=1}, \widehat{u}_{\mid x_{d}=1}\right\rangle-\left\langle A_{d} \widehat{u}_{\mid x_{d}=0}, \widehat{u}_{\mid x_{d}=0}\right\rangle+2 \gamma \int_{0}^{1}|\widehat{u}|^{2} \mathrm{~d} x_{d}=2 \operatorname{Re} \int_{0}^{1}\langle\widehat{f}, \widehat{u}\rangle \mathrm{d} x_{d}
$$

Then by strict dissipativity (see Definition-Proposition 3.1) and Cauchy-Schwarz/Young inequalities applied to the right hand side it follows:

$$
\gamma \int_{0}^{1}|\widehat{u}|^{2} \mathrm{~d} x_{d}+\left|u_{\mid x_{d}=0}\right|^{2}+\left|u_{\mid x_{d}=1}\right|^{2} \leq C\left(\frac{1}{\gamma} \int_{0}^{1}\left|\widehat{f}^{2} \mathrm{~d} x_{d}+\left|\widehat{g}_{0}\right|^{2}+\left|\widehat{g}_{1}\right|^{2}\right)\right.
$$

Finally we integrate over $\mathbb{R}^{d-1}$ with respect to the Fourier variable and over $\mathbb{R}$ with respect with the imaginary part of the Laplace variable. By Plancherel formula we obtain that $u$ satisfies 22).

### 3.1.2 Construction of a weak solution by duality method

To show the existence of a weak solution of (1) we use a duality method. This method is made possible due to the fact that strictly dissipative boundary conditions are suitable for the dual problem. More precisely we mean that if the primal problem (1) admits strictly dissipative boundary conditions then we are free to define a problem associated to (1) the so-called dual problem which also admits strictly dissipative boundary conditions.

Note that as for the half space geometry, as long as the boundary conditions are strictly dissipative, an other proof based on Hile-Yosida theorem Yosida, 1995 (we refer to Benzoni-Gavage and Serre, 2007 Chapter 3 for a precise proof) should also operate and be more straightfull. However as we will see in the general framework (that is to say when boundary conditions are not strictly dissipative anymore), see Paragraph 3.3.3, this duality method will be more relevant, that is why we decided to describ it in the setting of strictly dissipative boundary conditions.

To use this duality method we introduce the dual problem:

Definition 3.2 The strip problem:

$$
\begin{cases}L^{*}(\partial) v=f & \text { on } \Omega  \tag{24}\\ C_{0} v_{\mid x_{d}=0}=g_{0} & \text { on } \partial \Omega_{0} \\ C_{1} v_{\mid x_{d}=1}=g_{1} & \text { on } \partial \Omega_{1}\end{cases}
$$

is said to be dual for the strip problem (1) if for all regular $u$ and $v$ (for example $H^{1}(\Omega)$ we have the duality formula:

$$
\begin{equation*}
\langle L(\partial) u, v\rangle-\left\langle u, L^{*}(\partial) v\right\rangle=\sum_{j=0}^{1}(-1)^{j+1}\left(\left\langle N_{j} u_{\mid x_{d}=j}, C_{j} v_{\mid x_{d}=j}\right\rangle+\left\langle B_{j} u_{\mid x_{d}=j}, M_{j} v_{\mid x_{d}=j}\right\rangle\right) \tag{25}
\end{equation*}
$$

where the matrices $N_{0}, C_{0}, M_{1} \in \mathbf{M}_{N-p \times N}(\mathbb{R}), N_{1}, C_{1}, M_{0} \in \mathbf{M}_{p \times N}(\mathbb{R})$, appearing in the right hand side of (25) are characterized by the relations:

$$
\begin{equation*}
j=0,1, A_{d}=C_{j}^{T} N_{j}+M_{j}^{T} B_{j} \tag{26}
\end{equation*}
$$

The following proposition, whose proof is classical (see for example Benzoni-Gavage and Serre, 2007) ensures the existence of a dual problem to (1).

Proposition 3.3 Under Assumptions 2.2 and 2.3 let $L^{*}(\partial)$ be given by:

$$
L^{*}(\partial):=-L(\partial)
$$

then there exists matrices $N_{0}, C_{0}, M_{1} \in \mathbf{M}_{N-p \times N}(\mathbb{R}), N_{1}, C_{1}, M_{0} \in \mathbf{M}_{p \times N}(\mathbb{R})$ such that (24) is a dual problem for (1). The matrices $C_{j}$ are characterized by the relations:

$$
\begin{equation*}
\operatorname{ker} C_{j}=\left(A_{d} \operatorname{ker} B_{j}\right)^{\perp} \tag{27}
\end{equation*}
$$

the matrices $N_{j}$ are onto and satisfy:

$$
\operatorname{ker} B_{j} \cap \operatorname{ker} N_{j}=\{0\}
$$

The following result which is a slight modification of the arguments of Lax and Phillips, 1960 shows that if the strip problem (1) admits strictly dissipative boundary conditions then so do for the associated dual problem (24). More precisely:
Lemma 3.1 If the strip problem (8) satisfies Assumptions 2.2 2.3 and admits strictly dissipative boundary conditions in the sense of Definition-Proposition 3.1 then for all choices of $C_{0} \in \mathbf{M}_{N-p \times N}(\mathbb{R}), C_{1} \in$ $\mathbf{M}_{p \times N}(\mathbb{R})$ satisfying (27) the dual strip problem (24) satisfies Assumptions 2.2 .2.3 and the associated boundary condition are strictly dissipative.

Proof : The fact that $(24)$ also satisfies Assumptions $2.2 \mid 2.3$ is clear from the definition of $L^{*}(\partial)$ and from the characterization $(27)$. We now turn to the strict dissipativity of the boundary conditions.

As $L^{*}(\partial)=-L(\partial)$ we shall show that:

$$
\forall v \in \operatorname{ker} C_{0},\left\langle A_{d} v, v\right\rangle>0, \text { and } \forall w \in \operatorname{ker} C_{1},\left\langle A_{d} w, w\right\rangle<0
$$

We only show the first inequality, the proof of the second one is totally similar (except for the sign).
By contradiction we assume that there exists $\underline{v} \in \operatorname{ker} C_{0}$ such that

$$
\left\langle A_{d} \underline{v}, \underline{v}\right\rangle \leq 0
$$

Then for all $u \in \operatorname{ker} B_{0}$ and for all $\lambda \in \mathbb{R}$ we have:

$$
\left\langle A_{d}(u+\lambda \underline{v}), u+\lambda \underline{v}\right\rangle=\left\langle A_{d} u, u\right\rangle+2 \lambda\left\langle A_{d} u, \underline{v}\right\rangle+\lambda^{2}\left\langle A_{d} \underline{v}, \underline{v}\right\rangle<0
$$

Indeed from (27) the second term in the right hand side is zero, by strict dissipativity of the boundary condition on $\partial \Gamma_{0}$ the first one is strictly negative and the last one is negative by construction of $\underline{v}$.

As we assumed ker $B_{0}$ to be of maximal rank for the property $\left\langle A_{d} \cdot, \cdot\right\rangle<0$ it implies that $u+\lambda \underline{v} \in \operatorname{ker} B_{0}$ and consequently $\underline{v} \in \operatorname{ker} B_{0}$. As a consequence we have, by (26), that $A_{d} \underline{v}=C_{0}^{T} N_{0} \underline{v}$ and thus

$$
\left\langle A_{d} \underline{v}, \underline{v}\right\rangle=\left\langle C_{0} \underline{v}, N_{0} \underline{v}\right\rangle=0
$$

because $\underline{v} \in \operatorname{ker} C_{0}$. This contradicts the strict dissipativity of the boundary condition $\partial \Gamma_{0}$.

Lemma 3.1 also admits an other version in the general setting that is the one of boundary conditions that only satisfy the uniform Kreiss-Lopatinskii condition. More precisey, we refer for example to [ Benzoni-Gavage and Serre, 2007-Pagraph 4.4] that we have the following result:

Lemma 3.2 Under Assumptions 2.1 2.2 the primal problem satisfies the uniform Kreiss-Lopatinskii condition for the side $\partial \Gamma_{0}$ (resp. the side $\partial \Gamma_{1}$ ) if and only if the dual problem defined in (24) satisfies the backward uniform Kreiss-Lopatinskii condition for the side $\partial \Gamma_{0}$ (resp. the side $\partial \Gamma_{1}$ ).

More precisely it means that we have the equivalence

$$
\operatorname{ker} B_{0} \cap E^{s}(\zeta)=\{0\} \Leftrightarrow \operatorname{ker} C_{0} \cap \widetilde{E}^{s}(\zeta)=\{0\}
$$

and

$$
\operatorname{ker} B_{1} \cap E^{u}(\zeta)=\{0\} \Leftrightarrow \operatorname{ker} C_{1} \cap \widetilde{E}^{u}(\zeta)=\{0\}
$$

where $\widetilde{E}^{s}(\zeta)$ and $\widetilde{E}^{u}(\zeta)$ stand for the stable/unstable subspaces associated to $\widetilde{\mathscr{A}}(\sigma, \eta):=\mathscr{A}(-\sigma, \eta)$ the resolvent matrix of (24).

Consequently we define for the dual problem (24) we define the inverses $\widetilde{\phi}_{0}(\zeta):=C_{0_{\mid \tilde{E}^{s}(\zeta)}}$ and $\widetilde{\phi}_{1}(\zeta):=$ $C_{1_{\mid \tilde{E}^{u}(\zeta)}}$

This result will be the keystone in the construction of the symmetrizor in the general setting and also to construct a weak solution by duality method.

We now go back to the particular case of strictly dissipative boundary conditions and we show that Lemma 3.1 implies the existence of a weak solution $u \in L_{\gamma}^{2}(\Omega)$ of $\mathbb{1}$ to conclude this paragraph.

For simplicity, for $\chi \in \mathbb{R}$, we introduce the functional space $\mathscr{H}_{\chi}$ by

$$
\mathscr{H}_{\chi}:=L_{\chi}^{2}(\Omega) \times L_{\chi}^{2}\left(\partial \Omega_{0}\right) \times L_{\chi}^{2}\left(\partial \Omega_{1}\right)
$$

which is a Hilbert space with the product norm:

$$
\|\cdot\|_{\mathscr{H}_{\chi}}:=\sqrt{\|\cdot\|_{L_{\chi}^{2}(\Omega)}^{2}+\|\cdot\|_{L_{\chi}^{2}\left(\partial \Omega_{0}\right)}^{2}+\|\cdot\|_{L_{\chi}^{2}\left(\partial \Omega_{1}\right)}^{2}} .
$$

Proposition 3.4 Under Assumptions 2.2 [2.3 if the boundary conditions of the strip problem (1) are strictly dissipative in the sense of Definition-Proposition 3.1, then the strip problem (1) admits a weak solution (in the distributional sense) $u \in L_{\gamma}^{2}(\Omega)$.

The proof of Proposition 3.4 is rather classical. Indeed it is a straightforward adaptation of the one for the boundary value problem in the half space given, for example, in [ Benzoni-Gavage and Serre, 2007]Paragraph 4.5.3, see also [Sarason, 1962 for a slight modification of this proof]. However we recall this proof for a sake of completeness.

Proof : We define

$$
X_{-\gamma}:=\left\{L^{*}(\partial) v \text { for } v \in H_{-\gamma}^{1}(\Omega) \text { satisfying } C_{0} v_{\mid x_{d}=0}=0, C_{1} v_{\mid x_{d}=1}=0\right\}
$$

Using Lemma 3.1, the dual problem (24) satisfies the assumptions of Proposition 3.2 Consequently, as $v \in H_{-\gamma}^{1}(\Omega)$ satisfies $C_{0} v_{\mid x_{d}=0}=C_{1} v_{\mid x_{d}=1}=0$ we deduce that we have the energy estimate:

$$
\gamma\|v\|_{L_{-\gamma}^{2}(\Omega)}^{2}+\left\|v_{\mid x_{d}=0}\right\|_{L_{-\gamma}^{2}\left(\partial \Omega_{0}\right)}^{2}+\left\|v_{\mid x_{d}=1}\right\|_{L_{-\gamma}^{2}\left(\partial \Omega_{1}\right)}^{2} \leq \frac{C}{\gamma}\left\|L^{*}(\partial) v\right\|_{L_{-\gamma}^{2}(\Omega)}^{2},
$$

that is to say

$$
\begin{equation*}
\|v\|_{\mathscr{H}_{-\gamma}} \leq C_{\gamma}\left\|L^{*}(\partial) v\right\|_{L_{-\gamma}^{2}(\Omega)} \tag{28}
\end{equation*}
$$

Let $\left(f, g_{0}, g_{1}\right) \in \mathscr{H}_{\gamma}$ we define a linear form $\ell$ on $X_{-\gamma}$ by:

$$
\begin{equation*}
\ell\left(L^{*}(\partial) v\right):=\left\langle f, L^{*}(\partial) v\right\rangle-\left\langle g_{1}, M_{1} v_{\mid x_{d}=1}\right\rangle+\left\langle g_{0}, M_{0} v_{\mid x_{d}=1}\right\rangle \tag{29}
\end{equation*}
$$

Then combining Cauchy-Schwarz inequality and (28) we deduce that:

$$
\begin{aligned}
\left|\ell\left(L^{*}(\partial) v\right)\right| & \leq C\left(\|f\|_{L_{\gamma}^{2}(\Omega)}\|v\|_{L_{-\gamma}^{2}(\Omega)}+\left\|g_{0}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{0}\right)}\left\|v_{\mid x_{d}=0}\right\|_{L_{-\gamma}^{2}\left(\partial \Omega_{0}\right)}+\left\|g_{1}\right\|_{L_{\gamma}^{2}\left(\partial \Omega_{1}\right)}\left\|v_{\mid x_{d}=1}\right\|_{L_{-\gamma}^{2}\left(\partial \Omega_{1}\right)}\right) \\
& \leq C_{\gamma}\left\|L^{*}(\partial) v\right\|_{L_{-\gamma}^{2}(\Omega)} .
\end{aligned}
$$

Consequently, from Hahn-Banach theorem, $\ell$ can be extended by continuity up to $L_{-\gamma}^{2}(\Omega)$ and by Riesz representation theorem, there exists $\underline{u} \in L_{\gamma}^{2}(\Omega)$ such that for all $L^{*}(\partial) v \in L_{-\gamma}^{2}(\Omega)$ we have:

$$
\ell\left(L^{*}(\partial) v\right)=\left\langle\underline{u}, L^{*}(\partial) v\right\rangle
$$

which by definition of $\ell$ is equivalent to:

$$
\left\langle f, L^{*}(\partial) v\right\rangle-\left\langle\underline{u}, L^{*}(\partial) v\right\rangle=\left\langle g_{1}, M_{1} v_{\mid x_{d}=1}\right\rangle-\left\langle g_{0}, M_{0} v_{\mid x_{d}=1}\right\rangle .
$$

For $u \in L_{\gamma}^{2}(\Omega)$, the traces $u_{\mid x_{d}=0} \in H_{\gamma}^{-\frac{1}{2}}\left(\partial \Omega_{0}\right)$ and $u_{\mid x_{d}=1} \in H_{\gamma}^{-\frac{1}{2}}\left(\partial \Omega_{1}\right)$ are well-defined and by Green formula we have that for all $v \in H_{-\gamma}^{1}(\Omega)$ :

$$
\begin{align*}
&\langle L(\partial) u, v\rangle_{H_{\gamma}^{-1}(\Omega) ; H_{-\gamma}^{1}(\Omega)}-\left\langle u, L^{*}(\partial) v\right\rangle=\sum_{j=0}^{1}(-1)^{j+1}\left(\left\langle B_{j} u_{\mid x_{d}=j}, M_{j} v_{\mid x_{d}=j}\right\rangle_{H_{\gamma}^{-\frac{1}{2}}\left(\partial \Omega_{j}\right) ; H_{\gamma}^{\frac{1}{2}}\left(\partial \Omega_{j}\right)}\right. \\
&\left.+\left\langle N_{j} u_{\mid x_{d}=j}, C_{j} v_{\mid x_{d}=j}\right\rangle_{H_{\gamma}^{-\frac{1}{2}}\left(\partial \Omega_{j}\right) ; H_{\gamma}^{\frac{1}{2}}\left(\partial \Omega_{j}\right)}\right) \tag{30}
\end{align*}
$$

Testing (30) written for $u=\underline{u}$, against functions $L^{*}(\partial) v$ where $v \in \mathscr{D}(\Omega)$ gives that $L(\partial) \underline{u}=f$ in the distributional sense and that $L(\partial) \underline{u} \in L_{\gamma}^{2}(\Omega)$. Next test (30) against functions in $L^{*}(\partial) v$ where $v \in$ $\mathscr{D}\left(\mathbb{R} \times \mathbb{R}^{d-1} \times\left[0,1[)\right.\right.$ satisfies $v_{\mid x_{d}=0} \in \operatorname{ker} C_{0}\left(\right.$ resp. $\left.\left.v \in \mathscr{D}\left(\mathbb{R} \times \mathbb{R}^{d-1} \times\right] 0,1\right]\right)$ satisfies $\left.v_{\mid x_{d}=1} \in \operatorname{ker} C_{1}\right)$ gives

$$
\begin{equation*}
\left\langle g_{0}-B_{0} \underline{u}_{\mid x_{d}=0}, M_{0} v_{\mid x_{d}=0}\right\rangle_{H_{\gamma}^{-\frac{1}{2}}\left(\partial \Omega_{0}\right) ; H_{\gamma}^{\frac{1}{2}}\left(\partial \Omega_{0}\right)}=0,\left(\operatorname{resp} .\left\langle B_{1} \underline{u}_{\mid x_{d}=1}-g_{1}, M_{1} v_{\mid x_{d}=1}\right\rangle_{H_{\gamma}^{-\frac{1}{2}}\left(\partial \Omega_{1}\right) ; H_{\gamma}^{\frac{1}{2}}\left(\partial \Omega_{1}\right)}=0\right) . \tag{31}
\end{equation*}
$$

Using the characterization $\operatorname{ker} C_{j}=\left(A_{d} \operatorname{ker} B_{j}\right)^{\perp}$, the fact that $B_{j}$ is onto we deduce that $M_{j_{\mid \operatorname{ker} C_{j}}}$ is onto so that the test function in (31) can be remplaced by any test function. Consequently $B_{0} \underline{u}_{\mid x_{d}=0}=g_{0}$ and $B_{1} \underline{u}_{\mid x_{d}=1}=g_{1}$ in the distributional sense and $B_{0} \underline{u}_{\mid x_{d}=0} \in L_{\gamma}^{2}\left(\partial \Omega_{0}\right), B_{1} \underline{u}_{\mid x_{d}=1} \in L_{\gamma}^{2}\left(\partial \Omega_{1}\right)$, which concludes the existence of a weak solution for (1).

### 3.1.3 The "weak=strong" lemma

To conclude the proof of Theorem 3.1 it remains to show that the weak solution constructed in Proposition 3.4 admits the extra trace regularity (that is $u_{\mid x_{d}=j} \in L_{\gamma}^{2}\left(\partial \Omega_{j}\right)$ ) and that it satisfies the energy estimate (22), which immediately implies uniqueness by linearity.

The interesting point here is that, due to the fact that the strip geometry "looks like" the half space geometry, the classical proof for the half space (see, for example [Benzoni-Gavage and Serre, 2007], Lemma 4.7] or Chazarain and Piriou, 1981) can be used to show these points in the strip geometry ${ }^{1}$. More precisely, we have:

Lemma 3.3 ("Weak=strong") Let $u \in L_{\gamma}^{2}(\Omega)$ be a solution of 11 such that $f:=L(\partial) u \in L_{\gamma}^{2}(\Omega)$, $g_{0}=\left(B_{0} u\right)_{\mid x_{d}=0} \in L^{2}\left(\partial \Omega_{0}\right)$ and $g_{1}=\left(B_{1} u\right)_{\mid x_{d}=1} \in L^{2}\left(\partial \Omega_{1}\right)$ then:

[^0]- $u_{\mid x_{d}=0} \in L^{2}\left(\partial \Omega_{0}\right)$ and $u_{\mid x_{d}=1} \in L^{2}\left(\partial \Omega_{1}\right)$.
- u satisfies 22).

Note that this "weak=strong" lemma does not depend on the kind of considered boundary conditions. As a consequence, Lemma 3.3 can also be used in the general framework of non strictly dissipative boundary conditions.

Proof : Following Benzoni-Gavage and Serre, 2007, it is sufficient to consider the case where $u \in L_{\gamma}^{2}(\Omega)$ has a compact support. Let $\rho^{\varepsilon}$ be a mollifier in the tangential variables that is $\left(t, x^{\prime}\right)$ then $u^{\varepsilon}:=\rho^{\varepsilon} *_{\left(t, x^{\prime}\right)} u$ is a compactly supported function which is smooth in $\left(t, x^{\prime}\right)$.

The same holds for $L(\partial) u^{\varepsilon}=\rho^{\varepsilon} *_{\left(t, x^{\prime}\right)} L(\partial) u \in L_{\gamma}^{2}(\Omega)$ and also for $\left(B_{0} u^{\varepsilon}\right)_{\mid x_{d}=0}=\rho^{\varepsilon} *_{\left(t, x^{\prime}\right)}\left(B_{0} u\right)_{\mid x_{d}=0} \in$ $L_{\gamma}^{2}\left(\partial \Omega_{0}\right),\left(B_{1} u^{\varepsilon}\right)_{\mid x_{d}=1}=\rho^{\varepsilon} *_{\left(t, x^{\prime}\right)}\left(B_{1} u\right)_{\mid x_{d}=1} \in L_{\gamma}^{2}\left(\partial \Omega_{1}\right)$ because $\rho$ does not depend on $x_{d}$.

Let $j=1, \ldots, d-1$ then we have $\partial_{j} u^{\varepsilon}=\left(\partial_{j} \rho\right) *_{\left(t, x^{\prime}\right)} u$ so that $\partial_{j} u^{\varepsilon} \in L_{\gamma}^{2}(\Omega)$. The same holds for $\partial_{t} u^{\varepsilon}$. Consequently the trivial identity

$$
\partial_{d} u^{\varepsilon}=A_{d}^{-1}\left[L(\partial) u^{\varepsilon}-\partial_{t} u^{\varepsilon}-\sum_{j=1}^{d-1} A_{j} \partial_{j} u^{\varepsilon}\right]
$$

implies that $\partial_{d} u^{\varepsilon} \in L_{\gamma}^{2}(\Omega)$ so that $u^{\varepsilon} \in H_{\gamma}^{1}(\Omega)$. We deduce that $u^{\varepsilon}$ satisfies 22 and using the classical properties of mollifiers we obtain, by passing to the limit in 22, that $u$ satisfies (22) and that $u_{\mid x_{d}=0} \in$ $L_{\gamma}^{2}\left(\partial \Omega_{0}\right), u_{\mid x_{d}=1} \in L_{\gamma}^{2}\left(\partial \Omega_{1}\right)$.

This Lemma concludes the proof of Theorem 3.1 .

### 3.2 Necessary conditions for strong well-posedness in the general setting

To derive necessary conditions for strong well-posedness we consider the problem:

$$
\left\{\begin{array}{l}
\left.\frac{\mathrm{d}}{\mathrm{~d} x_{d}} u=\mathscr{A}(\zeta) u \quad \text { for } x_{d} \in\right] 0,1[,  \tag{32}\\
B_{0} u_{\mid x_{d}=0}=g \\
B_{1} u_{\mid x_{d}=1}=0
\end{array}\right.
$$

which is assumed to be strongly well-posed and we extend the solution by zero for $x_{d} \geq 1$. The extension, denoted by $v$, satisfies:

$$
\left\{\begin{array}{l}
\left.\frac{\mathrm{d}}{\mathrm{~d} x_{d}} v=\mathscr{A}(\zeta) v-\delta_{x_{d}=1} v_{\mid x_{d}=1}, \quad \text { for } x_{d} \in\right] 0,+\infty[ \\
B_{0} u_{\mid x_{d}=0}=g
\end{array}\right.
$$

Then we apply Duhamel's formula to obtain:

$$
\begin{align*}
& \Pi^{s}(\zeta) v=e^{x_{d} \mathscr{A}(\zeta)} \Pi^{s}(\zeta) v_{\mid x_{d}=0}-\int_{0}^{x_{d}} \delta_{\mid s=1} e^{\left(x_{d}-s\right) \mathscr{A}(\zeta)} \mathrm{d} s \Pi^{s}(\zeta) v_{\mid x_{d}=1}  \tag{33}\\
& \Pi^{u}(\zeta) v=\int_{x_{d}}^{\infty} \delta_{\mid s=1} e^{\left(x_{d}-s\right) \mathscr{A}(\zeta)} \mathrm{d} s \Pi^{u}(\zeta) v_{\mid x_{d}=1} \tag{34}
\end{align*}
$$

Thanks to the uniform Kreiss-Lopatinskii condition on the side $\partial \Gamma_{0}$ we have:

$$
\Pi^{s}(\zeta) v_{\mid x_{d}=0}=\phi_{0}(\zeta) g-\phi_{0}(\zeta) B_{0} \Pi^{u}(\zeta) v_{\mid x_{d}=0}
$$

which combined with (33)-(34) gives:

$$
\begin{align*}
u_{\mid x_{d}=0}=v_{\mid x_{d}=0} & =\phi_{0}(\zeta) g-\left(I-\phi_{0}(\zeta) B_{0}\right) \int_{0}^{+\infty} \delta_{\mid s=1} e^{-s \mathscr{A}(\zeta)} \mathrm{d} s \Pi^{u}(\zeta) u_{\mid x_{d}=1} \\
& =\phi_{0}(\zeta) g+\left(\phi_{0}(\zeta) B_{0}-I\right) e^{-\mathscr{A}(\zeta)} \Pi^{u}(\zeta) u_{\mid x_{d}=1} \tag{35}
\end{align*}
$$

Then we extend the solution $u$ by zero for $x_{d} \leq 0$ to derive a relation between $u_{\mid x_{d}=1}$ and $u_{\mid x_{d}=0}$. More precisely, by repeating exactly the same kind of computations, we obtain:

$$
\begin{equation*}
u_{\mid x_{d}=1}=\left(\phi_{1}(\zeta) B_{1}-I\right) e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) u_{\mid x_{d}=0} \tag{36}
\end{equation*}
$$

Combine (35) and (36) then leads to the compatibility condition:

$$
\begin{align*}
u_{\mid x_{d}=0}= & \phi_{0}(\zeta) g+\left(\phi_{0}(\zeta) B_{0}-I\right) e^{-\mathscr{A}(\zeta)} \Pi^{u}(\zeta)\left(\phi_{1}(\zeta) B_{1}-I\right) e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) u_{\mid x_{d}=0} \\
= & \phi_{0}(\zeta) g+\left(\phi_{0}(\zeta) B_{0}-I\right) e^{-\mathscr{A}(\zeta)} \Pi^{u}(\zeta) \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) u_{\mid x_{d}=0} \\
& -\left(\phi_{0}(\zeta) B_{0}-I\right) e^{-\mathscr{A}(\zeta)} \underbrace{\Pi^{u}(\zeta) e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta)}_{=0} u_{\mid x_{d}=0} \\
= & \phi_{0}(\zeta) g-\left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) u_{\mid x_{d}=0}, \tag{37}
\end{align*}
$$

because for all $\zeta \in \Xi, \operatorname{Ran}\left(\phi_{1}(\zeta)\right)=E^{u}(\zeta)$.
For simplicity, let:

$$
\begin{equation*}
T_{0 \rightarrow 1}(\zeta)=\phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \text { and } T_{1 \rightarrow 0}(\zeta)=\phi_{0}(\zeta) B_{0} e^{-\mathscr{A}(\zeta)} \tag{38}
\end{equation*}
$$

in such a way that 37 becomes:

$$
\left(I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right) \Pi^{s}(\zeta) u_{\mid x_{d}=0}+\Pi^{u}(\zeta) u_{\mid x_{d}=0}+e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) u_{\mid x_{d}=0}=\phi_{0}(\zeta) g
$$

On the one hand, by identification on $E^{s}(\zeta)$ and $E^{u}(\zeta)$, it follows (by using the fact that $E^{u}(\zeta)$ is a stable space of $\left.e^{-\mathscr{A}(\zeta)}\right)$ :

$$
\left\{\begin{array}{l}
\left(I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right) \Pi^{s}(\zeta) u_{\mid x_{d}=0}=\phi_{0}(\zeta) g  \tag{39}\\
\Pi^{u}(\zeta) u_{\mid x_{d}=0}=-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) u_{\mid x_{d}=0}
\end{array}\right.
$$

from which we deduce that the unstable component of the trace of the solution $u$ is uniquely determined in terms of the stable component.

Then the first equation of 39 immediately implies that for all $\zeta \in \Xi \backslash \Xi_{0}$, the operator $I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)$ : $E^{s}(\zeta) \longrightarrow E^{s}(\zeta)$ is invertible and thus determines in a unique way the value of $\Pi^{s}(\zeta) u_{\mid x_{d}=0}$. Consequently, the unstable part of the trace of the solution is uniquely determined in terms of $g$ by:

$$
\Pi^{u}(\zeta) u_{\mid x_{d}=0}=-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)}\left(I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right)_{\mid E^{s}(\zeta)}^{-1} \phi_{0}(\zeta) g
$$

from which we deduce that

$$
\begin{equation*}
u_{\mid x_{d}=0}=\left(I-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{+\mathscr{A}(\zeta)}\right)\left(I-T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta)\right)_{\mid E^{s}(\zeta)}^{-1} \phi_{0}(\zeta) g \tag{40}
\end{equation*}
$$

If one reiterates exactly the same reasoning but this time with a non zero source term on the side $\left\{x_{d}=1\right\}$ instead of the side $\left\{x_{d}=0\right\}$, then he will recover the following necessary condition for the lower exponential strong well-posedness: for all $\zeta \in \Xi \backslash \Xi_{0}$, the operator $I-T_{0 \rightarrow 1}(\zeta) T_{1 \rightarrow 0}(\zeta): E^{u}(\zeta) \longrightarrow E^{u}(\zeta)$ is invertible.

However, on the other hand, by identification on $E^{s}(\zeta)$ and ker $B_{0}$ (we recall that the associated projectors are defined in Definition 2.1) (37) reads:

$$
\left\{\begin{array}{l}
\bar{\Pi}^{s}(\zeta) u_{\mid x_{d}=0}=\phi_{0}(\zeta) g  \tag{41}\\
\bar{\Pi}_{\mathrm{ker} B_{0}}(\zeta) u_{\mid x_{d}=0}=\left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) u_{\mid x_{d}=0}
\end{array}\right.
$$

because $\operatorname{Ran}\left(I-\phi_{0}(\zeta) B_{0}\right) \subset$ ker $B_{0}$. The first equation of 41$)$ uniquely determines the stable part of the trace on $\partial \Omega_{0}$ and we shall make the second one more explicit in terms of $\bar{\Pi}_{\text {ker } B_{0}}(\zeta) u_{\mid x_{d}=0}(\zeta)$. To this aim we decompose:

$$
\Pi^{s}(\zeta) u_{\mid x_{d}=0}(\zeta)=\underbrace{\Pi^{s}(\zeta) \bar{\Pi}^{s}(\zeta)}_{=\bar{\Pi}^{s}(\zeta)} u_{\mid x_{d}=0}(\zeta)+\Pi^{s}(\zeta) \bar{\Pi}_{\text {ker } B_{0}}(\zeta) u_{\mid x_{d}=0}(\zeta)
$$

in the right hand side of the second equation of 41. However for all $X \in \mathbb{C}^{N}$ we have the identity:

$$
e^{-\mathscr{A}(\zeta)}\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}(\zeta)} X=e^{-\mathscr{A}(\zeta)} \underbrace{\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}(\zeta)} \Pi^{u}(\zeta)}_{=0} X+e^{-\mathscr{A}(\zeta)}\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) X,
$$

because both spaces $E^{s}(\zeta)$ and $E^{u}(\zeta)$ are stable subspaces for $e^{\mathscr{A}(\zeta)}$. So we deduce that the contribution of $\Pi^{s}(\zeta) \bar{\Pi}_{\text {ker } B_{0}}(\zeta) u_{\mid x_{d}=0}(\zeta)$ in the second equation of 41$)$ is:

$$
\begin{aligned}
& \left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) \bar{\Pi}_{\operatorname{ker} B_{0}}(\zeta) u_{\mid x_{d}=0}(\zeta) \\
& \quad=\left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)}\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}(\zeta)} \bar{\Pi}_{\mathrm{ker} B_{0}}(\zeta) u_{\mid x_{d}=0}(\zeta)-\underbrace{\left(I-\phi_{0}(\zeta) B_{0}\right) \Pi^{s}(\zeta)}_{=0} \bar{\Pi}_{\mathrm{ker} B_{0}}(\zeta) u_{\mid x_{d}=0}
\end{aligned}
$$

Consequently the second equation of (41) reads:

$$
\begin{equation*}
\left(I-\bar{T}_{1 \rightarrow 0}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right) \bar{\Pi}_{\operatorname{ker} B_{0}}(\zeta) u_{\mid x_{d}=0}(\zeta)=\left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \phi_{0}(\zeta) g \tag{42}
\end{equation*}
$$

where we set:

$$
\begin{equation*}
\bar{T}_{1 \rightarrow 0}(\zeta):=\left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)} \text { and } \bar{T}_{0 \rightarrow 1}(\zeta):=\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}(\zeta)} \tag{43}
\end{equation*}
$$

and where we used the first equation of (41) to express the right hand side of (42) in terms of $g$. In view of (42) (recall that $\left(I-\phi_{0}(\zeta) B_{0}\right) E^{u}(\bar{\zeta})=\left(I-\phi_{0}(\zeta) B_{0}\right) \mathbb{C}^{N}=$ ker $B_{0}$, we assume that the operator $\left(I-\bar{T}_{1 \rightarrow 0}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right)$ is invertible on ker $B_{0}$ with values in ker $B_{0}$ uniformly in terms of $\zeta \in \Xi \backslash \Xi_{0}$.

Once again, if one reiterates exactly the same reasoning but this time with a non zero source term on the side $\left\{x_{d}=1\right\}$ instead of the side $\left\{x_{d}=0\right\}$, then he will recover the following necessary condition for the strong well-posedness: for all $\zeta \in \Xi \backslash \Xi_{0}$, the operator $I-\bar{T}_{0 \rightarrow 1}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta): \operatorname{ker} B_{1} \longrightarrow \operatorname{ker} B_{1}$ is invertible, uniformly in terms of $\zeta$.

For simplicity let us define:

$$
T_{0 \rightarrow 0}(\zeta):=T_{1 \rightarrow 0}(\zeta) T_{0 \rightarrow 1}(\zeta) \text { and } T_{1 \rightarrow 1}(\zeta):=T_{0 \rightarrow 1}(\zeta) T_{1 \rightarrow 0}(\zeta)
$$

where $T_{0 \rightarrow 1}(\zeta)$ and $T_{1 \rightarrow 0}(\zeta)$ are defined in (38) and

$$
\bar{T}_{0 \rightarrow 0}(\zeta):=\bar{T}_{1 \rightarrow 0}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) \text { and } \bar{T}_{1 \rightarrow 1}(\zeta):=\bar{T}_{0 \rightarrow 1}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)
$$

where $\bar{T}_{0 \rightarrow 1}(\zeta)$ and $\bar{T}_{1 \rightarrow 0}(\zeta)$ are defined in 43).
We summarize this paragraph in the following proposition:
Proposition 3.5 If the strip problem (32) is strongly well posed then for all $\zeta \in \Xi \backslash \Xi_{0}$, the operator $\left(I-T_{0 \rightarrow 0}(\zeta)\right)\left(\right.$ resp. $\left.\left(I-T_{1 \rightarrow 1}(\zeta)\right)\right)$ is an isomorphism from $E^{s}(\zeta)$ to $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ to $E^{u}(\zeta)$ ). Moreover we assume the bounds: for all $\zeta \in \Xi \backslash \Xi_{0}$;

$$
\forall v \in E^{s}(\zeta),|v| \leq C_{0}\left|\left(I-T_{0 \rightarrow 0}(\zeta)\right) v\right|
$$

and

$$
\forall w \in E^{u}(\zeta),|w| \leq C_{1}\left|\left(I-T_{1 \rightarrow 1}(\zeta)\right) w\right|
$$

where $C_{0}, C_{1}>0$ do not depend on $\zeta$.
If the strip problem (32) is strongly well posed then for all $\zeta \in \Xi \backslash \Xi_{0}$, the operator $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ (resp. $\left.\left(I-\bar{T}_{1 \rightarrow 1}(\zeta)\right)\right)$ is an isomorphism from $\operatorname{ker} B_{0}$ to $\operatorname{ker} B_{0}$ (resp. $\operatorname{ker} B_{1}$ to $\operatorname{ker} B_{1}$ ). Moreover we assume the bounds: for all $\zeta \in \Xi \backslash \Xi_{0}$;

$$
\forall v \in \operatorname{ker} B_{0},|v| \leq \bar{C}_{0}\left|\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right) v\right|
$$

and

$$
\forall w \in \operatorname{ker} B_{1},|w| \leq \bar{C}_{1}\left|\left(I-\bar{T}_{1 \rightarrow 1}(\zeta)\right) w\right|
$$

where $\bar{C}_{0}, \bar{C}_{1}>0$ do not depend on $\zeta$.
The end of this section is devoted to the proof of the fact that these necessary conditions are also sufficient conditions for the strip problem $\sqrt{32}$ to be lower exponentially strongly well posed.

### 3.3 Sufficient conditions for lower exponential strong well-posedness via symmetrizors

We first make the following assumptions:
Assumption 3.1 For all $\zeta \in \Xi \backslash \Xi_{0}$ we assume that the operator ( $\left.I-T_{0 \rightarrow 0}(\zeta)\right)$ (resp. $\left(I-T_{1 \rightarrow 1}(\zeta)\right)$ ) is uniformly invertible (in terms of $\zeta$ ) from $E^{s}(\zeta)$ to $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ to $E^{u}(\zeta)$ ).

Assumption 3.2 For all $\zeta \in \Xi \backslash \Xi_{0}$ we assume that the operator $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ (resp. $\left(I-\bar{T}_{1 \rightarrow 1}(\zeta)\right)$ ) is uniformly invertible (in terms of $\zeta$ ) from $\operatorname{ker} B_{0}$ to $\operatorname{ker} B_{0}$ (resp. $\operatorname{ker} B_{1}$ to $\operatorname{ker} B_{1}$ ).

These assumptions has been shown to be necessary for the lower exponential strong well-posedness of (11) in Paragraph 3.2.

Before to turn to the proof of their sufficiency, let us describe a link between these two assumptions. More precisely we show in the following proposition that Assumption 3.1 implies Assumption 3.24]

Proposition 3.6 If the operator $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ is invertible over $E^{s}(\zeta)$ then the operator $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ is invertible over ker $B_{0}$.

Proof : Consider $v \in \mathbb{C}^{N}$ such that $v=\bar{T}_{0 \rightarrow 0}(\zeta) v$, then $v \in$ ker $B_{0}$ and we decompose $v=\Pi^{s}(\zeta) v+\Pi^{u}(\zeta) v$. We compute:

$$
\begin{aligned}
\bar{T}_{0 \rightarrow 0}(\zeta) v= & \left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)}\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) v, \\
& =\underbrace{\left(I-\phi_{0}(\zeta) B_{0}\right) \Pi^{s}(\zeta)}_{=0} v-\left(I-\phi_{0}(\zeta) B_{0}\right) e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) v, \\
& =-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) v+T_{0 \rightarrow 0}(\zeta) \Pi^{s}(\zeta) v .
\end{aligned}
$$

Identifying on $E^{s}(\zeta)$ and $E^{u}(\zeta)$ we derive:

$$
\left\{\begin{array}{l}
\left(I-T_{0 \rightarrow 0}(\zeta)\right) \Pi^{s}(\zeta) v=0, \\
\Pi^{u}(\zeta) v=-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \Pi^{s}(\zeta) v .
\end{array}\right.
$$

By assumption we have $\Pi^{s}(\zeta) v=\Pi^{u}(\zeta) v=0$ and consequently $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ is invertible over ker $B_{0}$.

Consequently Proposition 3.6 is a strong argument in favor of the fact that Assumption 3.1 implies 3.2 To show such an implication only the uniform bound of the inverse has to be considered and it is left for future studies.

### 3.3.1 Outline of the proof, and comments about Assumptions 3.1-3.2

As in the proof of the lower exponential strong well-posedness in the strictly dissipative setting the sketch of proof that we follow is a classical proof in three steps. More precisely, the first step is to obtain an $a$ priori energy estimate for all regular enough solution of (1). This first step uses Assumption 3.2 but not Assumption 3.1.

Then in a second step we will construct a weak solution of (17. This can be done by two distinct ways: the first one is to construct this weak solution "by hand" and will be done under a stronger assumption than Assumption 3.1, while the second one will use the dual problem (as it has been done in Paragraph 3.1.2.). More precisely, we will introduce a dual problem for (11), show the uniqueness of its solution from which we immediately deduce the existence of a weak solution for the primal problem.

The simplest way to show that the dual problem admits a unique solution is of course to show that it satisfies the a priori energy estimate and, from the first step of the proof, it will be sufficient that Assumption 3.2 holds for the dual problem. The main point in this proof is the quite surprising fact that Assumption

[^1]3.2 for the dual problem is nothing but Assumption 3.1 for the primal problem.

Consequently assume that the primal problem satisfies both Assumptions 3.1 and 3.2 then we have the a priori energy estimate and the existence of a weak solution. The "weak=strong" lemma used for the problem with strictly dissipative boundary conditions ends the proof of Theorem 2.2

### 3.3.2 The a priori energy estimate via symmetrizors

The proof relies on several substeps and follow essentially the ideas developed by Osher, 1973 to establish a priori energy estimate for the corner problem. Let us list the main arguments :
i) The first substep consists in a reduction to a homogeneous problem in the interior and used the fact that the symmetric problem with strictly dissipative boundary conditions is lower exponentially strongly well-posed (see Theorem 3.1).
ii)Then we show a priori energy estimate for the reduced problem thanks to symmetrizors. This adopt was already used by Kreiss, 1970 to treat the problem in the half space.
iii) Finally the last substep is to show that such symmetrizors exist. This point of the proof of the $a$ priori energy estimate is the only one that uses Assumption 3.2.

Step 1: Reduction to a homogeneous problem in the interior. The purpose of this paragraph is to show the following:

Proposition 3.7 (Reduction to an homogeneous problem in the interior) Under Assumptions 2.1. 2.2 and 2.3 consider the strip problem with homogeneous forcing term in the interior, that is

$$
\left\{\begin{array}{l}
\left.\frac{d}{d x_{d}} v=\mathscr{A}(\zeta) v \quad \text { for } x_{d} \in\right] 0,1[  \tag{44}\\
B_{0} v_{\mid x_{d}=0}=g_{0}(\zeta) \\
B_{1} v_{\mid x_{d}=1}=g_{1}(\zeta)
\end{array}\right.
$$

then the resolvent strip problem (8) is lower exponentially strongly well-posed if and only if the resolvent strip corner problem with homogeneous interior forcing term (44) is lower exponentially strongly well-posed in the sense that it admits a unique solution $v \in L^{2}([0,1])$, satisfying the energy estimate:

$$
\begin{equation*}
\left|v_{\mid x_{d}=0}\right|^{2}+\left|v_{\mid x_{d}=1}\right|^{2} \leq C\left(\left|g_{0}\right|^{2}+\left|g_{1}\right|^{2}\right) \tag{45}
\end{equation*}
$$

where we stress that $C$ does not depend on $\zeta$.
Proof : We only have to show that the lower exponential strong well-posedness of 44 implies the one of (8). To this aim we introduce two strictly dissipative boundary conditions $\widetilde{B}_{0}$ and $\widetilde{B}_{1}$ (note that such boundary conditions exist from Definition-Proposition 3.1 and we consider the auxiliary problem:

$$
\left\{\begin{array}{l}
\left.\frac{\mathrm{d}}{\mathrm{~d} x_{d}} w\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) w+A_{d}^{-1} f\left(\zeta, x_{d}\right) \quad \text { for } x_{d} \in\right] 0,1[  \tag{46}\\
\widetilde{B}_{0} w_{\mid x_{d}=0}(\zeta)=0 \\
\widetilde{B}_{1} w_{\mid x_{d}=1}(\zeta)=0
\end{array}\right.
$$

as the boundary conditions are strictly dissipative Theorem 3.1 applies that is there exists a unique solution $w \in L^{2}([0,1])$ of 46$)$. Moreover $w$ satisfies the energy estimate: there exists $C>0$ such that for all $\gamma>0$,

$$
\begin{equation*}
\gamma\|w\|_{L^{2}([0,1])}^{2}+\left|w_{\mid x_{d}=0}\right|^{2}+\left|w_{\mid x_{d}=1}\right|^{2} \leq \frac{C}{\gamma}\|f\|_{L^{2}([0,1])}^{2} . \tag{47}
\end{equation*}
$$

Let $v$ be the unique solution of (44), as we assume lower exponential strong well-posedness, we have, on the one hand (see 45):

$$
\begin{equation*}
\left|v_{\mid x_{d}=0}\right|^{2}+\left|v_{\mid x_{d}=1}\right|^{2} \leq C\left(\left|g_{0}\right|^{2}+\left|g_{1}\right|^{2}\right) . \tag{48}
\end{equation*}
$$

But on the other hand, test the evolution equation of (44) against $v$, integrate over $[0,1]$ and finally taking the real part gives the estimate (see the proof of Proposition 3.2):

$$
\begin{equation*}
\gamma\|v\|_{L^{2}([0,1])}^{2} \leq\left|v_{\mid x_{d}=0}\right|^{2}+\left|v_{\mid x_{d}=1}\right|^{2} \tag{49}
\end{equation*}
$$

Let $u$ be defined by $u:=w+v$ then $u \in L^{2}([0,1])$ is a solution of the strip problem:

$$
\left\{\begin{array}{l}
\left.\frac{\mathrm{d}}{\mathrm{~d} x_{d}} u=\mathscr{A}(\zeta) u+A_{d}^{-1} f\left(\zeta, x_{d}\right) \quad \text { for } x_{d} \in\right] 0,1[ \\
B_{0} u_{\mid x_{d}=0}=\widetilde{g}_{0}(\zeta) \\
B_{1} u_{\mid x_{d}=1}=\widetilde{g}_{1}(\zeta)
\end{array}\right.
$$

where the source terms on the boundaries are in $\mathbb{C}^{N}$ and are defined by $\widetilde{g}_{j}:=g_{j}+B_{j} w_{\mid x_{d}=j}, j=0,1$. Finally, from the energy estimates (47), (48), 49) and the triangle inequality we deduce that:

$$
\begin{aligned}
\gamma\|u\|_{L^{2}([0,1])}^{2}+\left|u_{\mid x_{d}=0}\right|^{2}+\left|u_{\mid x_{d}=1}\right|^{2} \leq & \gamma\|v\|_{L^{2}([0,1])}^{2}+\left|v_{\mid x_{d}=0}\right|^{2}+\left|v_{\mid x_{d}=1}\right|^{2} \\
& +\gamma\|w\|_{L^{2}([0,1])}^{2}+\left|w_{\mid x_{d}=0}\right|^{2}+\left|w_{\mid x_{d}=1}\right|^{2} \\
\leq & C\left(\frac{1}{\gamma}\|f\|_{L^{2}([0,1])}^{2}+\left|g_{0}\right|^{2}+\left|g_{1}\right|^{2}\right)
\end{aligned}
$$

which concluded the proof of the strong well-posedness of (8) (the uniqueness follows from the energy estimate).

Step 2: Definition of the symmetrizor and last reduction. In this second substep we show that the energy estimate $\sqrt{19}$ holds for any regular (let us say $\left.H^{1}([0,1])\right)$ solution of 88 . To show this result we will need the following object: the so-called symmetrizor ${ }^{3}$.

Definition 3.3 (Symmetrizor) We say that $\mathscr{S}: \Xi \backslash \Xi_{0} \times[0,1] \rightarrow \mathbf{M}_{N \times N}(\mathbb{C})$ is a symmetrizor for the resolvent strip problem (8) if it satisfies the following conditions:
i) For all $\zeta \in \Xi \backslash \Xi_{0}$ and for all $x_{d} \in[0,1]$, the matrix $\Sigma\left(\zeta, x_{d}\right):=\mathscr{S}\left(\zeta, x_{d}\right) A_{d}$ is hermitian.
ii) For all $\zeta \in \Xi \backslash \Xi_{0}, \Sigma(\zeta, 0)$ (resp. $\Sigma(\zeta, 1)$ ) is bounded, uniformly in terms of $\zeta$ on $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ ). In other words there exists $c_{0}, c_{1}>0$ such that for all $\zeta \in \Xi \backslash \Xi_{0}$ the following bounds holds:

$$
\forall v \in E^{s}(\zeta),|\Sigma(\zeta, 0) v| \leq c_{0}|v| \text { and } \forall w \in E^{u}(\zeta),|\Sigma(\zeta, 1) w| \leq c_{0}|w|
$$

iii) For all $\zeta \in \Xi \backslash \Xi_{0}, \mathscr{S}(\zeta, \cdot)$ solves the differential equation:

$$
\left.\frac{1}{2} \frac{d}{d x_{d}}\left(\mathscr{S}\left(\zeta, x_{d}\right)\right) A_{d}+\operatorname{Re}\left(\mathscr{S}\left(\zeta, x_{d}\right) M(\zeta)\right)=0, \text { for } x_{d} \in\right] 0,1[
$$

where we recall that $M(\zeta)=\left[\sigma+i A^{\prime}(\eta)\right]$.
iv) For $x_{d}=0$, for all $v \in \mathbb{C}^{N}$ and for all $\zeta \in \Xi \backslash \Xi_{0}$ we have the inequality:

$$
\langle v, \mathscr{S}(\zeta, 0) v\rangle \leq C_{0}\left|B_{0} v\right|^{2}-\varepsilon_{0}|v|^{2}
$$

for some constants $\varepsilon_{0}, C_{0}>0$ that do not depend on $\zeta$.
$v)$ For $x_{d}=1$, for all $w \in \mathbb{C}^{N}$ and for all $\zeta \in\left(\Xi \backslash \Xi_{0}\right)$ we have the inequality:

$$
-\langle w, \mathscr{S}(\zeta, 1) w\rangle \leq C_{1}\left|B_{1} w\right|^{2}-\varepsilon_{1}|w|^{2}
$$

for some constants $\varepsilon_{1}, C_{1}>0$ that do not depend on $\zeta$.

[^2]In a first time we assume that there exists a symmetrizor $\mathscr{S}$ for (8) and we conclude this substep by showing that the a priori energy estimate 18 is satisfied. The construction of a symmetrizor is postnone to the following substep.

Proposition 3.8 ( $\boldsymbol{A}$ priori energy estimate) Under Assumptions 2.1, and 2.2, and 2.3 if the resolvent strip problem (8) admits a symmetrizor in the sense of Definition 3.3 then the a priori energy estimate (18) is satisfied by all regular enough solution $u$ of (8).

Proof : From Proposition 3.7 we can assume that $f \equiv 0$, so that we only have to show (18) without the forcing term $f$ in the right hand side and the $L^{2}$-norm of the solution in the left hand side that is:

$$
\left|u_{\mid x_{d}=0}\right|^{2}+\left|u_{\mid x_{d}=1}\right|^{2} \leq C\left(\left|g_{0}\right|^{2}+\left|g_{1}\right|^{2}\right)
$$

where $u$ is a regular solution of 44 or equivalently of:

$$
\left\{\begin{array}{l}
\left.A_{d} \frac{\mathrm{~d}}{\mathrm{~d} x_{d}} u+M(\zeta) u=0 \quad \text { for } x_{d} \in\right] 0,1[ \\
B_{0} u_{\mid x_{d}=0}=g_{0}(\zeta) \\
B_{1} u_{\mid x_{d}=1}=g_{1}(\zeta)
\end{array}\right.
$$

We multiply the evolution equation of this system by $\mathscr{S}\left(\zeta, x_{d}\right)$ and test against $u$ to derive that:

$$
\begin{equation*}
\left\langle\Sigma\left(\zeta, x_{d}\right) \frac{\mathrm{d}}{\mathrm{~d} x_{d}} u, u\right\rangle+\left\langle\mathscr{S}\left(\zeta, x_{d}\right) M(\zeta) u, u\right\rangle=0 \tag{50}
\end{equation*}
$$

Then we use the relation

$$
\frac{\mathrm{d}}{\mathrm{~d} x_{d}}\left\langle\Sigma\left(\zeta, x_{d}\right) u, u\right\rangle=\left\langle\frac{\mathrm{d}}{\mathrm{~d} x_{d}}\left(\Sigma\left(\zeta, x_{d}\right)\right) u, u\right\rangle+2\left\langle\Sigma\left(\zeta, x_{d}\right) \frac{\mathrm{d}}{\mathrm{~d} x_{d}} u, u\right\rangle,
$$

which holds because of point $i$ ) of Definition 3.3 in 50 and we obtain:

$$
\begin{equation*}
\frac{1}{2} \frac{\mathrm{~d}}{\mathrm{~d} x_{d}}\left\langle\Sigma\left(\zeta, x_{d}\right) u, u\right\rangle-\frac{1}{2}\left\langle\frac{\mathrm{~d}}{\mathrm{~d} x_{d}}\left(\Sigma\left(\zeta, x_{d}\right)\right) u, u\right\rangle+\left\langle\mathscr{S}\left(\zeta, x_{d}\right) M(\zeta) u, u\right\rangle=0 \tag{51}
\end{equation*}
$$

From point $i i i$ ) of Definition 3.3, this equation is equivalent to:

$$
\frac{\mathrm{d}}{\mathrm{~d} x_{d}}\left\langle\Sigma\left(\zeta, x_{d}\right) u, u\right\rangle=0
$$

Integrate over $[0,1]$ gives:

$$
\left\langle\Sigma(\zeta, 1) u_{\mid x_{d}=1}, u_{\mid x_{d}=1}\right\rangle=\left\langle\Sigma(\zeta, 0) u_{\mid x_{d}=0}, u_{\mid x_{d}=0}\right\rangle
$$

Finally, using points $i v$ ) and $v$ ) of Definition 3.3 we obtain as desired:

$$
\varepsilon_{0}\left|u_{\mid x_{d}=0}\right|^{2}+\varepsilon_{1}\left|u_{\mid x_{d}=1}\right|^{2} \leq C_{0}\left|g_{0}\right|^{2}+C_{1}\left|g_{1}\right|^{2} .
$$

Let us remark that in this proof we do not use point $i i$ ) of Definition 3.3. Indeed, $i i$ ) of Definition 3.3 is in fact assumed to introduce a new notion of symmetrizor for which the control on the trace on $\partial \Gamma_{0}$ (resp. $\partial \Gamma_{1}$ ) is only imposed on ker $B_{0}$ (resp. ker $B_{1}$ ). More precisely we define the following notion of symmetrizor, the reduced symmetrizor:

Definition 3.4 [Reduced symmetrizor] We say that $\mathscr{S}:\left(\Xi \backslash \Xi_{0}\right) \times[0,1] \rightarrow \mathbf{M}(\mathbb{C})$ is a reduced symmetrizor for the resolvent strip problem (8) if it satisfies the conditions $i$ ) - iii) of Definition 3.3 and if in addition it satisfies:
iv $\left.{ }^{\prime}\right)$ For $x_{d}=0$, for all $v \in \operatorname{ker} B_{0}$ and for all $\zeta \in \Xi \backslash \Xi_{0}$ we have the inequality:

$$
\langle v, \mathscr{S}(\zeta, 0) v\rangle \leq-\varepsilon_{0}|v|^{2}
$$

for some constant $\varepsilon_{0}>0$ that does not depend on $\zeta$.
$\left.v^{\prime}\right)$ For $x_{d}=1$, for all $w \in \operatorname{ker} B_{1}$ and for all $\zeta \in \Xi \backslash \Xi_{0}$ we have the inequality:

$$
-\langle w, \mathscr{S}(\zeta, 1) w\rangle \leq-\varepsilon_{1}|w|^{2}
$$

for some constant $\varepsilon_{1}>0$ that does not depend on $\zeta$.
The main interest of this reduced symmetrizor is that thanks to point ii) of Definition 3.3 the control on the traces can be relaxed to the kernel of the boundary conditions while the reduced symmetrizor still enable to obtain a priori energy estimate. This result is given in the following Proposition

Proposition 3.9 Under Assumptions 2.1, 2.2 and 2.3 if the strip problem (8) admits a reduced symmetrizor in the sense of Definition 3.4 then the a priori energy estimate (45) (and consequently 18) is satisfied.

Proof : We reiterate the proof of Proposition 3.8 (note that this part of the proof only require $i$ ) and $i i i$ ) in Definition 3.3 to derive the equality

$$
\left\langle\Sigma(\zeta, 1) u_{\mid x_{d}=1}, u_{\mid x_{d}=1}\right\rangle=\left\langle\Sigma(\zeta, 0) u_{\mid x_{d}=0}, u_{\mid x_{d}=0}\right\rangle .
$$

Using point $i v^{\prime}$ ) of Definition 3.4 for all $u^{b} \in \operatorname{ker} B_{0}$, we have $\left\langle u^{b}, \mathscr{S}(0, \zeta) u^{b}\right\rangle \leq-\varepsilon_{0}\left|u^{b}\right|^{2}$. By the uniform Kreiss-Lopatinskii condition we decompose $u_{x_{d}=0}=u^{b}+u^{\sharp}$, where $u^{b} \in \operatorname{ker} B_{0}$ and $u^{\sharp} \in E^{s}(\zeta), B_{0} u^{\sharp}=g_{0}$.

Then because of point $i$ ) in Definition 3.3 we have:

$$
\begin{aligned}
\left\langle\Sigma(\zeta, 0) u_{\mid x_{d}=0}, u_{\mid x_{d}=0}\right\rangle & =\left\langle\Sigma(\zeta, 0) u^{b}, u^{b}\right\rangle+2 \operatorname{Re}\left\langle\Sigma(\zeta, 0) u^{b}, u^{\sharp}\right\rangle+\left\langle u^{\sharp}, \Sigma(\zeta, 0) u^{\sharp}\right\rangle, \\
& \leq-\frac{\varepsilon_{0}}{2}\left|u^{b}\right|^{2}+C_{\varepsilon_{0}}\left|\Sigma(\zeta, 0) u^{\sharp}\right|^{2}+\left|u^{\sharp}\right|^{2},
\end{aligned}
$$

where we used point $i v^{\prime}$ ) of Definition 3.4 in the first term of the right hand side, Cauchy-Schwarz and Young inequalities in the second and third terms. From point $i i$ ) in Definition 3.3 we obtain:

$$
\left\langle\Sigma(\zeta, 0) u_{\mid x_{d}=0}, u_{\mid x_{d}=0}\right\rangle \leq-\frac{\varepsilon_{0}}{2}\left|u^{b}\right|^{2}+C\left|u^{\sharp}\right|^{2},
$$

where we stress that the constant $C$ in the right hand side may depend on $\varepsilon_{0}$ but does not depend on $\zeta$.
Finally by the uniform Kreiss-Lopatinskii condition for the side $\partial \Gamma_{0}$ we obtain:

$$
\left\langle\Sigma(\zeta, 0) u_{\mid x_{d}=0}, u_{\mid x_{d}=0}\right\rangle \leq-\frac{\varepsilon_{0}}{2}\left|u^{b}\right|^{2}+C\left|B_{0} u^{\sharp}\right|^{2},
$$

which is nothing but the inequality $i v$ ) of Definition 3.3. The derivation of the inequality $v$ ) of Definition 3.3 follows exactly the same lines and will be ommited here.

We then conclude exactly as in the end of the proof of Proposition 3.8.

With Proposition 3.9 in hand, to show the a priori energy estimate for the strip problem (1) it only remains to construct a reduced symmetrizor in the sense of Definition 3.4. This construction occupies the end of this subsection.

Step 3: Construction of a reduced symmetrizor.
The construction of a reduced symmetrizor for the strip problem has to be compared with the one given by Osher, 1973 to treate corner problems. In the author opinion, this fact is interesting in its own because it shows that the symmetrizor is not linked to the geometry of the space in which the problem lies but that it is constructed to deal with autointeraction phenomena which are proper to boundary value problems with
several boundaries.
We are looking for a reduced symmetrizor, in the sense of Definition 3.4, under the form:

$$
\begin{equation*}
\mathscr{S}\left(\zeta, x_{d}\right):=A_{d} J\left(\zeta, x_{d}\right) N(\zeta) J^{*}\left(\zeta, x_{d}\right) \tag{52}
\end{equation*}
$$

A symmetrizor under this special form will of course satisfies the point $i$ ) of Definition 3.3 provided that

$$
\begin{equation*}
N^{*}(\zeta)=N(\zeta) \tag{53}
\end{equation*}
$$

The first substep of this construction is to choose $J$ in 52 in such a way that point $i i i$ ) of Definition 3.3 is achieved. As we will see, a way to determine such a $J$ involves the operators $\bar{T}_{0 \rightarrow 1}$ and $\bar{T}_{1 \rightarrow 0}$ defined in (43).

Then the second substep is to choose $N$ in (52) as a suitable weight to ensure points $i v^{\prime}$ ) and $v^{\prime}$ ) of Definition 3.4. An important remark for what follows is that because $N$ will not depend on the variable $x_{d}$, it will not destroy the fact that a symmetrizor reading under the form (52) will still satisfy point iii) in Definition 3.3. In some sense, $N$ will depend on $J$ (and consequently $N$ will depend on $\bar{T}_{0 \rightarrow 1}$ and $\bar{T}_{1 \rightarrow 0}$ ) but $J$ will not depend on $N$, so that we are have enough liberty to choose $N$ after having fixed $J$.

At last, we check that the constructed $\mathscr{S}$ satisfies point $i i$ ) of Definition 3.3.
Substep 3.1: Construction of $J$
A good candidate for $J$ is a solution of the equation: for all $\left.\zeta \in \Xi \backslash \Xi_{0}, x_{d} \in\right] 0,1[$

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} A_{d} J\left(\zeta, x_{d}\right)+M^{*}(\zeta) J\left(\zeta, x_{d}\right)=0 \tag{54}
\end{equation*}
$$

or equivalently (recall that $A_{d}$ is assumed to be symmetric)

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} J^{*}\left(\zeta, x_{d}\right) A_{d}+J^{*}\left(\zeta, x_{d}\right) M(\zeta)=0 \tag{55}
\end{equation*}
$$

Such a $J$ ensures that $\mathscr{S}$ reading under the form satisfies point iii) of Definition 3.3. Indeed we have:

$$
\begin{aligned}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} \mathscr{S}\left(\zeta, x_{d}\right) A_{d} & =\frac{\mathrm{d}}{\mathrm{~d} x_{d}} A_{d} J\left(\zeta, x_{d}\right) N(\zeta) J^{*}\left(\zeta, x_{d}\right) A_{d} \\
& =\frac{\mathrm{d}}{\mathrm{~d} x_{d}}\left(A_{d} J\left(\zeta, x_{d}\right)\right) N(\zeta) J^{*}\left(\zeta, x_{d}\right) A_{d}+A_{d} J\left(\zeta, x_{d}\right) N(\zeta) \frac{\mathrm{d}}{\mathrm{~d} x_{d}}\left(J^{*}\left(\zeta, x_{d}\right) A_{d}\right), \\
& =-M^{*}(\zeta) J\left(\zeta, x_{d}\right) N(\zeta) J^{*}\left(\zeta, x_{d}\right) A_{d}-A_{d} J\left(\zeta, x_{d}\right) N(\zeta) J^{*}\left(\zeta, x_{d}\right) M(\zeta), \\
& =-2 \operatorname{Re}\left(\mathscr{S}\left(\zeta, x_{d}\right) M(\zeta)\right)
\end{aligned}
$$

that is to say exactly $i i i$ ) of Definition 3.3 .
So it remains to construct $J$ as a solution of (54). Note that this equation is equivalent to

$$
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} J\left(\zeta, x_{d}\right)=\widetilde{\mathscr{A}}(\zeta) J\left(\zeta, x_{d}\right)
$$

where $\widetilde{\mathscr{A}}(\zeta)$ is the resolvent matrice of the dual problem of $\sqrt[12]{ }$ ) given by $\widetilde{\mathscr{A}}(\sigma, \eta)=\mathscr{A}(-\sigma, \eta)$ (see Paragraph 3.1.2). So that $J\left(\zeta, x_{d}\right)$ is a solution of the dual problem.

We choose $J$ under the form $J\left(\zeta, x_{d}\right):=K_{0}\left(\zeta, x_{d}\right)+K_{1}\left(\zeta, x_{d}\right)$ where $K_{0}$ and $K_{1}$ are solutions of:

$$
\forall \zeta \in \Xi \backslash \Xi_{0},\left\{\begin{array}{l}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} K_{0}\left(\zeta, x_{d}\right)=\widetilde{\mathscr{A}}(\zeta) K_{0}\left(\zeta, x_{d}\right) \quad \text { for } x_{d} \geq 0  \tag{56}\\
K_{0_{\mid x_{d}=0}}=\widetilde{\phi}_{0}(\zeta) C_{0} v
\end{array}\right.
$$

and

$$
\forall \zeta \in \Xi \backslash \Xi_{0},\left\{\begin{array}{l}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} K_{1}\left(\zeta, x_{d}\right)=\widetilde{\mathscr{A}}(\zeta) K_{1}\left(\zeta, x_{d}\right) \quad \text { for } x_{d} \leq 1,  \tag{57}\\
K_{1 \mid x_{d}=1}=\widetilde{\phi}_{1}(\zeta) C_{1} v,
\end{array}\right.
$$

where $v$ is a given element of $\mathbb{C}^{N}$ so that $K_{0}$ and $K_{1}$ are matrices. In (56)- (57) we recall that the terms $\widetilde{\phi}_{0}(\zeta)$ and $\widetilde{\phi}_{1}(\zeta)$ are the inverse given by the uniform Kreiss-Lopatinskii condition applied to the dual problem.

Such a choice of $J$ is perfectly possible because the sum $K_{0}+K_{1}$ is well-defined on $[0,1]$ and because no boundary conditions are imposed for $J$ in Definition 3.3.

Using Duhamel's formula we define $\widetilde{K_{0}}$ a solution of (56) and $\widetilde{K_{1}}$ a solution (57) by:

$$
\begin{equation*}
\widetilde{K_{0}}\left(\zeta, x_{d}\right) v=e^{x_{d} \widetilde{\mathscr{I}}(\zeta)} \widetilde{\phi_{0}}(\zeta) C_{0} v \tag{58}
\end{equation*}
$$

and

$$
\begin{equation*}
\widetilde{K_{1}}\left(\zeta, x_{d}\right) v=e^{\left(x_{d}-1\right) \widetilde{\mathscr{A}}(\zeta)} \widetilde{\phi_{1}}(\zeta) C_{1} v, \tag{59}
\end{equation*}
$$

So $\widetilde{J}:=\widetilde{K_{0}}+\widetilde{K_{1}}$ solves (54). However, in order to make the dependency of $\widetilde{J}(\zeta, 0)$ and $\widetilde{J}(\zeta, 1)$ explicit in terms of $\bar{T}_{0 \rightarrow 1}(\zeta)$ and $\bar{T}_{1 \rightarrow 0}(\zeta)$ we shall modify a bit the definition of $\widetilde{K_{0}}$ and $\widetilde{K}_{1}$. This is made in the following substep after a study of the boundary conditions $\left.i v^{\prime}\right)$ and $v^{\prime}$ ).

Substep 3.2: The adjoint operator $\widetilde{J}^{*}$
With a reduced symmetrizor $\mathscr{S}$ under the form (52), points $i v^{\prime}$ ) and $v^{\prime}$ ) of Definition 3.4 read:

$$
\left\{\begin{array}{l}
\forall v \in \operatorname{ker} B_{0},\left\langle v, A_{d} \widetilde{J}(\zeta, 0) N(\zeta) \widetilde{J}^{*}(\zeta, 0) A_{d} v\right\rangle \leq-\varepsilon_{0}|v|^{2}, \\
\forall w \in \operatorname{ker} B_{1},-\left\langle w, A_{d} \widetilde{J}(\zeta, 1) N(\zeta) \widetilde{J^{*}}(\zeta, 1) A_{d} w\right\rangle \leq-\varepsilon_{1}|w|^{2},
\end{array}\right.
$$

or equivalently

$$
\left\{\begin{array}{l}
\forall v \in \operatorname{ker} B_{0},\left\langle\widetilde{J}^{*}(\zeta, 0) A_{d} v, N(\zeta) \widetilde{J}^{*}(\zeta, 0) A_{d} v\right\rangle \leq-\varepsilon_{0}|v|^{2},  \tag{60}\\
\forall w \in \operatorname{ker} B_{1},-\left\langle\widetilde{J}^{*}(\zeta, 1) A_{d} w, N(\zeta) \widetilde{J^{*}}(\zeta, 1) A_{d} w\right\rangle \leq-\varepsilon_{1}|w|^{2}
\end{array}\right.
$$

So points $i v^{\prime}$ ) and $v^{\prime}$ ) of Definition 3.4 do not involve the operator $\widetilde{J}$, but only $\widetilde{J}^{*}$.
We have (recall the definitions of $\widetilde{K_{0}}$ and $\widetilde{K_{1}}$ (see (58) and (59)):

$$
\widetilde{J}^{*}(\zeta, 0)=\left(\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*}+\left(e^{-\tilde{\mathscr{A}}(\zeta)} \widetilde{\phi}_{1}(\zeta) C_{1}\right)^{*} \text { and } \widetilde{J}^{*}(\zeta, 1)=\left(e^{\tilde{\mathscr{A}}(\zeta)} \widetilde{\phi}_{0}(\zeta) C_{0}\right)^{*}+\left(\widetilde{\phi}_{1}(\zeta) C_{1}\right)^{*} \text {. }
$$

Next we use the following lemmas:
Lemma 3.4 Let $\widetilde{E}^{s}(\zeta)$ (resp. $\widetilde{E}^{u}(\zeta)$ be the stable (resp. unstable) subspace of $\widetilde{\mathscr{A}}(\zeta)$ then for all $\zeta \in \Xi$ we have the equalities:

$$
\widetilde{E}^{s}(\zeta)=\left(A_{d} E^{s}(\zeta)\right)^{\perp} \text { and } \widetilde{E}^{u}(\zeta)=\left(A_{d} E^{u}(\zeta)\right)^{\perp}
$$

Proof : We refer, for example, to [Benzoni-Gavage and Serre, 2007]-Paragraph 4.4] for a proof.
Lemma 3.5 For all $\zeta \in \Xi$ we have the equalities:

$$
\left(\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*}=A_{d}\left(I-\phi_{0}(\zeta) B_{0}\right) A_{d}^{-1} \text { and }\left(\widetilde{\phi_{1}}(\zeta) C_{0}\right)^{*}=A_{d}\left(I-\phi_{1}(\zeta) B_{1}\right) A_{d}^{-1} .
$$

Proof : We give the poof of the first equality, the proof of the second one follows exactly the same lines.
Firstly we show that:

$$
\left(\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*}=A_{d} \overline{\mathrm{\Pi}}_{\mathrm{ker} B_{0}}(\zeta) A_{d}^{-1}
$$

where we recall that $\bar{\Pi}_{\text {ker } B_{0}}(\zeta)$ is the projector upon ker $B_{0}$ with respect with the decomposition given by the uniform Kreiss-Lopatinskii condition on $\partial \Gamma_{0}$, that is $\mathbb{C}^{N}=\operatorname{ker} B_{0} \oplus E^{s}(\zeta)$. By definition of $\widetilde{\phi_{0}}(\zeta), \widetilde{\phi_{0}}(\zeta) C_{0}$
is a projector so it is uniquely determined by its the kernel and its range of $\left.\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*}$. We have by Lemma 3.4 .

$$
\begin{aligned}
\operatorname{ker}\left(\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*} & =\left(\operatorname{Ran} \widetilde{\phi_{0}}(\zeta) C_{0}\right)^{\perp}=\widetilde{E}^{s}(\zeta)^{\perp}=A_{d} E^{s}(\zeta), \\
\operatorname{Ran}\left(\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*} & =\left(\operatorname{ker} \widetilde{\phi_{0}}(\zeta) C_{0}\right)^{\perp}=\operatorname{ker} C_{0}^{\perp}=A_{d} \operatorname{ker} B_{0},
\end{aligned}
$$

where we used Lemma 3.4 to derive the first equality and 27 to derive the second one. Consequently we have:

$$
\left(\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*}=A_{d} \bar{\Pi}_{\text {ker } B_{0}}(\zeta) A_{d}^{-1}
$$

Then we remark that for all $X \in \mathbb{C}^{N}$ we have:

$$
\left.\left(I-\phi_{0}(\zeta) B_{0}\right)\right) X=\underbrace{\left(I-\phi_{0}(\zeta) B_{0}\right) \bar{\Pi}^{s}(\zeta) X}_{=0}+\left(I-\phi_{0}(\zeta) B_{0}\right) \bar{\Pi}_{\text {ker } B_{0}}(\zeta) X=\bar{\Pi}_{\text {ker } B_{0}}(\zeta) X
$$

from which we immediately deduce that

$$
\left(\widetilde{\phi_{0}}(\zeta) C_{0}\right)^{*}=A_{d}\left(I-\phi_{0}(\zeta) B_{0}\right) A_{d}^{-1}
$$

From Lemma 3.5 we deduce that

$$
\begin{aligned}
& \widetilde{J}^{*}(\zeta, 0) A_{d}=A_{d}\left[\left(I-\phi_{0}(\zeta) B_{0}\right)+\bar{T}_{0 \rightarrow 1}(\zeta)\right] \\
& \widetilde{J}^{*}\left(\zeta, 1 A_{d}\right)=A_{d}\left[\bar{T}_{1 \rightarrow 0}(\zeta)+\left(I-\phi_{1}(\zeta) B_{1}\right)\right]
\end{aligned}
$$

where we recall that $\bar{T}_{0 \rightarrow 1}(\zeta)$ and $\bar{T}_{1 \rightarrow 0}(\zeta)$ are defined in 43). Consequently to conclude the construction of $J$ we define $J$ such that $J^{*}\left(\zeta, x_{d}\right)=A_{d}^{-1} \widetilde{J}^{*}\left(\zeta, x_{d}\right)$, which satisfies (54). Moreover we have the following traces values:

$$
\begin{align*}
& J^{*}(\zeta, 0) A_{d}=\left(I-\phi_{0}(\zeta) B_{0}\right)+\bar{T}_{0 \rightarrow 1}(\zeta)  \tag{61}\\
& J^{*}(\zeta, 1) A_{d}=\bar{T}_{1 \rightarrow 0}(\zeta)+\left(I-\phi_{1}(\zeta) B_{1}\right) \tag{62}
\end{align*}
$$

Once $J$ is fixed, it remains to choose the operator $N$ to ensure (written with $J$ instead of $\widetilde{J}$ ). Before to give a precise construction for $N$, let us be a little more precise about these boundary conditions. Indeed, we remark that the first equation of 60 is assumed to hold for $v \in \operatorname{ker} B_{0}$, the second one for $w \in \operatorname{ker} B_{1}$ so that we in fact have:

$$
\begin{equation*}
J^{*}(\zeta, 0) A_{d} v=\left(I+\bar{T}_{0 \rightarrow 1}(\zeta)\right) v, \text { and } J^{*}(\zeta, 1) A_{d} w=\left(\bar{T}_{1 \rightarrow 0}(\zeta)+I\right) w \tag{63}
\end{equation*}
$$

The next substep of the proof is devoted to the construction of $N$ to make sure that 60 holds.
Substep 3.3: Construction of $N$.
We are looking for $N$ under the form:

$$
\begin{equation*}
N(\zeta):=\Pi_{\text {ker } B_{0}}^{*} N_{00}(\zeta) \Pi_{\text {ker } B_{0}}+\Pi_{\text {ker } B_{1}}^{*} N_{11}(\zeta) \Pi_{\text {ker } B_{1}}+2 \operatorname{Re}\left(\Pi_{\text {ker } B_{0}}^{*} N_{01}(\zeta) \Pi_{\text {ker } B_{1}}\right) \tag{64}
\end{equation*}
$$

where we recall that $\Pi_{\text {ker } B_{0}}$ (resp. $\Pi_{\text {ker } B_{1}}$ ) is the projector upon ker $B_{0}$ (resp. ker $B_{1}$ ) with respect with the decompositior ${ }^{4} \mathbb{R}^{N}=\operatorname{ker} B_{0} \oplus \operatorname{ker} B_{1}$. Note that such a $N$ satisfies (53) independently of $N_{00}, N_{11}$ and $N_{01}$ so that $\mathscr{S}$ defined in (52) satisfies point $i$ ) of Definition 3.3 .

To make the first equation of (60) more explicit we shall compute $N(\zeta) J(\zeta, 0)^{*} A_{d} v$ for $v \in$ ker $B_{0}$. Because of (63) and (also recall that for all $\left.\zeta \in \Xi, \operatorname{Ran}\left(\bar{T}_{0 \rightarrow 1}(\zeta)\right) \subset \operatorname{ker} B_{1}\right)$ we obtain that for all $v \in \operatorname{ker} B_{0}$

$$
\begin{aligned}
N(\zeta) J(\zeta, 0)^{*} A_{d} v= & \Pi_{\text {ker } B_{0}}^{*} N_{00}(\zeta) v+\Pi_{\text {ker } B_{1}}^{*} N_{11}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) v \\
& -\Pi_{\text {ker } B_{0}}^{*} N_{01}(\zeta) T_{0 \rightarrow 1}(\zeta) v+\Pi_{\text {ker } B_{1}}^{*} N_{01}(\zeta)^{*} \Pi_{\text {ker } B_{0}} v
\end{aligned}
$$

[^3]so that the first equation of 60 becomes:
$$
\forall v \in \operatorname{ker} B_{0},\left\langle v,\left(N_{00}(\zeta)+\bar{T}_{0 \rightarrow 1}^{*}(\zeta) N_{11}(\zeta) T_{0 \rightarrow 1}(\zeta)+N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+\bar{T}_{0 \rightarrow 1}^{*}(\zeta) N_{01}^{*}(\zeta)\right) v\right\rangle \leq-\varepsilon_{0}|v|^{2}
$$

The same reasoning applied to the second equation of 60 gives:
$\forall w \in \operatorname{ker} B_{1}, \quad-\left\langle w,\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{00}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)+N_{11}(\zeta)+\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)+N_{01}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)\right) w\right\rangle \leq-\varepsilon_{1}|w|^{2}$.
In terms of operators these equations are equivalent to:

$$
\left\{\begin{array}{l}
N_{00}(\zeta)+\bar{T}_{0 \rightarrow 1}^{*}(\zeta) N_{11}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+2 \operatorname{Re}\left(N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right) \leq-I,  \tag{65}\\
-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{00}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)-N_{11}(\zeta)-2 \operatorname{Re}\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)\right) \leq-I,
\end{array}\right.
$$

where without loss of generality we assumed that $\varepsilon_{0}=\varepsilon_{1}=1$.
To solve 65 we essentially follow the proof given in Osher, 1973. Note that in 65 we have three degrees of freedom to solve two equations that is why the choice of the coefficients of $N(\zeta)$ is not unique. To determine good candidates we multiply the second equation of 65 by $\bar{T}_{0 \rightarrow 1}^{*}(\zeta)$ in the left hand side and by $\bar{T}_{0 \rightarrow 1}(\zeta)$ in the right hand side and then sum to the first equation of 65). We obtain

$$
\begin{equation*}
N_{00}(\zeta)-\bar{T}_{0 \rightarrow 0}^{*}(\zeta) N_{00}(\zeta) \bar{T}_{0 \rightarrow 0}(\zeta)+2 \operatorname{Re}\left(\left(I-\bar{T}_{0 \rightarrow 0}^{*}(\zeta)\right) N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right) \leq-\left(I+\bar{T}_{0 \rightarrow 1}^{*}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right) \tag{66}
\end{equation*}
$$

where we recall that $\bar{T}_{0 \rightarrow 0}(\zeta)=\bar{T}_{1 \rightarrow 0}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)$ and where we used the identity:

$$
\begin{aligned}
2 \operatorname{Re}\left(N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right) & -2 \bar{T}_{0 \rightarrow 1}^{*}(\zeta) \operatorname{Re}\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)\right) \bar{T}_{0 \rightarrow 1}^{*}(\zeta) \\
& =2 \operatorname{Re}\left(N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right)-2 \operatorname{Re}\left(\bar{T}_{0 \rightarrow 1}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right) \\
& =2 \operatorname{Re}\left(\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{*} N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right)
\end{aligned}
$$

Let formally ${ }^{5} N_{00}(\zeta):=-I$ and $N_{01}(\zeta):=\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-*} V(\zeta)$ and we choose $V(\zeta)$ in order to ensure the equality in (66). So we shall solve:

$$
V(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+\bar{T}_{0 \rightarrow 1}^{*}(\zeta) V^{*}(\zeta)+\bar{T}_{0 \rightarrow 0}^{*}(\zeta) \bar{T}_{0 \rightarrow 0}(\zeta)=-\bar{T}_{0 \rightarrow 1}^{*}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)
$$

and we verify that $V(\zeta):=-\frac{1}{2} \bar{T}_{0 \rightarrow 1}^{*}(\zeta)\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)+I\right)$ is suitable. Under this choice of $N_{00}(\zeta)$ and $N_{01}(\zeta)$ we now determine $N_{11}$ in such a way that 65 is satisfied.

Firstly, we consider the second equation of 65 in which we choose $N_{11}(\zeta)$ in order to impose the equality. It implies that:

$$
N_{11}(\zeta):=I-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}^{*}-2 \operatorname{Re}\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)\right)
$$

Secondly it remains to show that the first equation of 65 is satisfied under this choice of $N_{00}(\zeta), N_{01}(\zeta)$ and $N_{11}(\zeta)$. This equation is equivalent to

$$
\bar{T}_{0 \rightarrow 1}^{*}(\zeta)\left(I-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)-2 \operatorname{Re}\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)\right)\right) \bar{T}_{0 \rightarrow 1}(\zeta)+2 \operatorname{Re}\left(N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}^{*}(\zeta)\right) \leq 0
$$

which can be developed as

$$
\begin{aligned}
-\bar{T}_{0 \rightarrow 1}^{*}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+\bar{T}_{0 \rightarrow 0}^{*}(\zeta) \bar{T}_{0 \rightarrow 0}(\zeta)-\bar{T}_{0 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) & -\bar{T}_{0 \rightarrow 1}^{*}(\zeta) N_{01}^{*}(\zeta) \bar{T}_{0 \rightarrow 0}(\zeta) \\
& +N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+\bar{T}_{0 \rightarrow 1}^{*}(\zeta) N_{01}^{*}(\zeta) \leq 0
\end{aligned}
$$

that is

$$
\begin{equation*}
-\bar{T}_{0 \rightarrow 1}^{*}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+\bar{T}_{0 \rightarrow 0}^{*}(\zeta) \bar{T}_{0 \rightarrow 0}(\zeta)+2 \operatorname{Re}\left(\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{*} N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)\right) \leq 0 \tag{67}
\end{equation*}
$$

[^4]From the definition of $N_{01}(\zeta)$ equation b7) becomes:

$$
\begin{aligned}
& -\bar{T}_{0 \rightarrow 1}^{*}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+\bar{T}_{0 \rightarrow 0}^{*}(\zeta) \bar{T}_{0 \rightarrow 0}(\zeta)+V(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta)+\bar{T}_{0 \rightarrow 1}^{*}(\zeta) V^{*}(\zeta) \leq 0, \\
\Leftrightarrow & -2 \bar{T}_{0 \rightarrow 1}^{*}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) \leq 0
\end{aligned}
$$

Consequently with

$$
\begin{gather*}
N_{00}:=-I, N_{01}(\zeta):=-\frac{1}{2}\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-*} \bar{T}_{0 \rightarrow 1}^{*}(\zeta)\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)+I\right)  \tag{68}\\
\text { and } N_{11}(\zeta):=I-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)-2 \operatorname{Re}\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)\right),
\end{gather*}
$$

the symmetrizor $\mathscr{S}$ defined in (52), with $J$ characterized as a solution of (54) and $N$ under the form 64 with coefficients defined in (68) solves, at the formal level, points $i$ ) and $i i i$ ) of Definition 3.3 as well as $i v^{\prime}$ ) and $v^{\prime}$ ) of Definition 3.4 .

Substep 3.4: Justification and proof of point ii) in Definition 3.3
To show that $\mathscr{S}$ is a reduced symmetrizor it remains to show that it is well-defined, which amounts to show that $N$ is well-defined, and that it also satisfies $i i$ ) of Definition 3.3

Firstly we justify that with the coefficients defined in 68), $N$ is well-defined under Assumption 3.2 . By definition, we have:

$$
\begin{align*}
N(\zeta)= & -\Pi_{\text {ker } B_{0}}^{*} \Pi_{\text {ker } B_{0}}+\Pi_{\text {ker } B_{1}}^{*}\left(I-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)-N_{01}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)\right) \Pi_{\text {ker } B_{1}}  \tag{69}\\
& +\Pi_{\text {ker } B_{0}}^{*} N_{01}(\zeta) \Pi_{\text {ker } B_{1}}+\Pi_{\text {ker } B_{1}}^{*} N_{01}^{*}(\zeta) \Pi_{\text {ker } B_{0}}
\end{align*}
$$

so that to ensure that $N(\zeta)$ is well-defined we have to make sure that each terms involving $N_{01}(\zeta)$ or $N_{01}^{*}(\zeta)$ in 69) are well-defined under Assumption 3.2

Concerning the terms involving $N_{01}^{*}(\zeta)$, using the fact that $\operatorname{Ran} \bar{T}_{1 \rightarrow 0}(\zeta) \subset$ ker $B_{0}$, we only have to justify that $N_{01}^{*}(\zeta) v$ makes sense for all $v \in \operatorname{ker} B_{0}$. It is trivially the case under Assumption 3.2 because we have:

$$
N_{01}^{*}(\zeta) v=V^{*}(\zeta)\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-1} v
$$

Then we remark that the only terms involving $N_{01}(\zeta)$ in 69 are the terms $\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)$ or $\Pi_{\text {ker } B_{0}}^{*} N_{01}(\zeta)$. These terms can be rewritten under the form:

$$
\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)=\left(\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-1} \bar{T}_{1 \rightarrow 0}(\zeta)\right)^{*} V \text { and } \Pi_{\text {ker } B_{0}}^{*} N_{01}(\zeta)=\left(\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-1} \Pi_{\text {ker } B_{0}}\right)^{*} V
$$

and consequently they are well-defined under Assumption 3.2 .
So to show that the symmetrizor $\mathscr{S}$ defined in (52) is a reduced symmetrizor in the sense of Definition 3.4 it remains to show that it satisfies point $i i$ ) of Definition 3.3. We recall that point $i i$ ) of Definition 3.3 means that there exists $c_{0}, c_{1}>0$ such that for all $\zeta \in \Xi \backslash \Xi_{0}$ we have the bounds:

$$
\forall v \in E^{s}(\zeta),\left|\mathscr{S}(\zeta, 0) A_{d} v\right| \leq c_{0}|v| \text { and } \forall w \in E^{u}(\zeta),\left|\mathscr{S}(\zeta, 1) A_{d} w\right| \leq c_{1}|w|
$$

From (61) we have that for all $v \in E^{s}(\zeta)$ :

$$
J^{*}(\zeta, 0) A_{d} v=\underbrace{\left(I-\phi_{0}(\zeta) B_{0}\right) v}_{=0}+\bar{T}_{0 \rightarrow 1}(\zeta) v
$$

so that, because $\operatorname{Ran} \bar{T}_{0 \rightarrow 1}(\zeta) \subset \operatorname{ker} B_{1}$, we have:

$$
N(\zeta) J^{*}(\zeta, 0) A_{d} v=\left(\Pi_{\mathrm{ker} B_{1}}^{*} N_{11}(\zeta)+\Pi_{\mathrm{ker} B_{0}}^{*} N_{01}(\zeta)\right) \bar{T}_{0 \rightarrow 1}(\zeta) v
$$

from which we deduce that for all $v \in E^{s}(\zeta)$

$$
\begin{align*}
\mathscr{S}(\zeta, 0) A_{d} v & =A_{d} J(\zeta, 0) N(\zeta) J^{*}(\zeta, 0) A_{d} v \\
& =A_{d}\left(\widetilde{\phi}_{0}(\zeta) C_{0}+e^{-\widetilde{\mathscr{A}}(\zeta)} \widetilde{\phi}_{1}(\zeta) C_{1}\right) A_{d}^{-1}\left(\Pi_{\mathrm{ker} B_{1}}^{*} N_{11}(\zeta)+\Pi_{\mathrm{ker} B_{0}}^{*} N_{01}(\zeta)\right) \bar{T}_{0 \rightarrow 1}(\zeta) w \tag{70}
\end{align*}
$$

In order to simplify as much as possible 70 we use the following lemma:

Lemma 3.6 Let $C_{0}$ and $C_{1}$ denote the boundary conditions for the dual problem characterized by (27) then we have the decomposition:

$$
\mathbb{R}^{N}=\operatorname{ker} C_{0} \oplus \operatorname{ker} C_{1},
$$

let $\Pi_{\mathrm{ker} C_{1}}$ (resp. $\Pi_{\mathrm{ker} C_{0}}$ ) be the projector upon $\operatorname{ker} C_{1}$ (resp. ker $C_{0}$ ) with respect with this decomposition. Then we have the equalities

$$
\Pi_{\mathrm{ker} B_{1}}^{*}=A_{d} \Pi_{\mathrm{ker} C_{0}} A_{d}^{-1} \text { and } \Pi_{\mathrm{ker} B_{0}}^{*}=A_{d} \Pi_{\mathrm{ker} C_{1}} A_{d}^{-1}
$$

Proof : The proof of this lemma is essentialy the same as the proof of Lemma 3.5. However let us give the proof of the first equality for a sake of completeness. Considering that $\Pi_{\text {ker } B_{1}}^{*}$ is a projector we just have to determine its kernel and its range. We have:

$$
\begin{aligned}
\operatorname{ker} \Pi_{\operatorname{ker} B_{1}}^{*} & =\operatorname{ker} B_{1}^{\perp}=A_{d} \operatorname{ker} C_{1} \\
\operatorname{Ran} \Pi_{\operatorname{ker} B_{1}}^{*} & =\operatorname{ker} B_{0}^{\perp}=A_{d} \operatorname{ker} C_{0}
\end{aligned}
$$

where we used the characterization 27 . Consequently, we have $\Pi_{\text {ker } B_{1}}^{*}=A_{d} \Pi_{\text {ker } C_{0}} A_{d}^{-1}$

From Lemma 3.6, 70 becomes for $v \in E^{s}(\zeta)$

$$
\begin{equation*}
\mathscr{S}(\zeta, 0) A_{d} v=A_{d}\left(\widetilde{\phi}_{0}(\zeta) C_{0} \Pi_{\mathrm{ker} C_{1}} A_{d}^{-1} N_{01}(\zeta)+e^{-\tilde{\mathscr{A}}(\zeta)} \widetilde{\phi}_{1}(\zeta) C_{1} \Pi_{\mathrm{ker} C_{0}} A_{d}^{-1} N_{11}(\zeta)\right) \bar{T}_{0 \rightarrow 1}(\zeta) v \tag{71}
\end{equation*}
$$

The same kind of computations leads to the fact for all $w \in E^{u}(\zeta)$

$$
\begin{equation*}
\mathscr{S}(\zeta, 1) A_{d} w=A_{d}\left(\widetilde{\phi}_{1}(\zeta) C_{1} \Pi_{\mathrm{ker} C_{0}} A_{d}^{-1} N_{01}^{*}(\zeta)+e^{\widetilde{\mathscr{A}}(\zeta)} \widetilde{\phi}_{0}(\zeta) C_{0} \Pi_{\mathrm{ker} C_{1}} A_{d}^{-1}\right) \bar{T}_{1 \rightarrow 0}(\zeta) v \tag{72}
\end{equation*}
$$

So to show that point $i i$ ) of Definition 3.3 is satisfied, from 71 and 72 , the uniform Kreiss-Lopatinskii for the dual problem it is sufficient to show that the terms

$$
\begin{equation*}
\Pi_{\text {ker } C_{1}} A_{d}^{-1} N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) v, N_{11}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) v \text { for } v \in E^{s}(\zeta) \tag{73}
\end{equation*}
$$

and

$$
\begin{equation*}
N_{01}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta) w \text { for } w \in E^{u}(\zeta) \tag{74}
\end{equation*}
$$

are uniformly controlable in terms of $v$ and $w$ respectively.
To treat the first term in 73 , from Lemma 3.6 and 68 we write:

$$
\Pi_{\mathrm{ker} C_{1}} A_{d}^{-1} N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) v=-\frac{1}{2} A_{d}^{-1}\left[\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-1} \Pi_{\mathrm{ker} B_{0}}\right]^{*} \bar{T}_{0 \rightarrow 1}^{*}(\zeta)\left(\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)+I\right) \bar{T}_{0 \rightarrow 1}(\zeta) v
$$

From Assumption 3.2 the fact that the operators $\bar{T}_{0 \rightarrow 1}^{*}(\zeta)$ and $\bar{T}_{1 \rightarrow 0}^{*}(\zeta)$ are respectively given by:

$$
\bar{T}_{0 \rightarrow 1}^{*}(\zeta)=A_{d} e^{-\widetilde{\mathscr{A}}(\zeta)} \widetilde{\phi}_{1}(\zeta) C_{1} A_{d}^{-1} \text { and } \bar{T}_{1 \rightarrow 0}^{*}(\zeta)=A_{d} e^{\tilde{\mathscr{A}}(\zeta)} \widetilde{\phi}_{0}(\zeta) C_{0} A_{d}^{-1}
$$

so that (from the fact that $\operatorname{Ran} \widetilde{\phi}_{0}(\zeta) \subset \widetilde{E}^{s}(\zeta)$ and $\left.\operatorname{Ran} \widetilde{\phi}_{1}(\zeta) \subset \widetilde{E}^{u}(\zeta)\right)$ they are uniformly bounded with respect to the parameter $\zeta$ (we refer to substep 3.2 or to Paragraph 3.3.3 for explicit computations). We only have to justify that for all $v \in E^{s}(\zeta)$

$$
\left|\bar{T}_{0 \rightarrow 1}(\zeta) v\right| \leq C|v| \text { and }\left|\bar{T}_{1 \rightarrow 0} \bar{T}_{0 \rightarrow 1}(\zeta) v\right| \leq C|v|
$$

The first inequality is clear from the definition of $\bar{T}_{0 \rightarrow 1}(\zeta)$ (see 43) and from the fact that $v \in E^{s}(\zeta)$, while the second one is not clear at first glance because Ran $\bar{T}_{0 \rightarrow 1}(\zeta) \subset \operatorname{ker} B_{1}$. However from the uniform Kreiss-Lopatinskii condition on the side $\partial \Gamma_{0}$ we have:

$$
\begin{align*}
\left|\bar{T}_{0 \rightarrow 0}(\zeta) v\right| \leq C\left|e^{-\mathscr{A}(\zeta)}\left(I-\phi_{1}(\zeta) B_{1}\right) e^{\mathscr{A}} v\right| & \leq C\left(|v|+\left|e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}} v\right|\right) \\
& \leq C|v| \tag{75}
\end{align*}
$$

because $v \in E^{s}(\zeta)$ and $\operatorname{Ran} \phi_{1}(\zeta) \subset E^{u}(\zeta)$.
Consequently we obtain that:

$$
\left|\Pi_{\mathrm{ker} C_{1}} A_{d}^{-1} N_{01}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) v\right| \leq C|v|
$$

which conclude the discussion for the first term in $\sqrt[73]{ }$. We then turn to the second term in $(73)$. From the definition of $N_{11}(\zeta)$ (see 68) it reads:

$$
\begin{equation*}
N_{11}(\zeta) \bar{T}_{0 \rightarrow 1}(\zeta) v=\left(I-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)-\bar{T}_{1 \rightarrow 0}^{*}(\zeta) N_{01}(\zeta)-N_{01}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)\right) \bar{T}_{0 \rightarrow 1}(\zeta) v \tag{76}
\end{equation*}
$$

and we treat each term in (76) separately. The first term in 76 is bounded from the fact that $v \in E^{s}(\zeta)$, the second one is bounded from $(75)$ and the proof for the third one follows the one of the first term in (73) (with $\bar{T}_{1 \rightarrow 0}^{*}(\zeta)$ instead of $\Pi_{\operatorname{ker} C_{1}} A_{d}^{-1}$ ) so we essentially only have to treat the last term in $(76)$. From (68) we write:

$$
\begin{aligned}
N_{01}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta) & =\left(\bar{T}_{1 \rightarrow 0}(\zeta)^{*} N_{01}(\zeta)\right)^{*} \\
& =-\frac{1}{2}\left(\left[\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-1} \bar{T}_{1 \rightarrow 0}(\zeta)\right]^{*} \bar{T}_{0 \rightarrow 1}^{*}(\zeta)\left(I+\bar{T}_{1 \rightarrow 0}^{*}(\zeta) \bar{T}_{1 \rightarrow 0}(\zeta)\right)\right)^{*}
\end{aligned}
$$

and we conclude as for the proof of the first term in (73). The proof of the uniform boundedness of 74 follows exactly the same reasoning.

This concludes the proof of the fact that $\mathscr{S}$ defined in (52) satisfies ii) of Definition 3.3
Consequenlty we have shown the following proposition:
Proposition 3.10 Under Assumptions 2.1, 2.2, 2.3, 2.4 and 3.2. Also assume that each boundary condition satisfies the uniform Kreiss-Lopatinskii condition then the strip problem (8) admits a reduced symmetrizor in the sense of Definition 3.4.

From which we immediately deduce the following corollary by using Proposition 3.9 .
Corollary 3.1 Under Assumptions 2.1, 2.2, 2.3, 2.4 and 3.2. Also assume that each boundary condition satisfies the uniform Kreiss-Lopatinskii condition then the strip problem (1) satisfies the a priori energy estimate, that is to say that (18) is satisfied by all regular enough solution of (1).

### 3.3.3 Existence of a weak solution

In this paragraph we give two constructions of a weak solution of (8). The first one uses a stronger version of Assumption 3.1 but it has the advantage to give an explicit formula for the solution, whereas the second one only uses Assumption 3.1 and is an adaptation of the construction of the weak solution in the strictly dissipative framework (see Paragraph 3.1).

Construction of a weak solution "by hand" In this paragraph we assume the following:
Assumption 3.3 The restriction of $T_{0 \rightarrow 0}(\zeta)$ (resp. $T_{1 \rightarrow 1}(\zeta)$ ) to $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ ) is a uniform (in terms of $\zeta$ ) contraction. More precisely, this means that there exists $\left.C_{0} \in\right] 0,1\left[\right.$ (resp. $\left.C_{1} \in\right] 0,1[$ ) such that for all $\zeta \in \Xi \backslash \Xi_{0}:$

$$
\forall v \in E^{s}(\zeta)\left(\text { resp. } \forall w \in E^{u}(\zeta)\right),\left|T_{0 \rightarrow 0}(\zeta) v\right| \leq C_{0}|v| \quad\left(\text { resp. }\left|T_{1 \rightarrow 1}(\zeta) w\right|<C_{1}|w|\right)
$$

where we stress that $C_{0}$ (resp. $C_{1}$ ) does not depend on $\zeta$.
Proposition 3.11 Under Assumptions 2.1 3.3, also assume that each boundary condition satisfies the uniform Kreiss-Lopatinskii condition then the problem (1) admits a weak solution u.

Proof : By linearity we can assume that $g_{1} \equiv 0$ and we thus have to consider:

$$
\left\{\begin{array}{l}
\left.\frac{\mathrm{d}}{\mathrm{~d} x_{d}} u\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) u\left(\zeta, x_{d}\right), \text { for } x_{d} \in\right] 0,1[  \tag{77}\\
B_{0} u_{\mid x_{d}=0}(\zeta)=g_{0}(\zeta) \\
B_{1} u_{\mid x_{d}=1}(\zeta)=0
\end{array}\right.
$$

We define $u_{0}$ the solution of

$$
\left\{\begin{array}{l}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} u_{0}\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) u_{0}\left(\zeta, x_{d}\right), \text { for } x_{d} \geq 0  \tag{78}\\
u_{0_{\mid x_{d}=0}}(\zeta)=\phi_{0}(\zeta) g(\zeta)
\end{array}\right.
$$

Duhamel's formula gives:

$$
u_{0}\left(\zeta, x_{d}\right)=e^{x_{d} \mathscr{A}(\zeta)} \phi_{0}(\zeta) g(\zeta), \text { and consequently } u_{0_{\mid x_{d}=1}}(\zeta)=e^{\mathscr{A}(\zeta)} \phi_{0}(\zeta) g(\zeta)
$$

Then define $u_{1}$ the solution of the system:

$$
\left\{\begin{array}{l}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} u_{1}\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) u_{1}\left(\zeta, x_{d}\right), \text { for } x_{d} \leq 1  \tag{79}\\
u_{1_{\mid x_{d}=1}}(\zeta)=-\phi_{1}(\zeta) B_{1} u_{0_{\mid x_{d}=1}}(\zeta),
\end{array}\right.
$$

once again Duhamel's formula gives:

$$
u_{1}\left(\zeta, x_{d}\right)=-e^{\left(x_{d}-1\right) \mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} u_{0_{\mid x_{d}=1}}(\zeta)=-e^{\left(x_{d}-1\right) \mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \phi_{0}(\zeta) g(\zeta)
$$

from which we deduce:

$$
u_{1_{\mid x_{d}=0}}(\zeta)=-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \phi_{0}(\zeta) g(\zeta)
$$

Note that by construction we have:

$$
u_{0_{\mid x_{d}=0}}+u_{1_{\mid x_{d}=0}}=\left(I-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)}\right) \phi_{0}(\zeta) g(\zeta),
$$

which has to be compared with (which we recall is expected to give the "true value" of the trace of the solution on $\left.\partial \Gamma_{0}\right)$ and

$$
u_{0_{\mid x_{d}=1}}+u_{1_{\mid x_{d}=1}}=\left(I-\phi_{1}(\zeta) B_{1}\right) u_{0_{\mid x_{d}=1}}
$$

which immediately implies that $B_{1}\left(u_{0_{\mid x_{d}=1}}+u_{1_{\mid x_{d}=1}}\right)=0$. We then define recursively the sequence $\left(u_{n}\right)_{n \geq 2}$ where $u_{n}$ is the solution of:

$$
\left\{\begin{array}{l}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} u_{n}\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) u_{n}\left(\zeta, x_{d}\right), \text { for } x_{d} \geq 0 \\
u_{n_{\mid x_{d}=0}}(\zeta)=-\phi_{0}(\zeta) B_{0} u_{n-1}\left(\begin{array}{l}
\mid x_{d}=0
\end{array}\right. \\
(\zeta),
\end{array} \quad \text { if } n\right. \text { is even }
$$

and

$$
\left\{\begin{array}{l}
\frac{\mathrm{d}}{\mathrm{~d} x_{d}} u_{n}\left(\zeta, x_{d}\right)=\mathscr{A}(\zeta) u_{n}\left(\zeta, x_{d}\right), \text { for } x_{d} \leq 1 \\
u_{n_{\mid x_{d}=1}}(\zeta)=-\phi_{1}(\zeta) B_{1} u_{n-1_{\mid x_{d}=1}}(\zeta),
\end{array} \text { if } n\right. \text { is odd. }
$$

Duhamel's formula (combined with a recursive argument) immediately gives the following expressions: for all $p \in \mathbb{N}$

$$
\begin{aligned}
u_{2 p}\left(\zeta, x_{d}\right) & =e^{x_{d} \mathscr{A}(\zeta)} T_{0 \rightarrow 0}(\zeta)^{p} \phi_{0}(\zeta) g(\zeta) \\
u_{2 p+1}\left(\zeta, x_{d}\right) & =-e^{\left(x_{d}-1\right) \mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} T_{0 \rightarrow 0}(\zeta)^{p} \phi_{0}(\zeta) g(\zeta)
\end{aligned}
$$

Finally, let $S_{p}\left(\zeta, x_{d}\right):=\sum_{q=0}^{p}\left(u_{2 q}+u_{2 q+1}\right)\left(\zeta, x_{d}\right)$. In view of its definition it is clear that for all $p \in \mathbb{N}, S_{p}$ solves the evolution equation of 77 . Moreover for all $p \in \mathbb{N}$, we have:

$$
B_{1} S_{p_{\mid x_{d}=1}}(\zeta)=\sum_{q=0}^{p} \underbrace{B_{1}\left(I-\phi_{1}(\zeta) B_{1}\right)}_{=0} u_{2 q_{\mid x_{d}=1}}(\zeta)=0
$$

and

$$
S_{p_{\mid x_{d}=0}}(\zeta)=\left(I-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)}\right) \sum_{q=0}^{p} T_{0 \rightarrow 0}(\zeta)^{q} \phi_{0}(\zeta) g
$$

and once again this equation has to be compared with where $\left(I-T_{0 \rightarrow 0}(\zeta)\right)^{-1}$ has been remplaced by the truncated Neumann expansion $\sum_{q=0}^{p} T_{0 \rightarrow 0}(\zeta)^{q}$.

From Assumption $3.3, \underline{u}\left(\zeta, x_{d}\right):=\lim _{p \rightarrow \infty} S_{p}\left(\zeta, x_{d}\right)$ is well-defined and $\underline{u}(\zeta, \cdot) \in L^{2}([0,1])$ solves the evolution equation of 77 as well as the boundary condition on $x_{d}=1$. Finally:

$$
\begin{aligned}
B_{0} \underline{u}_{\mid x_{d}=0}(\zeta) & =B_{0}\left(I-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)}\right) \sum_{q=0}^{\infty} T_{0 \rightarrow 0}(\zeta)^{q} \phi_{0}(\zeta) g \\
& =g(\zeta)+B_{0} T_{0 \rightarrow 0}(\zeta) \sum_{q=0}^{\infty} T_{0 \rightarrow 0}(\zeta)^{q} \phi_{0}(\zeta) g-B_{0} e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)} \sum_{q=0}^{\infty} T_{0 \rightarrow 0}(\zeta)^{q} \phi_{0}(\zeta) g \\
& =g(\zeta)
\end{aligned}
$$

where we used the relation:

$$
B_{0} T_{0 \rightarrow 0}(\zeta)=\underbrace{B_{0} \phi_{0}(\zeta)}_{=I} B_{0} e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)}
$$

Consequently $\underline{u}$ is a weak solution of 77 . Then to treate the case $g_{1} \neq 0$, we repeat essentially the same construction.

Before to end this paragraph let us make some comments about the constructed weak solution. This weak solution is constructed as the superposition of wave reflected wave packets. Each wave packet reads under the form $T_{0 \rightarrow 0}(\zeta)^{p}$ so that it corresponds to the initial boundary trace $\phi_{0}(\zeta) g$ which has been reflected on $\partial \Gamma_{1}$ and on $\partial \Gamma_{0}, p$ times.

Consequently it seems natural to impose that $T_{0 \rightarrow 0}(\zeta)$ is a contraction (that is the energy of reflected rays decreases) because if it is not then we expect the solution to blow up in finite time.

Existence of a weak solution by duality To show that (1) admits a solution we use the duality method. We essentially just have to show that Assumption 3.2 is satisfied by the dual problem of (1) (if it is satisfied by the primal problem (1) and we can then conclude that there exists a weak solution of (1) exactly as it has been done in Paragraph 3.1.2.
For convenience we recall that the dual problem reads:

$$
\begin{cases}L^{*}(\partial) v=f & \text { on } \Omega  \tag{80}\\ C_{0} v_{\mid x_{d}=0}=g_{0} & \text { on } \partial \Omega_{0} \\ C_{1} v_{\mid x_{d}=1}=g_{1} & \text { on } \partial \Omega_{1}\end{cases}
$$

and that the matrices $C_{0} \in \mathbf{M}_{(N-p) \times N}(\mathbb{R})$ and $C_{1} \in \mathbf{M}_{p \times N}(\mathbb{R})$ are characterized by:

$$
\begin{equation*}
\operatorname{ker} C_{j}=\left(A_{d} \operatorname{ker} B_{j}\right)^{\perp} \tag{81}
\end{equation*}
$$

and also recall $\widetilde{\phi}_{0}(\zeta)$ and $\widetilde{\phi}_{1}(\zeta)$ stand for the inverse given by the uniform Kreiss-Lopatinskii condition for the dual problem 80).

In fact we will show that if the primal problem (1) satisfies Assumption 3.1 then the operator $\left(I-\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)\right)$ (where $\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)$ denotes the operator $\bar{T}_{0 \rightarrow 0}(\zeta)$ for the dual problem) is uniformly invertible over ker $C_{0}$. This result will however be sufficient to conclude because to perform the remaining existence part of the proof in

Paragraph (3.1.2) we just use the fact that the solution of the dual problem is unique, thanks to the fact that this problem satisfies the a priori energy estimate.

Recall that in the proof of the a priori energy estimate in Paragraph 3.3.2 we used the fact that $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ is uniformly invertible over $\operatorname{ker} B_{0}$ (that is Assumption 3.2) and not that $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ is uniformly invertible on $E^{s}(\zeta)$ (that is Assumption 3.1).

We introduce the operators $\tilde{\bar{T}}_{0 \rightarrow 1}(\zeta)$ and $\widetilde{\bar{T}}_{1 \rightarrow 0}(\zeta)$ associated to the dual problem $(80)$. They are given by:

$$
\widetilde{\bar{T}}_{0 \rightarrow 1}(\zeta):=\left(I-\widetilde{\phi}_{1}(\zeta) C_{1}\right) e^{\tilde{\mathscr{A}}(\zeta)} \text { and } \widetilde{\bar{T}}_{1 \rightarrow 0}(\zeta):=\left(I-\widetilde{\phi}_{0}(\zeta) C_{0}\right) e^{-\tilde{\mathscr{A}}(\zeta)}
$$

The main result of this paragraph is the following :
Proposition 3.12 The primal problem (1) satisfies Assumption 3.1 if and only if the dual problem 800 satisfies Assumption 3.2.

Proof : We want to show that for all $\zeta \in \Xi \backslash \Xi_{0}$ the operator $\left(I-\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)\right)$ is uniformly invertible over ker $C_{0}$. Noting that for all $v \in \operatorname{ker} C_{0}$ we have $\left(I-\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)\right)\left(I-\widetilde{\phi}_{0}(\zeta) C_{0}\right) v=\left(I-\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)\right) v$ it is sufficient to show that $\left(I-\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)\right)\left(I-\widetilde{\phi}_{0}(\zeta) C_{0}\right)$ is uniformly invertible over ker $C_{0}$.

To do this we use the following lemma:
Lemma 3.7 Let $A, B \subset \mathbb{C}^{N}$ such that $E=A \oplus B$ and let $T: A \rightarrow A$. Then $T: A \rightarrow A$ is invertible if and only if $T^{*}: B^{\perp} \rightarrow B^{\perp}$ is invertible. Moreover $T: A \rightarrow A$ is uniformly invertible in terms of some parameter $\zeta$ if and only if $T^{*}: B^{\perp} \rightarrow B^{\perp}$ is uniformly invertible.

From Lemma 3.7 applied to the decomposition $\mathbb{C}^{N}=\widetilde{E}^{s}(\zeta) \oplus \operatorname{ker} C_{0}$ is it equivalent to show that $\left(I-\widetilde{\phi}_{0}(\zeta) C_{0}\right)^{*}\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{*}$ is uniformly invertible over $\widetilde{E}^{s}(\zeta)^{\perp}=A_{d} E^{s}(\zeta)$ (by Lemma 3.4). By Lemma 3.5 we have $\left(I-\widetilde{\phi}_{0}(\zeta) C_{0}\right)^{*}=A_{d} \phi_{0}(\zeta) B_{0} A_{d}^{-1}$ and we compute:

$$
\begin{align*}
& \widetilde{\bar{T}}_{0 \rightarrow 1}^{*}(\zeta)=\left(e^{\tilde{\mathscr{A}}(\zeta)}\right)^{*}\left(I-\widetilde{\phi}_{1}(\zeta) C_{1}\right)^{*}=A_{d} e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} A_{d}^{-1} \\
& \widetilde{\bar{T}}_{1 \rightarrow 0}^{*}(\zeta)=\left(e^{-\tilde{\mathscr{A}}(\zeta)}\right)^{*}\left(I-\widetilde{\phi}_{0}(\zeta) C_{0}\right)^{*}=A_{d} e^{\mathscr{A}(\zeta)} \phi_{0}(\zeta) B_{0} A_{d}^{-1} \tag{82}
\end{align*}
$$

where we refer to Paragraph 3.3 .2 for the details about $\left(e^{\widetilde{\mathscr{A}(\zeta)}}\right)^{*}$. Consequently; for $X \in A_{d} E^{s}(\zeta), X=A_{d} Y$ with $Y \in E^{s}(\zeta)$ we have:

$$
\begin{aligned}
\left(I-\widetilde{\phi}_{0}(\zeta) C_{0}\right)^{*}\left(I-\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)\right)^{*} X & =A_{d} \phi_{0}(\zeta) B_{0} A_{d}^{-1}\left(I-A_{d} e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{+\mathscr{A}(\zeta)} \phi_{0}(\zeta) B_{0} A_{d}^{-1}\right) X \\
& =A_{d} \phi_{0}(\zeta) B_{0}\left(I-e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{+\mathscr{A}(\zeta)}\right) Y \\
& =A_{d}\left(\phi_{0}(\zeta) B_{0}-T_{0 \rightarrow 0}(\zeta)\right) Y \\
& =A_{d}\left(I-T_{0 \rightarrow 0}(\zeta)\right) Y
\end{aligned}
$$

because $Y \in E^{s}(\zeta)$. Consequently Assumption 3.1 for the primal problem 8 is equivalent to Assumption 3.2 for the dual problem 80 which concludes the proof.

With Proposition 3.12 we can reiterate exactly the same reasoning as for the problem with strictly dissipative boundary conditions (see Paragraph 3.1.2). We immediately obtain the existence of a weak solution $u \in L_{\gamma}^{2}(\Omega)$ for (1).
Reiterating exactly the same kind of computations as those described in the proof of Proposition 3.12 but for the operator $\left(I-\widetilde{T}_{0 \rightarrow 0}(\zeta)\right)$ instead of $\left(I-\widetilde{\bar{T}}_{0 \rightarrow 0}(\zeta)\right)$ immediately gives the following result :

Proposition 3.13 The primal problem (1) satisfies Assumption 3.2 if and only if the dual problem 80 satisfies Assumption 3.1.

### 3.3.4 Summary of the proof of Theorem 2.2 .

Assume that the strip problem (1) satisfies Assumptions 2.1, 2.2 $2.3 \cdot 2.3$ and that each boundary condition satisfies the uniform Kreiss-Lopatinskii condition.

Assume that Assumption 3.2 holds, so that (1) admits a reduced symmetrizor and Corollary 3.1 applies and (1) satisfies the a priori energy estimate.

Assume that Assumption 3.1 (or its stronger version that is Assumption 3.3) holds then one of the constructions of a weak solution given in Paragraph 3.3 .3 applies and we deduce that 1 admits a weak solution $u \in L_{\gamma}^{2}(\Omega)$.

Reiterating exactly the same reasoning as the one described for strictly dissipative boundary conditions (see Lemma 3.3) we show that this weak solution is a strong solution so that it satisfies the a priori energy estimate and by linearity it is unique. This concludes the proof of Theorem 2.2.

Before to study some examples let us give more details about the links between Assumptions 3.1] and 3.2 . As already mentioned Proposition 3.6 shows that if $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ is invertible over $E^{s}(\zeta)$ then $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ is invertible over ker $B_{0}$.

Assume that we can show the following stronger version of Proposition 3.6
Proposition 3.14 If the strip problem (1) satisfies Assumption 3.1 then it satisfies Assumption 3.2.
Then we can also show the following version of Theorem 2.2 (that is sharp, in the author's opinion)
Theorem 3.2 Under Assumptions 2.1 2.2 2.3 and 2.4, assume that Proposition 3.13 holds then Assumption 3.1 and Assumption 3.2 are equivalent.

Consequently the hyperbolic boundary value problem in the strip (1) is lower exponentially strongly wellposed in the sense of Definition 2.2 if and only if each boundary condition satisfies the uniform KreissLopatinskii condition (see Assumption-Definition 2.1) and if moreover it satisfies Assumption 3.1.

Proof : We just have to justify the equivalence between Assumption 3.1 and Assumption 3.2. From Proposition 3.13 Assumption 3.1 umplies Assumption 3.2 .

The fact that Assumption 3.2 implies Assumption 3.1 is a consequence of the duality method. Assume that Assumption 3.2 holds for the primal problem (1), then by Proposition 3.13. Assumption 3.1 holds for the dual problem (80). By Proposition 3.14 , the dual problem satisfies Assumption 3.2. This implies, by Proposition 3.12 that the primal problem (1) satisfies Assumption 3.1.

### 3.4 Examples

### 3.4.1 An example in the spirit of Osher, 1974a

In this paragraph we consider the following hyperbolic boundary value problem inspired from Osher, 1974a:

$$
\begin{cases}\partial_{t} u+A_{1} \partial_{x} u=\partial_{t} u+\left[\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right] \partial_{x} u=f & \text { for }(t, x) \in \mathbb{R} \times] 0,1[,  \tag{83}\\
B_{0} u_{\mid x=0}:=\left[\begin{array}{cc}
1 & -\alpha_{0}
\end{array}\right] u_{\mid x=0}=g_{0} & \text { for } t \in \mathbb{R}, \\
B_{1} u_{\mid x=1}:=\left[\begin{array}{ll}
-\alpha_{1} & 1
\end{array}\right] u_{\mid x=1}=g_{1} & \text { for } t \in \mathbb{R}, \\
u_{\mid t \leq 0}=0 & \text { for } x \in[0,1],\end{cases}
$$

where $\alpha_{0}$ and $\alpha_{1}$ are fixed real numbers and where the source terms $f, g_{0}, g_{1}$ are zero for negative times. Equation (83) is nothing but two transport equations (a left going one and a right going one) coupled by the boundary conditions. To ensure that Assumption 2.4 is satisfied we impose that $\left(\alpha_{0}, \alpha_{1}\right) \neq(1,1)$. It is
also easy to check that Assumptions 2.12 .2 and 2.3 are satisfied, moreover we can show that the boundary conditions are strictly dissipative if and only if

$$
\begin{equation*}
\left|\alpha_{0}\right|<1 \text { and }\left|\alpha_{1}\right|<1 \tag{84}
\end{equation*}
$$

consequently if $\alpha_{0}$ and $\alpha_{1}$ satisfy (84), Theorem 3.1 applies and 83 is strongly well-posed.
Then we compute $\mathscr{A}(\sigma)=-\sigma A_{1}$ which immediately implies that $E^{s}(\sigma)=E^{s}=$ vect $\{(1,0)\}$ and $E^{u}(\sigma)=E^{u}=\operatorname{vect}\{(0,1)\}$. So the uniform Kreiss-Lopatinskii condition is satisfied on the side $\partial \Gamma_{0}$ (resp. $\left.\partial \Gamma_{1}\right)$ for all $\alpha_{0}\left(\right.$ resp. $\left.\alpha_{1}\right)$.

We then study Assumptions 3.1 and 3.2 , we compute:

$$
\left(I-T_{0 \rightarrow 0}(\sigma)\right)\left[\begin{array}{l}
1 \\
0
\end{array}\right]=\left(I-\phi_{0}(\sigma) B_{0} e^{-\mathscr{A}(\sigma)} \phi_{1}(\sigma) B_{1} e^{\mathscr{A}(\sigma)}\right)\left[\begin{array}{l}
1 \\
0
\end{array}\right]=\left(1-\alpha_{0} \alpha_{1} e^{-2 \sigma}\right)\left[\begin{array}{l}
1 \\
0
\end{array}\right]
$$

from which we deduce that Assumption 3.1 is satisfied if and only if $\left|\alpha_{0} \alpha_{1}\right|<1$. We now study Assumption 3.2 by a direct computation, we have:

$$
\left(I-\bar{T}_{0 \rightarrow 0}(\sigma)\right)\left[\begin{array}{c}
\alpha_{0} \\
1
\end{array}\right]=\left(I-\left(I-\phi_{0}(\sigma) B_{0}\right) e^{-\mathscr{A}(\sigma)}\left(1-\phi_{1}(\sigma) B_{1}\right) e^{\mathscr{A}(\sigma)}\right)\left[\begin{array}{c}
\alpha_{0} \\
1
\end{array}\right]=\left(\alpha_{0} \alpha_{1} e^{-2 \sigma}-1\right)\left[\begin{array}{c}
\alpha_{0} \\
1
\end{array}\right]
$$

from which we deduce that Assumption 3.2 is satisfied if and only if $\left|\alpha_{0} \alpha_{1}\right|<1$. In particular we recall that Assumption 3.2 holds if and only if Assumption 3.1 holds.

Consequently if $\left|\alpha_{0} \alpha_{1}\right|<1$ we can apply Theorem 2.2 and therefore 83 is lower exponentially strongly well-posed.

So we have shown on this example that the strip problem (83) can be lower exponentially strongly wellposed for boundary conditions which are not strictly dissipative and that Assumptions $3.1 / 3.2$ are satisfied for more boundary conditions that only the strictly dissipative ones.

### 3.4.2 The wave equation

In this paragraph we consider the two dimensional wave equation:

$$
\begin{cases}\partial_{t} u+\left[\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right] \partial_{1} u+\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right] \partial_{2} u=0 & \text { for }(t, x) \in \mathbb{R} \times] 0,1[\times \mathbb{R}  \tag{85}\\
B_{0} u_{\mid x_{1}=0}:=\left[\begin{array}{ll}
1 & -\alpha_{0}
\end{array}\right] u_{\mid x_{1}=0}=g & \text { for }\left(t, x_{2}\right) \in \mathbb{R} \times \mathbb{R} \\
B_{1} u_{\mid x_{1}=1}:=\left[\begin{array}{ll}
1 & -\alpha_{1}
\end{array}\right] u_{\mid x_{1}=1}=0 & \text { for }\left(t, x_{2}\right) \in \mathbb{R} \times \mathbb{R} \\
u_{\mid t \leq 0}=0 & \text { for } x \in[0,1] \times \mathbb{R}\end{cases}
$$

where $\alpha_{0}$ and $\alpha_{1}$ are fixed real numbers satisfying $\alpha_{0} \neq \alpha_{1}$ to ensure Assumption 2.4. We can verify that the boundary conditions are strictly dissipative if and only if

$$
\begin{equation*}
\left|\alpha_{0}\right|<1 \text { and }\left|\alpha_{1}\right|>1 \tag{86}
\end{equation*}
$$

As the system (85) only has two equations we can apply the result of Strang, 1969 establishing that the boundary conditions of (85) satisfy the uniform Kreiss-Lopatinskii condition if and only they are strictly dissipative. Consequently the strip problem 85) is strongly well-posed if and only if (86) holds.

Our aim is here to verify that under 86, Assumption 3.1 is satisfied. To this end, we compute:

$$
\mathscr{A}(\zeta)=\left[\begin{array}{cc}
-\sigma & -i \eta \\
i \eta & \sigma
\end{array}\right]
$$

and let $X(\zeta) \in \mathbb{C}$ be a solution of $\operatorname{det}(\mathscr{A}(\zeta)-X(\zeta))=0$. We have

$$
\begin{equation*}
X(\zeta)^{2}=\sigma^{2}+\eta^{2} \tag{87}
\end{equation*}
$$

and we define $X^{s}:=X^{s}(\zeta)=X_{R}^{s}+i X_{I}^{s}$ the solution of 87) with strictly negative real part and $X^{u}:=X^{u}(\zeta)$ the one with strictly positive real part. We have $X^{s}=-X^{u}$.

Thus we can parametrized $E^{s}(\zeta)=\operatorname{vect}\left\{\left(-i \eta, \sigma+X^{s}\right)\right\}$ and $E^{u}(\zeta)=\operatorname{vect}\left\{\left(-i \eta, \sigma-X^{s}\right)\right\}$ from which we deduce the inverses given by the uniform Kreiss-Lopatinskii condition, we have:

$$
\phi_{0}(\zeta) x:=-\frac{x}{i \eta+\alpha_{0}\left(\sigma+X^{s}\right)}\left[\begin{array}{c}
-i \eta \\
\sigma+X^{s}
\end{array}\right] \text { and } \phi_{1}(\zeta) y:=-\frac{y}{i \eta+\alpha_{1}\left(\sigma-X^{s}\right)}\left[\begin{array}{c}
-i \eta \\
\sigma-X^{s}
\end{array}\right]
$$

Some computations give:

$$
\left(I-\phi_{0}(\zeta) B_{0} e^{-\mathscr{A}(\zeta)} \phi_{1}(\zeta) B_{1} e^{\mathscr{A}(\zeta)}\right)\left[\begin{array}{c}
-i \eta \\
\sigma+X^{s}
\end{array}\right]=\left(1-e^{2 X^{s}} \frac{i \eta+\alpha_{1}\left(\sigma+X^{s}\right)}{i \eta+\alpha_{1}\left(\sigma-X^{s}\right)} \frac{i \eta+\alpha_{0}\left(\sigma-X^{s}\right)}{i \eta+\alpha_{0}\left(\sigma+X^{s}\right)}\right)\left[\begin{array}{c}
-i \eta \\
\sigma+X^{s}
\end{array}\right]
$$

from which it follows that for Assumption 3.1 (or equivalently Assumption 3.1) to hold it is sufficient that

$$
\begin{equation*}
\left|\frac{i \eta+\alpha_{1}\left(\sigma+X^{s}\right)}{i \eta+\alpha_{1}\left(\sigma-X^{s}\right)} \frac{i \eta+\alpha_{0}\left(\sigma-X^{s}\right)}{i \eta+\alpha_{0}\left(\sigma+X^{s}\right)}\right|<1, \tag{88}
\end{equation*}
$$

because $\left|e^{2 X^{s}}\right| \leq 1$.
Some tedious (but not difficult) computations show that 88 reads under the form $-\lambda<\lambda$, where

$$
\begin{equation*}
\lambda:=\left(\alpha_{0}-\alpha_{1}\right)\left(1+\alpha_{0} \alpha_{1}\right) \eta X_{I}^{s}+\left(\alpha_{0}^{2}-\alpha_{1}^{2}\right)\left(\tau X_{I}^{s}+\gamma X_{R}^{s}\right) \tag{89}
\end{equation*}
$$

So 888 holds if and only if $\lambda>0$. By definition we have $\gamma X_{R}^{s}<0$ and from 87) (in which we take the imaginary part) we obtain that $X_{R}^{s} X_{I}^{s}=\gamma \tau$ so that $\tau X_{I}^{s} \leq 0$ and the second term in (89) is strictly positive if $\left|\alpha_{0}\right|<\left|\alpha_{1}\right|$ (which holds in particular for all strictly dissipative boundary conditions).

The first term in 89 is unfortunately much more difficult to deal with because it is not clear that the product $\eta X_{I}^{s}$ has a constant sign. To deal with we choose to cancel it by choosing $\left(1+\alpha_{0} \alpha_{1}\right)=0$. An consequently we show that for $\left(1+\alpha_{0} \alpha_{1}\right)=0$ and $\left|\alpha_{0}\right|<\left|\alpha_{1}\right|$, the operator $T_{0 \rightarrow 0}(\zeta)$ is a contraction with respect to $\zeta$ and we can conclude in a (really) particular case.

A more clever analysis to treat (at least) the strictly dissipative boundary conditions is to remark that for all strictly dissipative boundary conditions we have $\left(\alpha_{0}-\alpha_{1}\right)\left(1+\alpha_{0} \alpha_{1}\right)>\left(\alpha_{0}^{2}-\alpha_{1}^{2}\right)$ so that:

$$
\lambda>\left(\alpha_{0}^{2}-\alpha_{1}^{2}\right)\left(\eta X_{I}^{s}+\tau X_{I}^{s}+\gamma X_{R}^{s}\right)
$$

and to determine the sign of $X_{I}^{s}(\eta+\tau)$.

## 4 Conclusion and conjectures

In this article we gave two theorems establishing the lower exponential strong well-posedness of hyperbolic boundary value problems in a strip. The first one (see Theorem 3.1) applies in the particular framework of strictly dissipative boundary conditions. In particular it gives easily checkable sufficient conditions for lower exponential strong well-posedness. The second one, that is Theorem 2.2 is stated, in terms of the boundary conditions, in the more general possible setting. It states that the lower exponential strong well-posedness is equivalent to the fullfilment of two new inversibility condition. These conditions asks that the operators $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ and $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ are uniformly, in terms of the frequency parameter $\zeta$, invertible on the stable subspace $E^{s}(\zeta)$.

Of course, when Theorem 3.1 applies then Theorem 2.2 should also applies and consequently for the consistency of the theory it should be interesting to show that when the boundary conditions are strictly dissipative then the operators $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ and $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ are uniformly invertible.

About this question, our main conjecture is that when the boundary condition on the side partial $\Gamma_{0}$ (resp. $\partial \Gamma_{1}$ ) is strictly dissipative then the operator $T_{1 \rightarrow 0}(\zeta): E^{u}(\zeta) \rightarrow E^{s}(\zeta)$ (resp. $T_{0 \rightarrow 1}(\zeta)$ : $\left.E^{s}(\zeta) \rightarrow E^{u}(\zeta)\right)$ is a contraction so that the same holds for $T_{0 \rightarrow 0}(\zeta): E^{s}(\zeta) \rightarrow E^{s}(\zeta)$. The same conjecture holds for the operators $\bar{T}_{1 \rightarrow 0}(\zeta): \operatorname{ker} B_{1} \rightarrow \operatorname{ker} B_{0}\left(\right.$ resp. $\left.\bar{T}_{0 \rightarrow 1}(\zeta): \operatorname{ker} B_{0} \rightarrow \operatorname{ker} B_{1}\right)$.

In Paragraph 3.2 we have shown that both conditions on the operators $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ and $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)$ are necessary for the lower exponential strong well-posedness of the boundary value problem in the strip. Moreover the result of Proposition 3.6 combined with the result of Proposition 3.13 seems to indicate that in fact Assumptions 3.1 and 3.2 are equivalent. We believe that establish such an equivalent should be an interesting for the all theory (because it permits to give a version of Theorem 2.2 with Assumption 3.1 (or 3.2 only, and this question is left for future studies. The only point that we should show is a stronger version of Proposition 3.6 including the uniform bound for the inverse.

However as described in paragraph 3.4 .2 , this condition (specifically the uniform part of the inversibility) seems to be really difficult to establish for a given hyperbolic boundary value problem in the strip with general boundary conditions.

In the author's opinion, this difficulty is a reminiscence of the same one in the half space geometry. Let us briefly explain what we mean by this assertion.

In this article we used a version of the uniform Kreiss-Lopatinskii condition (see Assumption-Definition 2.1) stated on the all space $\Xi$. This condition is equivalent to the following condition: for all $\zeta \in \Xi \backslash \Xi_{0}$ the restriction of $B_{0}$ (resp. $B_{1}$ ) to $E^{s}(\zeta)$ (resp. $E^{u}(\zeta)$ ) is uniformly invertible. However when one wants to check that a given boundary condition satisfies (or not) the uniform Kreiss-Lopatinskii condition the first version of the uniform Kreiss-Lopatinskii condition (see see Assumption-Definition 2.1) is much more easy to handle with because it gives for free the uniform part of the bound of the inverse. Indeed by homogeneity we can restrict the study to the unit sphere of $\Xi$ which is compact and consequently the uniform part of the bound comes for free.

Our new inversibility condition is stated on the open space $\Xi \backslash \Xi_{0}$ so that, by analogy, with the uniform Kreiss-Lopatinskii condition, we believe that a simple way (and possibly the simplest one) to show that the inversibility condition holds uniformly on the space $\Xi \backslash \Xi_{0}$ is to consider the boundary frequencies, that is the $\zeta \in \Xi_{0}$. Extend the operator $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ on $E^{s}(\zeta)$ up to the boundary $\Xi_{0}$ is not difficult because as noticed by Kreiss, 1970 and Métivier, 2000 the stable subspace $E^{s}(\zeta)$ admits a continuous extension up to $\Xi_{0}$. Moreover, instead of showing that $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ is uniformly invertible on $E^{s}(\zeta)$, it may be easier to show that $T_{0 \rightarrow 0}(\zeta)$ is a uniform contraction on $E^{s}(\zeta)$.

However the difficulty by considering $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ for $\zeta \in \Xi_{0}$ is that this operator will generically not be invertible on $E^{s}(\zeta)$ as soon as $\zeta$ is a glancing frequency ${ }^{6}$

Indeed, for simplicity, let us consider the wave equation (85) for which the glancing area (we recall that glancing modes are the ones such that Theorem 2.1 is satisfied with a block of type $i v)$ ) is parametrized by $\gamma=0,|\tau|=|\eta|$. For such a $\underline{\zeta}$ we have: $X^{s}(\underline{\zeta})=X^{u}(\underline{\zeta})=0$ and $E^{s}(\underline{\zeta})=E^{u}(\underline{\zeta})$. So that if $v_{s}$ stands for a generator of $E^{s}(\underline{\zeta})$ we obtain:

$$
T_{0 \rightarrow 0}(\underline{\zeta}) v_{s}=e^{2 X^{s}(\underline{\zeta})} \phi_{0}(\underline{\zeta}) B_{0} \phi_{1}(\underline{\zeta}) B_{1} v_{s}=v_{s}
$$

and $\left(I-T_{0 \rightarrow 0}(\underline{\zeta})\right)$ can not be invertible on $E^{s}(\underline{\zeta})$. Moreover it can also not be a contraction.
In more generality we obtain essentially

$$
T_{0 \rightarrow 0}(\underline{\zeta})=e^{2 X^{s}(\underline{\zeta})}
$$

on $E^{s}(\underline{\zeta})$ which does not prevent $\left(I-T_{0 \rightarrow 0}(\underline{\zeta})\right)$ to be invertible but which prevent $T_{0 \rightarrow 0}(\underline{\zeta})$ to be a contraction (because $X^{s}(\underline{\zeta}) \in i \mathbb{R}$ for glancing modes).

The fact that $\left(I-T_{0 \rightarrow 0}(\underline{\zeta})\right)$ is sometimes clearly not invertible for the glancing frequencies $\underline{\zeta}$ may sound alarming. However we believe that it should not be.

Indeed, if we see the inversibility condition on $\left(I-T_{0 \rightarrow 0}(\zeta)\right)$ as a non amplification condition on selfinteracting modes, then this condition should not be imposed on glancing modes because these modes are linked to boundary layers and so they do not propagate the information (see Williams, 1996) from one side to the other. So we believe, even if at present time some technical issues remain, that the construction of

[^5]the symmetrizor, described in paragraph 3.3 .2 , can be performed for $\zeta \in \Xi_{0}$ under an inversibility condition which does not involve glancing modes.

In order to do so we believe that a complete geometric optics expansion of the problem (1) should help to understand the action of $T_{0 \rightarrow 0}(\zeta)$ near glancing modes which is probably the keystone of the construction. The author aims to give such a WKB expansion in a forthcoming contribution.

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[^0]:    ${ }^{1}$ Note that this was not the case for the quarter space geometry in which it is not possible to regularized the solution whithout modify the value of (at least) one of its traces. However in the quarter space geometry this analysis can also be done by modifying the notion of weak solution and by using the "weak=strong" lemma of Sarason, 1962. We refer to Benoit, 2015-Chapter 4 for a precise construction.

[^1]:    ${ }^{2}$ In the core of the proof of Theorem 2.2 more precisely in Paragraph 3.3 .3 , we will obtain that in fact Assumption 3.2 and Assumption 3.1 are equivalent. However, at present time, we do not have a direct proof of the reverse implication.

[^2]:    ${ }^{3}$ The terminology of symmetrizor is possibly not the best one to use. Indeed compared to Kreiss, 1970, the matrix defined in Definition 3.3 does not symmetrize the operator in the sense that we assumed that the operator satisfies Assumption 2.3 This is due to the fact that $i i i$ ) of Definition 3.3 is based on a equality and not on an inequality allowing to compensate the lack of symmetry of the operator as in Kreiss, 1970. That is why the term of "positizor" may be more relevant.

[^3]:    ${ }^{4}$ It is the only point in the proof where we use Assumption 2.4

[^4]:    ${ }^{5}$ Note that it is not clear at first glance that such $N_{01}$ is well-defined. It will depend on Ran $V$, because Assumption 3.2 only ensures that $\left(I-\bar{T}_{0 \rightarrow 0}(\zeta)\right)^{-1}$ is well-defined on ker $B_{0}$

[^5]:    ${ }^{6}$ Recall that glancing modes are generic for hyperbolic boundary value problems, except for the degenerated example 83 because from Coulombel, 2011a.

