On the $L^2$ well posedness of Hyperbolic Initial Boundary Value Problems

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Abstract

In this paper we give a class of hyperbolic systems, which includes systems with constant multiplicities but significantly wider, for which the initial boundary value problem with source term and initial and boundary data in $L^2$, is well posed in $L^2$, provided that the necessary uniform Lopatinski condition is satisfied. Moreover, the speed of propagation is the speed of the interior problem. In the opposite direction, we show on an example that, even for symmetric systems in the sense of Friedrichs, with variable coefficients and variable multiplicities, the uniform Lopatinski condition is not sufficient to ensure the well posedness of the IBVP.

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1 Introduction

This paper is concerned with the solvability in $L^2$ of the initial boundary value problem for first order $N \times N$ systems

$$
\begin{align*}
& Lu := A_0(t, x) \partial_t u + \sum_{j=1}^{d} A_j(t, x) \partial_{x_j} u + B(t, x) u = f \\
& u|_{t=0} = u_0, \\
& M u|_{x_d=0} = g
\end{align*}
$$

We consider only the case of noncharacteristic boundaries, which means that $A_d$ is invertible when $x_d = 0$. For simplicity, we have assumed here that the boundary is flat and the equation holds for $t > 0$ and $x \in \mathbb{R}^d = \{x_d > 0\}$.

The starting point is the well-known theory of hyperbolic symmetric systems in the sense of Friedrichs ([Fr1, Fr2]): if the matrices $A_j$ are bounded and Lipschitz continuous on $\mathbb{R} \times \mathbb{R}^d_+$, hermitian symmetric, if $A_0$ is definite positive with $A_0^{-1}$ bounded and if the boundary condition is maximal
strictly dissipative, then for all \( T \geq 0, u_0 \in L^2(\mathbb{R}^d), f \in L^1([0, T]; L^2(\mathbb{R}^d)) \) and \( g \in L^2([0, T] \times \mathbb{R}^{d-1}) \), the equation (1.3) has a unique solution \( u \in C^0([0, T]; L^2(\mathbb{R}^d)) \) which satisfies
\[
\|u(t)\|_{L^2(\mathbb{R}^d)} + \|u|_{x_d=0}\|_{L^2([0, t] \times \mathbb{R}^{d-1})} \leq C\|u_0\|_{L^2(\mathbb{R}^d)} + C \int_0^t \|Lu(s)\|_{L^2(\mathbb{R}^d)} ds + \|g\|_{L^2([0, t] \times \mathbb{R}^{d-1})}
\]
for some constant \( C \) independent of \( u_0 \). We call this kind of inequality semi group estimates. Applied to \( e^{\gamma t}u \), they imply the following resolvent estimate: there are constants \( C \) and \( \gamma_0 \) such that for all \( \gamma \geq \gamma_0 \) and all \( u \in C_0^\infty(\mathbb{R}^{1+d}) \),
\[
\gamma \|u\|_{L^2(\mathbb{R}^{1+d})} + \|u|_{x_d=0}\|_{L^2(\mathbb{R}^d)} \leq C\gamma^{-1}\|(L + \gamma A_0)u\|_{L^2(\mathbb{R}^{1+d})} + \|Mu|_{x_d=0}\|_{L^2(\mathbb{R}^d)}.
\]
The purpose of this paper is to understand better in which conditions these properties remain true. The focus is put on linear problems, but, by differentiating the equation, the maximal estimates above imply similar a-priori estimates in Sobolev spaces, and, using iterative schemes, they ultimately imply the local solvability in time of nonlinear problems.

In the interior, a necessary condition for (1.3) has been given by V.Ivrii and V.Petkov ([IvPe]): the principal symbol \( L_1(t, x, \tau, \xi) \) must admit a bounded microlocal symmetrizer \( S(t, x, \xi) \) (the precise definition is recalled below). This property, called strong hyperbolicity of the symbol in ([Met3]), is equivalent to the existence of a constant \( C \) such that for all \( (t, x) \in ]0, T[ \times \mathbb{R}^d_+, \xi \in \mathbb{R}^d, \gamma > 0 \) and \( u \in \mathbb{C}^N \),
\[
\gamma \|u\|_{L^2(\mathbb{R}^{1+d})} \leq C\|L_1(t, x, \tau - i\gamma, \xi)\|.
\]
From now on, this condition is assumed to be satisfied.

Similarly, on the boundary, applying the estimate (1.3) to \( u(\lambda(t - \ell), \lambda(x - x)) \) and performing a Fourier transform in the tangential variables, one obtains that a necessary condition for (1.3) is that for all \( (t, x') \) in the boundary \([0, T] \times \mathbb{R}^{d-1} \), all \( (\tau, \xi') \in \mathbb{R} \times \mathbb{R}^{d-1} \), all \( \gamma > 0 \) and all \( u \in C_0^\infty(\mathbb{R}^d_+) \) one has
\[
\gamma \|u\|_{L^2(\mathbb{R}_+)} + \|u(0)\|^2 \leq C\gamma^{-1}\|(L_1(t, x', 0, i\tau + \gamma, i\xi', \partial_{\xi_d})u\|_{L^2(\mathbb{R}_+)}^2 + CM(t, x, 0)u(0)|.
\]
Denoting by \( C_- = \{ \tau \in \mathbb{C}; \text{Im} \tau < 0 \} \), this leads to introduce for \((t, x')\) in the boundary and \( \zeta = (\tau, \xi') \in C_- \times \mathbb{R}^{d-1} \), the space

\[
E^{in}(t, x', \zeta) = L^2(\mathbb{R}_+) \cap \ker L_1(t, x', 0, \tau, \xi', \partial_{x_d}).
\]

It is the invariant space of \( G(t, x, \zeta) = A_d^{-1}(\tau A_0 + \sum_{j=1}^{d-1} \xi_j A_j) \), associated to eigenvalues in \( \{ \text{Im} \lambda < 0 \} \). Hyperbolicity implies that for \( \text{Im} \tau < 0 \), \( G(t, x, \zeta) \) has no real eigenvalues and that the dimension of \( E^{in} \) is constant and equal to the number \( N^+ \) of positive eigenvalues of \( A_0^{-1}A_d \). \( N^+ \) is the correct number of boundary conditions for (1.1) and we assume from now on, that the boundary condition \( M \) is a \( N^+ \times N \) matrix. More generally, one could consider boundary conditions where \( M \) takes its values in a \( N^+ \) dimensional vector bundle.

Applied to \( u \in E^{in} \), the estimate \((\text{estim}^3)_1.5\) implies that

\[
\forall u \in E^{in}(t, x', \zeta), \quad |u| \leq C|u(t, x', 0)u|.
\]

Thus, a necessary condition is that the uniform Lopatinski condition must be satisfied (see [Kre]):

**Definition 1.1.** The boundary condition \( M \) is said to satisfy the uniform Lopatinski condition (in short ULC) for the system \( L \), when there is a constant \( C \) such that for \((t, x')\) in the boundary and all \( \zeta = (\tau, \xi') \in C_- \times \mathbb{R}^{d-1} \), the estimate \((\text{estim}^3)_1.6\) is satisfied.

For symmetric systems in the sense of Friedrichs, this condition is therefore satisfied for maximal strictly dissipative boundary conditions. However, it is satisfied by a much wider class of boundary conditions, see e.g. [Ma-Os] or other examples below. Another important motivation for considering general boundary condition is the analysis of the stability of multidimensional shock waves initiated by A.Majda ([Maj]). H.Kreiss has shown that for strictly hyperbolic systems, the uniform Lopatinski condition implies the a priori estimate \((\text{estim}^3)_1.3\). In [Ma-Os] and [Maj] it was noticed that Kreiss' proof extended to the case where the so-called block structure condition was satisfied and in [Me1] it is shown that this latter condition is satisfied for hyperbolic systems with constant multiplicities. More recently, in [MeZu1], this result has been extended to some cases where the multiplicity varies, with applications to MHD.

At this point, several questions can be raised, and it is the goal of this paper to give them partial answers.
**Question 1.** About the resolvent estimates: to what extent can one push Kreiss' construction of symmetrizers? Recall that their existence implies estimates (1.3) for the direct problem and for the dual problem, implying the well posedness of the boundary value problem in weighted spaces $e^{\gamma t}L^2$ for $\gamma$ large; next a causality principle follows, showing that if $f$ and $g$ vanish for $t \leq t_0$, then the solution also vanishes there. This allows to solve the initial boundary value problem (1.1) with vanishing initial data $u_0$, and finally for smooth initial data.

Obviously, the obstacle to the construction of Kreiss symmetrizers is the existence of variable multiplicities. In Section 3, we give a reasonable condition which ensures the existence of smooth Kreiss symmetrizers.

**Definition 1.2.** The system $L$ belongs to the class $\mathcal{M}$ if near each point of the characteristic variety one of following condition is satisfied:

i) $L$ is analytically diagonalizable,

ii) denoting by $n$ the conormal to the boundary, either $n$ or $-n$ belongs to the cone of hyperbolic directions for the localized system.

It belongs to the class $s\mathcal{M}$ if in addition it admits a smooth symmetrizer $S(t,x,\xi)$. It belongs to the class $a\mathcal{M}$ if in addition there is smooth symmetrizer $S(t,x,\xi)$ which admits an holomorphich extension in a cone $\{|\text{Im} \xi| \leq \delta |\text{Re} \xi|\}$.

We refer to Section 3 for details. Condition i) is the geometrical form of the block structure condition (see [MeZu2]). It is satisfied in particular where the characteristic variety is smooth. The condition ii) extends the condition in [MeZu1] which concerned symmetric systems. We refer to this paper for examples.

**Theorem 1.3.** If the system belongs to the class $s\mathcal{M}$, there are families of smooth $K$-symmetrizers for $L$.

If in addition the boundary conditions satisfies the ULC, the boundary value problem is well posed in spaces $e^{\gamma t}L^2$ for $\gamma$ large enough.

The existence of Kreiss symmetrizers implies the continuity of $E^m$ up to the boundary $\text{Im} \tau = 0$ (see [MeZu2]), which is a strong limitation at points where the multiplicities of the eigenvalues varies. This question is discussed in Section 2.

**Question 2.** Is the uniform Lopatinski condition sufficient in general for the validity of (1.3)? In the constant coefficients case, the analysis in [GMWZ] shows that, if the estimate (1.2) (or (1.5)) is satisfied for one boundary matrix $M_0$ (and then $M_0$ is necessary UL), then it is satisfied for
all UL boundary condition $M$. This applies to symmetric systems which always have strictly dissipative boundary condition.

But, in general, the answer to the question is negative:

**Theorem 1.4.** There are symmetric hyperbolic systems in the sense of Friedrichs and boundary conditions which satisfy the uniform Lopatinski condition, such that the boundary value problem is ill posed in $C^\infty$.

An example is given in Section 6. Of course, it has variable coefficients, variable multiplicities and the boundary conditions are not dissipative. The strength of the result is that the well posed-ness is ruined not only in $L^2$ but also in $C^\infty$.

**Question 3.** What can be said about the local theory, in particular about local uniqueness and finite speed of propagation? We tackle this question under the angle of the invariance of the assumptions by change of time. In Section 2 we prove the following result:

**Theorem 1.5.** If $L$ is of class $sM$, the validity of the uniform Lopatinski condition is preserved by any change of time preserving hyperbolicity.

In particular, this proves that the speed of propagation for the boundary value problem does not exceed the speed of propagation for the interior problem. This is in sharp contrast with the case of weakly well posed problems, where the weak Lopatinski condition holds, where surface waves can propagate faster than interior waves (see [Ben, BeSe, Hor, Gar, Her]).

**Question 4.** Are the semigroup estimates (1.2) satisfied for Kreiss systems? This is easily proved when the system is symmetric, using the obvious energy balance, since the boundary term which involves the $L^2$ norm of the trace of the solution is controlled by (1.3). The general case is much more delicate. A positive answer has been established for strictly hyperbolic systems [Rau] and extended to systems with constant multiplicities [Aud]. An important consequence of this question is the solvability of the initial-boundary value problem (1.1) with all data, including $u_0$, in $L^2$. In Section 5 we obtain the following result which extends [Aud]:

**Theorem 1.6.** Assume that the coefficients of $L$ and $M$ are $W^{1,\infty}([0,T] \times \mathbb{R}^d_+)$. Suppose that the uniform Lopatinski condition is satisfied and that $L$ is of class $aM$. Then, for all $f \in L^1([0,T]; L^2(\mathbb{R}^d_+))$, $g \in L^2([0,T] \times \mathbb{R}^{d-1})$ and $u_0 \in L^2(\mathbb{R}^d_+)$, the problem (1.1) has a unique solution $u \in C^0([0,T] \times \mathbb{R}^d_+)$. Moreover, there is a constant $C$ such that the estimate (1.2) is satisfied.
2 Symbolic analysis

In this section we extend the known properties of symbols of hyperbolic boundary value problems in two directions, considering variable multiplicities and giving intrinsic definitions which make clear the invariance of these properties under a change of time direction.

Changing slightly the notations of the introduction, we do not specify the time direction and consider a family of symbols

\[ L(a, \tilde{\xi}) = \sum_{j=0}^{d} \xi_j A_j(a) \]

where \( \tilde{\xi} = (\xi_0, \ldots, \xi_d) \in \mathbb{R}^{1+d} \). The parameter \( a \) varies in a compact set \( A \), and the coefficients of the \( N \times N \) matrices are supposed to be at least continuous in \( a \). To avoid repetition, we use the following terminology: a function \( p(a, \tilde{\xi}) \) is said to be smooth [resp. analytic] if it is \( C^\infty \) [resp. real analytic] in \( \tilde{\xi} \) and of class \( W \) in \( a \in A \). Similarly, a family of spaces \( E(a, \tilde{\xi}) \) is smooth [analytic] if there is it admits locally a smooth [analytic] basis.

\( L(a, \cdot) \) is assumed to be strongly hyperbolic in some direction \( \nu^1 \). Denote by \( \Gamma_a \) the cone of hyperbolic directions of \( L(a, \cdot) \), containing \( \nu \). Then \( L(a, \cdot) \) is strongly hyperbolic in any direction \( \nu \in \Gamma_a \) (see e.g. [Met]). Shrinking \( A \) if necessary, we make the following assumption:

**Assumption 2.1.** \( \Gamma \) is a closed convex cone in \( \mathbb{R}^{1+d} \backslash 0 \) which is contained in \( \Gamma_a \) for all \( a \in A \) and there is a constant \( C \) such that for all \( a \in A \), all \( \tilde{\xi} = \text{Re} \tilde{\xi} + i \text{Im} \tilde{\xi} \in \mathbb{R}^{1+d} \) \( - i \Gamma \subset \mathbb{C}^{1+d} \backslash 0 \) and all \( u \in \mathbb{C}^N \);

\[ |\text{Im} \tilde{\xi}| u | \leq C |L(a, \tilde{\xi})u| . \]

**Remark 2.2.** Changing \( \tilde{\xi} \) to \( -\tilde{\xi} \), we see that the estimate (2.2) is satisfied as well when \( \text{Im} \tilde{\xi} \in \Gamma \).

Together with \( L \) we consider boundary operator

\[ M(a)u|_{x_n=0} \]

where \( x_n = n \cdot x \) and \( n \in \mathbb{R}^{1+d} \backslash \{0\} \) is the inner conormal vector to the boundary of \( \Omega = \{ x_n > 0 \}^2 \). More intrinsically, the data is \( \mathcal{K}(a) = \ker M(a) \subset \mathbb{C}^N \).

We assume that

1 In the notations of the introduction, \( x_0 = t, \tilde{\xi}_0 = \tau \) and \( \nu = dt = (1, 0, \ldots, 0) \).
2 In the notations of the introduction, \( n = dx_d = (0, 0, \ldots, 1) \), and \( x_n = x_d \).
Assumption 2.3. For all $a \in A$, the boundary matrix $L(a, n)$ is invertible, $K$ is a bundle over $A$ of class $W$ and $\dim K(a) = N_-$, the number of negative eigenvalues of $L(a, \nu)^{-1}L(a, n)$.

2.1 Localization and microhyperbolicity

The (real) characteristic variety of $L(a, \cdot)$ is

$$C_a = \{ \hat{\xi} \in \mathbb{R}^{1+d} \setminus \{0\}, \det L(a, \hat{\xi}) = 0 \}.$$ 

We denote by $C$ the set of $(a, \hat{\xi})$ with $a \in A$ and $\hat{\xi} \in C_a$. At $(a, \hat{\xi}) \in C$, invariant data are the kernel and the image of $L(a, \hat{\xi})$. Denoting by $\iota_{a, \hat{\xi}}$ the embedding $\ker L(a, \hat{\xi}) \mapsto \mathbb{C}^N$ and by $\varpi_{a, \hat{\xi}}$ the projection $\mathbb{C}^N \mapsto \mathbb{C}^N/\text{range} L(a, \hat{\xi})$, the localized symbol at $(a, \hat{\xi})$ is

$$L_{a, \hat{\xi}}(\tilde{\eta}) = \varpi_{a, \hat{\xi}} L(a, \tilde{\xi}) \iota_{a, \hat{\xi}}.$$

It acts from $\ker L(a, \hat{\xi})$ to $\mathbb{C}^N/\text{range} L(a, \hat{\xi})$. The characteristic variety of $L_{a, \hat{\xi}}$ is denoted by $C_{a, \hat{\xi}} \subset \mathbb{R}^{1+d}$.

Strong hyperbolicity implies that for $\hat{\xi} \in C_a$ one has

$$p(\tilde{\eta}) := \det \left( L(a, \tilde{\xi} + \tilde{\eta}) \right) = O(|\tilde{\eta}|^m)$$

where $m = \dim \ker L(a, \tilde{\xi})$ is the order of the root $\tau = 0$ of $p(\tilde{\xi} + \tau \nu) = 0$. The limit

$$p_0(\tilde{\eta}) = \lim_{\varepsilon \to 0} \varepsilon^{-m} p(\varepsilon \tilde{\eta})$$

exits and is homogeneous of degree $m$. Moreover, (2.14) implies that $p_0$ is hyperbolic any direction $\nu \in \Gamma_a$ (see. Lemma 8.7.2 in [Hor]). Denoting by $\Gamma_{a, \hat{\xi}}$ the cone of hyperbolic directions for $L_{a, \hat{\xi}}$ containing $\nu$ this means that

$$\forall (a, \hat{\xi}) \in C, \quad \Gamma \subset \Gamma_a \subset \Gamma_{a, \hat{\xi}}.$$

Following the terminology of [Hor] (see [KK] for the original definition) $\Gamma_{a, \hat{\xi}}$ is the cone of microhyperbolic directions near $\hat{\xi}$. Moreover, the strong form of hyperbolicity is preserved. This is the content of the next proposition.

Proposition 2.4. Let $\hat{\xi} \in C_a$ of multiplicity $m$ and let $\Gamma'$ be a closed convex subcone of $\Gamma_{a, \hat{\xi}}$. Then there is a neighborhood $\nu'$ of $(a, \hat{\xi})$ in $A \times \mathbb{C}^{1+d}$ and there are $\gamma_1 > 0$, $r > 0$ and $C$ such that :
Lemma 2.5. \( \text{Lemma } 2.5. \) There is a neighborhood \( \mathcal{V} \) of \((a, \xi)\) and there are \( \gamma_1 > 0, r > 0 \) and \( M > 0 \), such that for all matrix \( B \) with \( |B| \leq M \) and all \((a, \xi) \in \mathcal{V} \) with \( \gamma = |\text{Im} \hat{\xi}| \leq \gamma_1 \), \( \eta \in \mathbb{R}^{1+d} \) with \( |\eta| \leq r \):

\[
\text{Im} \hat{\xi} \in -\Gamma, \quad \eta \in \Gamma' \quad \Rightarrow \quad \det (A(a, \xi - i\eta) + \gamma B) \neq 0.
\]

Proof. a) Consider the polynomial in \( s \)

\[
p(b, s) = \det \left( L(a, \xi - i\gamma \nu - is\theta) + \gamma B \right)
\]

Where \( \xi \in \mathbb{R}^{1+d}, \nu \in \Gamma, \theta \in \Gamma' \) with \( |\nu| = |\theta| = 1 \) and \( b \) stands for \((a, \xi, \nu, \theta, B, \gamma)\). The assumption \((2.2)\) implies that for all matrix \( B \) with \( |B| < M = 1/C \), all real \( \xi \) and all \( \gamma > 0 \), \( L(a, \xi - i\gamma \nu) + \gamma B \) is invertible. Therefore, \( p(b, \cdot) \) has no root on the imaginary axis when \( \gamma > 0 \).

b) When \((a, \xi) = (a, \xi_0)\) and \( \gamma = 0 \), \( p(b, s) = (-is)^m p_0(\theta) + O(s^{m+1}) \) where \( p_0(\theta) \neq 0 \) and the set of \( \theta \) is compact, as well as the sets of \( B \) and \( \nu \), there is a real neighborhood \( \gamma_0 \) of \((a, \xi_0)\) and there are \( \gamma_1 \) and \( r > 0 \) such that for \((a, \xi) \in \gamma_0, |\gamma| \leq \gamma_1 \) and \( |B| \leq M \), \( p(b, \cdot) \) has exactly \( m \) roots counted with their multiplicity in the open disc \( D := \{ |s| < r \} \) and no root in \( r \leq s \leq 2r \).

c) When \((a, \xi) = (a, \xi_0), B = 0 \) and \( \gamma > 0 \), \( q(\gamma, \sigma) = (-i\gamma)^{-m} p(b, \gamma \sigma) \) is a polynomial in \( \sigma \). It extends to \( \gamma = 0 \) and at \( \gamma = 0 \), \( q(0, \sigma) = p_0(\nu + \sigma \theta) \) is a polynomial of degree \( m \) in \( \sigma \). Because both \( \nu \) and \( \theta \) belong to the cone \( \Gamma \) of hyperbolicity of \( p_0 \), \( q(0, \sigma) = 0 \) has only real negative roots (see e.g. \cite{Gar} or Lemma 8.7.3 in \cite{Hor}). By compactness in \( \nu \) and \( \theta \), there are \( R > R_1 > 0 \) such that these roots remain in \( \{ -R \leq \sigma \leq -R_1 \} \). By continuity, for \( \gamma \) small and positive, \( q(\gamma, \sigma) \) has \( m \) roots in \( |\sigma| \leq 2R \) which all satisfy \( \text{Re} \sigma < -\frac{1}{2} R_1 \).

This shows that for \((a, \xi) = (a, \xi_0), B = 0 \) and \( \gamma > 0 \) small, \( p(b, s) \) has \( m \) roots in \( \{ |s| \leq \gamma 2R, \text{Re} s < 0 \} \).
Decreasing \( \gamma_1 \) if necessary, we can assume that \( 2\gamma_1 R \leq r \), and this shows that for \( \gamma \in [0, \gamma_1] \) and \( (a, \xi) = (a, \hat{\xi}) \), \( B = 0 \), the \( m \) roots of \( p(b, \cdot) \) in the disc \( D \), are located in \( D_\gamma = \{ s \in D, \Re s < 0 \} \).

By a) and b), there are no root in \( \partial D_\gamma \) for \( (a, \xi) \in \mathcal{V}, |B| \leq M \) and \( \gamma \in [0, \gamma_1] \). Therefore, the number of roots in \( D_\gamma \) is constant and independent of \( b \) when \( \gamma > 0 \), if we have chosen, as we can, \( \mathcal{V}_R \) connected. Hence \( p(b, s) \) has no roots in \( \{ |s| \leq r, \Re s \geq 0 \} \) when \( \gamma > 0 \) and the lemma is proved. \( \Box \)

**Proof of Proposition 2.4.** Lemma 2.5 implies that for all \( (a, \xi) \in \mathcal{V} \), all \( \Im \tilde{\xi} \in -\Gamma \) and \( |\Im \tilde{\xi}| \leq \gamma_1 \) and all \( \tilde{\eta} \in \Gamma' \) with \( |\tilde{\eta}| \leq r \), \( L(a, \tilde{\xi} + i\tilde{\eta}) \) is invertible and

\[
(2.7) \quad |\Im \tilde{\xi}| |L(a, \tilde{\xi} + i\tilde{\eta})^{-1}| \leq 1/M.
\]

Because \( \Gamma_{a,\tilde{\xi}} \) is open and because \( \Gamma \) and \( \Gamma' \) are closed convex cones, there is \( \varepsilon > 0 \) such that for

\[
\tilde{\eta} \in \Gamma', \tilde{\xi} \in \Gamma, |\tilde{\xi}| \leq \varepsilon |\tilde{\eta}| \quad \Rightarrow \quad \tilde{\eta} + \tilde{\xi} \in \Gamma''
\]

where \( \Gamma'' \) is another closed subcone of \( \Gamma_{a,\tilde{\xi}} \) which contains \( \Gamma' \) in its interior. Choose \( \nu \in \Gamma \) with \( |\nu| = 1 \). There is a neighborhood \( \mathcal{V}_1 \) of \( (a, \tilde{\xi}) \) and there is \( \gamma_2 > 0 \) such that for \( (a, \Re \tilde{\xi} - i\tilde{\eta}) \in \mathcal{V}_1 \) with \( \eta \in \Gamma' \) and \( |\eta| \leq \gamma_2 \), one has \( (a, \Re \tilde{\xi} - i\varepsilon |\tilde{\eta}|\nu) \in \mathcal{V}, -\varepsilon |\tilde{\eta}|\nu \in \Gamma, | -\varepsilon |\tilde{\eta}|\nu| \leq \gamma_1 \) and \( \tilde{\eta}^1 = \tilde{\eta} - \varepsilon |\tilde{\eta}|\nu \in \Gamma_{a,\tilde{\xi}} \) with \( |\tilde{\eta}^1| \leq r \). Thus the estimate (2.7) which is valid on \( \Gamma'' \) implies that

\[
\varepsilon |\tilde{\eta}| |L(a, \Re \tilde{\xi} - i\tilde{\eta})^{-1}| \leq 1/M
\]

and (2.6) follows with \( C = 1/(\varepsilon M) \).

Part b) of the proof of the lemma above implies that for \( (a, \tilde{\xi}) \) close to \( (a, \hat{\xi}) \) and \( \theta \) of length 1 in \( \Gamma' \), \( \det L(a, \tilde{\xi} + s\theta) = 0 \) has exactly \( m \) roots in \( s \) in the disc \( \{ |s| < r \} \). Part c) says they are in \( \Im s > 0 \) when \( \Im \tilde{\xi} \in -\Gamma \).

If \( \Im \tilde{\xi} = 0 \), then (2.6) shows that the roots are located in \( \Im s \leq 0 \). Now we note that the assumption satisfied by \( (\nu, \theta) \) are also satisfied by \( (-\nu, -\theta) \) and therefore, shrinking the neighborhoods if necessary, the \( m \) roots of \( \det L(a, \xi - \theta) \) in the disc of radius \( r \) have also nonpositive imaginary part, therefore the \( m \) roots of \( \det L(a, \xi + s\theta) \) in the disc are real. This proves finishes the proof of ii). \( \Box \)
2.2 Smooth modes and the class \( \mathcal{M} \)

**Definition 2.6.** \( \mathcal{C} \) is said to be smooth at \((a, \tilde{\xi})\) if there is a neighborhood \( \mathcal{V} \) of this point in \( \mathcal{A} \times \mathbb{R}^{1+d} \) and a smooth function \( \varphi \) on \( \mathcal{V} \), such that \( d\xi \varphi (a, \tilde{\xi}) \neq 0 \) and \( \mathcal{C} \cap \mathcal{V} = \{(a, \tilde{\xi}) \in \mathcal{V} : \varphi (a, \tilde{\xi}) = 0\} \).

**Lemma 2.7.** Suppose that \( \mathcal{C} \) is smooth at \((a, \tilde{\xi})\) and given locally by the equation \( \varphi = 0 \). Then one can choose \( \varphi \) analytic in \( \tilde{\xi} \) and

i) The characteristic variety of \( L_{\xi} \tilde{\xi} \) is the hyperplane \( \{\eta \cdot d\xi \varphi (a, \tilde{\xi}) = 0\} \).

ii) There is a neighborhood \( \mathcal{V} \) of \((a, \tilde{\xi})\), and an analytic family of spaces \( E(a, \xi) \) on \( \mathcal{V} \), such that \( E(a, \xi) = \ker L(a, \xi) \) for all \((a, \tilde{\xi}) \in \mathcal{C} \cap \mathcal{V} \).

In particular, the dimension of \( \ker L(a, \tilde{\xi}) \) is constant for \((a, \tilde{\xi}) \in \mathcal{C} \cap \mathcal{V} \).

**Proof.** Consider the polynomials \( p \) and \( p_0 \) as in (2.4) and (2.5). Consider \( \nu \in \Gamma \) and choose an hyperplane \( H \) such that \( \mathbb{R}^{1+d} = \mathbb{R} \nu \oplus H \). By hyperbolicity and continuity of the roots of polynomials, the roots in \( s \) of \( p_0(\eta + s\nu) = 0 \) are the limits of \( s_\varepsilon \) where \( p(\varepsilon \eta + \varepsilon s_\varepsilon) = 0 \) for sequences \( \varepsilon \to 0 \). Thus the characteristic the set \( \{p_0 = 0\} \) is the tangent space to \( \mathcal{C} \) at \((a, \tilde{\xi})\) and this proves i).

Moreover, \( \nu \cdot \partial_\xi \varphi (a, \tilde{\xi}) \neq 0 \) since \( \nu \) is a direction of hyperbolicity, and thus non characteristic, for \( L_{\xi} \tilde{\xi} \). By the implicit function theorem, there are neighborhoods \( \mathcal{V} \) of \((a, \tilde{\xi})\) and \( \mathcal{V}_1 \) of \((a, 0)\) and a smooth function \( \lambda(a, \eta) \) on \( \mathcal{V}_1 \) such that

\[
(2.8) \quad \mathcal{C} \cap \mathcal{V} = \{(a, \tilde{\xi} + \eta + s\nu), \xi \in H, s \in \mathbb{R}, s + \lambda(a, \eta) = 0\}.
\]

In particular, for \((a, \eta) \in \mathcal{V}_1, -\lambda(a, \eta) \) is the unique eigenvalue close to 0 of \( L(a, \nu)^{-1} L(a, \xi + \eta) \) and this eigenvalue is semi-simple because of Assumption 2.1. Thus \( \lambda \) is analytic in \( \eta \) and the corresponding eigenspace \( E_\lambda(a, \eta) \) depends analytically on \( \eta \). This proves ii). \( \square \)

**Definition 2.8.** \( L \) is said to be smoothly [analytically] diagonalizable at \((a, \tilde{\xi}) \in \mathcal{C} \) if there is a neighborhood \( \mathcal{V} \) of this point in \( \mathcal{A} \times \mathbb{R}^{1+d} \), smooth functions \( \varphi_j \) on \( \omega \), and smooth family of spaces \( E_j(a, \tilde{\xi}) \) on \( \omega \), such that

i) \( \varphi_j(a, \tilde{\xi}) = 0 \) and \( d\xi \varphi_j(a, \tilde{\xi}) \neq 0 \) on \( \omega \),

ii) \( \mathcal{C} \cap \mathcal{V} = \bigcup \mathcal{C}_j \) where \( \mathcal{C}_j = \{(a, \tilde{\xi}) \in \mathcal{V}, \varphi_j(a, \tilde{\xi}) = 0\} \),

iii) the \( E_j(a, \tilde{\xi}) \) are in direct sum,

iv) for all \((a, \tilde{\xi}) \in \mathcal{C} \cap \mathcal{V} \), \( \ker L(a, \tilde{\xi}) \) is the direct sum of the \( E_k(a, \tilde{\xi}) \) for those indices \( k \) such that \((a, \tilde{\xi}) \in \mathcal{C}_k \).
Fix $\nu \in \Gamma$ and $H$ as before. $\nu$ is not characteristic for the localized symbol and, shrinking $\omega$, there are smooth [analytic] functions $\lambda_j$ for $(a, \xi) \in A \times H$ close to $(a, 0)$:

\[
\mathcal{C}_j \cap \omega = \{(a, \tilde{\xi} + \xi + sv, \xi \in H, s \in \mathbb{R}, s + \lambda_j(a, \xi) = 0\}.
\]

Hence, the $-\lambda_j(a, \xi)$ are the eigenvalue close to 0 of $L(a, \nu)^{-1}L(a, \tilde{\xi} + \xi)$. They are semi-simple because of the strong hyperbolicity.

**Remark 2.9.** This condition is very restrictive at non smooth points of $\mathcal{C}$. It is not satisfied in the example of MHD or non-isotropic Maxwell equations. However, it is interesting for two reasons:

- It is an almost necessary and sufficient condition for the validity of the block structure condition (see [MeZu2]) which is the key structural assumption for the construction of Kreiss-symmetrizers, see Section 3 below. Moreover, the definition above is intrinsic and in particular, this shows that the block structure condition is preserved by change of time.

- When all the $\mathcal{C}_j$, of codimension one, cross on a analytic submanifold $\Sigma$ of codimension 2, then, after a block reduction, we are left, locally, with the spectral analysis of a matrix of the form $\lambda(\sigma)\text{Id} + A(\sigma, \eta)$ where $\sigma \in \Sigma$, $A(\sigma, 0) = 0$ and $\eta$ is a single variable transversal to $\Sigma$. In this case, one can expect to be able to follow analytically in $\eta$ both the eigenvalues close to zero and associated eigenvectors of $A$.

At regular point $(a, \tilde{\xi}) \in \mathcal{C}$, the localized operator has the form

\[
L_{a, \tilde{\xi}}(\eta) = \tilde{\eta} \cdot d\varphi(a, \tilde{\xi}) \cdot J
\]

where $\{\varphi = 0\}$ is the local equation of $\mathcal{C}$ and $J$ an isomorphism from $\ker L(a, \xi)$ to $\mathbb{C}^N/\text{range} L(a, \xi)$. The vector field $H_\varphi$ with symbol $\tilde{\eta} \cdot d\varphi$ determines the propagation of singularities. In presence of a boundary, this depends on the position of $H_\varphi$ relatively to that boundary: tangent, incoming or outgoing. That is $\partial_\nu \varphi = n \cdot d\varphi = 0, > 0$ or $< 0$ (assuming as we may that $\nu \cdot d\varphi > 0$). In the first case, the classical terminology is that the mode $\tilde{\xi}$ is glancing, and in the other cases that it is hyperbolic. Another formulation is that $n$ is characteristic for $L_{a, \tilde{\xi}}$: $n \in \Gamma_{a, \tilde{\xi}}$ or $-n \in \Gamma_{a, \tilde{\xi}}$. These three properties make sense in general and we are led to the following definition.

**Definition 2.10.** Given the domain $\Omega = \{n \cdot x > 0\}$, $(a, \tilde{\xi}) \in \mathcal{C}$ is said hyperbolic incoming [resp. outgoing] if $n \in \Gamma_{a, \tilde{\xi}}$ [resp. $-n \in \Gamma_{a, \tilde{\xi}}$].

In this case, the boundary value problem for the localized operator needs full [resp. no] boundary conditions and no precise analysis of the singularities.
of \( C \) near \((a, \hat{\xi})\) is needed. According to the discussion before Proposition 2.4, a more correct terminology would be to say that the mode is microhyperbolic.

The condition that \( n \) is characteristic for \( L_{a, \hat{\xi}} \) also makes sense in general. However, in contrast with the situation at smooth points, at there is in general a gap between this condition and the hyperbolicity.

If \( L \) is smoothly diagonalizable near \((a, \hat{\xi})\), the characteristic variety is singular as soon as there are different sheets \( C_j \). But at these points the localized operator has a particular structure: it is block diagonal (see [MeZu1] and below) with blocks \( H_{\varphi_j J_j} \). Each of the \( H_{\varphi_j J_j} \) can be glancing, incoming or outgoing, but the analysis can be carried on because of the strong decoupling of these modes.

Summing up, the technical motivation for introducing of the class \( M \) in as in Definition 1.2 is to rule out the difficult case where the localized operator is not hyperbolic and cannot be decoupled into a diagonal system of vector fields which can be handled separately. There are other and more profound motivations that are explained in the sequel.

### 2.3 The incoming bundle, block decomposition

The Fourier-Laplace analysis of the boundary value problem relies on the spectral properties of the matrix

\[
G(a, \hat{\xi}) = L(a, n)^{-1} L(a, \hat{\xi})
\]

for complex \( \hat{\xi} \in \mathbb{R}^{1+d} - i\Gamma \), in particular in the limit \( \text{Im } \hat{\xi} \to 0 \).

For \( \hat{\xi} \in \mathbb{R}^{1+d} - i\Gamma \), the hyperbolicity implies that \( G(a, \hat{\xi}) \) has no eigenvalues on the real axis. The incoming space \( E_{\text{in}}(a, \hat{\xi}) \) is defined as the invariant space of \( G(a, \hat{\xi}) \) associated to the eigenvalues in \( \{ \text{Im } \lambda < 0 \} \). \( E_{\text{in}}(a, \hat{\xi}) \) is holomorphic in \( \hat{\xi} \in \mathbb{R}^{1+d} - i\Gamma \), and in particular, the dimension of \( E_{\text{in}} \) is constant.

If \( n \in \Gamma \) [resp. \(-n \in \Gamma\)], then one can choose above \( \hat{\xi} = -in \) [resp. \( \hat{\xi} = n \)] and since \( G(a, n) = \text{Id} \), \( \dim E_{\text{in}} = N \) [resp. \( \dim E_{\text{in}} = 0 \)]. Hence, for all \( \hat{\xi} \in \mathbb{R}^{1+d} - i\Gamma \), \( E_{\text{in}}(a, \hat{\xi}) = \mathbb{C}^N \) [resp. \( E_{\text{in}}(a, \hat{\xi}) = \{0\} \)].

So we now exclude these trivial cases and assume that

\[
(2.10) \quad n \notin \pm \Gamma.
\]

We first show that \( E_{\text{in}} \) only depends on the tangential frequencies.

**Lemma 2.11.** If \( \hat{\xi} \in \mathbb{R}^{1+d} - i\Gamma \), then for all complex number \( s \) such that \( \hat{\xi} + sn \in \mathbb{R}^{1+d} - i\Gamma \), one has

\[
E_{\text{in}}(a, \hat{\xi} + sn) = E_{\text{in}}(a, \hat{\xi}).
\]
Proof. Because $\Gamma$ is a convex cone, for all $t \in [0,1]$, $\tilde{\xi} + tsn \in \mathbb{R}^{1+d} - i\Gamma$ and the eigenvalues of $G(a, \tilde{\xi} + tsn)$ do not cross the real axis. Because the invariant spaces of $G(a, \tilde{\xi} + tsn) = G(a, \tilde{\xi}) + ts\text{Id}$ do not depend on $t$, this implies that the invariant space associated to the eigenvalues in $\{\text{Im } \lambda < 0\}$ is constant.

Consider the projection $\varpi : \mathbb{R}^{1+d} \mapsto \mathbb{R}^{1+d}/\mathbb{R}n = T^*\partial\Omega$ and its complex extension $C^{1+d} \mapsto C^{1+d}/Cn = C \otimes T^*\partial\Omega$. The image by $\varpi$ of $\mathbb{R}^{1+d} - i\Gamma$ is $T^*\partial\Omega - i\Gamma^b$ where $\Gamma^b = \varpi \Gamma$ is a closed convex cone in $T^*\partial\Omega \{0\}$. The invariance (2.11) legitimates the definition of $E^{in}$ for frequencies for $\zeta \in T^*\partial\Omega - i\Gamma^b$:

\begin{equation}
E^{in}(a, \zeta) = E^{in}(a, \tilde{\xi}), \quad \tilde{\xi} \in \mathbb{R}^{1+d} - i\Gamma, \quad \varpi\tilde{\xi} = \zeta.
\end{equation}

Another important remark is that for $\alpha \in C \{0\}$, $G(a, \alpha\tilde{\xi}) = \alpha G(a, \tilde{\xi})$ and therefore they have the same invariant spaces. Therefore, by continuity,

\begin{equation}
E^{in}(a, \alpha\zeta) = E^{in}(a, \tilde{\xi})
\end{equation}

as long as $\text{Im } \xi \in \Gamma$ and $\text{Im } (\alpha \tilde{\xi}) \in \Gamma$, since the set of $\alpha \in C$ such that $\text{Im } (\alpha \zeta) \in \Gamma$ is an open convex cone which contains 1. Introduce the open set

\begin{equation}
\mathcal{Z} = \{\alpha \tilde{\xi}, \text{ Im } \tilde{\xi} \in -\Gamma, \alpha \in C \{0\}\} \subset C^{1+d}\{0\}
\end{equation}

and its projection $\mathcal{Z}^b = \varpi \mathcal{Z} \subset C^{1+d}/Cn \approx C \otimes T^*\partial\Omega$

\begin{equation}
\mathcal{Z}^b = \{\zeta, \exists \alpha \in C \{0\} : \text{ Im } \alpha \zeta \in -\Gamma^b\}
\end{equation}

This set is conic and stable by multiplication by complex numbers $\neq 0$, but is not convex. It does not contain 0. Moreover, if $\alpha \xi = \beta \eta \in \mathcal{Z}$, with $\text{Im } \tilde{\xi}$ and $\text{Im } \tilde{\eta}$ in $-\Gamma$, then $\tilde{\eta} = \alpha/\beta \tilde{\xi}$ and by (2.13), $E^{in}(a, \zeta) = E^{in}(a, \tilde{\eta})$. Therefore, this legitimates the definition

\begin{equation}
E^{in}(a, \zeta) = E^{in}(a, \alpha\zeta), \quad \text{Im } \alpha \zeta \in -\Gamma^b
\end{equation}

for $\zeta \in \mathcal{Z}^b$, and the property (2.13) is satisfied on $\mathcal{Z}$.

An important issue is to understand the structure of the bundle $E^{in}$ in the limit $\text{Im } \zeta \to 0$.

Though this is not necessary, we simplify the exposition by choosing $\mathcal{Z} \subset \mathbb{R}^{1+d}$ an hyperplane which does not contain $n$. We identify $\mathcal{Z}$ to $T^*\partial\Omega$ considering the projection $\tilde{\varpi}$ from $\mathbb{R}^{1+d} \mapsto \mathcal{Z}$ which corresponds to the
decomposition $\tilde{\xi} = \xi + \xi_n n \in \mathbb{Z} \oplus \mathbb{R}n$. The complex cotangent space is identified with $\mathbb{Z}^C = \mathbb{Z} + i\mathbb{Z}$. In these coordinates, $\Gamma^b = \tilde{\pi}\Gamma \subset \mathbb{Z}$ and $E^{in}(a, \xi)$ is defined for $\xi \in \mathbb{Z} - i\Gamma^b$. We denote by $\Gamma^b = \Gamma^b \cup \{0\}$ the closure of $\Gamma^b$ in $\mathbb{Z}$.

Fix $\xi \in \mathbb{Z} - i\Gamma$. We study the spectral decomposition of $G(a, \xi)$ for $(a, \xi)$ close to $(a, \xi)$. Consider the distinct complex eigenvalues $\mu_k$, $k \in \{1, \ldots, k\}$ of $G(a, \tilde{\xi})$. The invariant spaces of $G(a, \xi)$ associated to $\mu_k$ can be holomorphically continued on a neighborhood $\mathcal{V}$ of $(a, \xi)$ in $\mathcal{A} \times \mathcal{Z}^C$. Denote by $E_k(a, \xi)$ these spaces. Taking holomorphic basis, this yields a block reduction

\[
G(a, \xi) = W^{-1}(a, \xi) \text{diag}(G_k(a, \xi))W(a, \xi).
\]

where the $W$ and $G_k$ are holomorphic in $\xi$ and the spectrum of $G_k(a, \tilde{\xi})$ is reduced to $\{\mu_k\}$. If $\text{Im}\xi \notin \Gamma^b$ the eigenvalues of $G_k(a, \tilde{\xi})$ are not real and the invariant subspace $E^{in}_k(a, \tilde{\xi})$ associated to eigenvalues in $\text{Im}\lambda < 0$ is well defined and holomorphic on this domain.

**Case 1.** If $\xi \in \Gamma^b$, none of the eigenvalues $\mu_k$ is real and for $(a, \xi)$ in a complex neighborhood of $(a, \xi)$, $E^{in}(a, \xi) = E_k(a, \xi)$ [resp. $E^{in}(a, \xi) = \{0\}$] if $\text{Im}\mu_k < 0$ [resp. $\text{Im}\mu_k < 0$].

**Case 2.** Suppose now that $\text{Im}\xi = 0$.

**Subcase 2.1.** If $\mu_k \notin \mathbb{R}$, then again there is a complex neighborhood $\mathcal{V}$ of $(a, \xi)$ such that for $(a, \xi) \in \mathcal{V}$ with $\text{Im}\xi \notin \Gamma^b$, $E^{in}(a, \xi) = E_k(a, \xi)$ [resp. $E^{in}(a, \xi) = \{0\}$] if $\text{Im}\mu_k < 0$ [resp. $\text{Im}\mu_k > 0$]. In particular, $E^{in}$ has an holomorphic extension to $\mathcal{V}$, which is $E_k$ or $\{0\}$.

**Subcase 2.1.** Suppose now that $\mu_k \in \mathbb{R}$. This means that $\xi^k = \xi - \mu_k n$ belongs to the real characteristic variety $\mathcal{C}$. We consider first the case where it is an hyperbolic point in the sense of Definition 2.10.

**Proposition 2.12.** If $(a, \xi^k) \in \mathcal{C}$ is hyperbolic incoming [resp. outgoing] in the sense of Definition 2.10, then there is a complex neighborhood $\mathcal{V}$ such that for $(a, \xi) \in \mathcal{V}$ with $\text{Im}\xi \in \Gamma^b$, $E^{in}(a, \xi) = E_k(a, \xi)$ [resp. $E^{in}(a, \xi) = \{0\}$]. In particular, $E^{in}$ has an holomorphic extension to $\mathcal{V}$, which is $E_k$ or $\{0\}$.

**Proof.** For $(a, \xi)$ near $(a, \xi)$, the invariant space of $G(a, \xi)$ for eigenvalues close to $\mu_k$ is the invariant space of $G(a, \tilde{\xi})$ for eigenvalues close to zero for $(a, \tilde{\xi})$ close to $(a, \xi^k)$. Suppose that $n \in \Gamma_{\xi^k}$. We apply by Proposition 2.4 with $\Gamma'$ a cone containing $\Gamma$ and $n$ and $\theta = n$. For $(a, \tilde{\xi})$ close to $(a, \xi^k)$
and $\tilde{\xi} \in \Gamma$, $\det G(a, \tilde{\xi} + sn) = 0$ has $m_k$ roots near 0 and they all belong to $\{\text{Im } s > 0\}$. Thus $\det G(a, \tilde{\xi})$ has $m_k$ eigenvalues counted with multiplicities near 0 and they all belong to $\{\text{Im } \mu < 0\}$. Projecting on $Z$ gives the result.

If $-n \in \Gamma \frac{a}{\tilde{\xi}}$, then the roots in $\text{Im } s < 0$ and the eigenvalues in $\text{Im } \mu > 0$ implying that $E^{in} = \{0\}$.

Next we consider the case where the system is smoothly diagonalizable near $(a, \tilde{\xi})$.

**Proposition 2.13.** Suppose that $L$ is analytically diagonalizable near $(a, \tilde{\xi})$. Then there on a neighborhood $V$ of $(a, \tilde{\xi})$, $E^{in} (a, \xi)$ has a continuous extension to $V \cap Z - i\Gamma^\flat$.

**Proof.** It is proved in [MeZu2] (see also Remark 3.8 below) that if a matrix $\tilde{G}(p, \zeta, \gamma)$ with parameters $p$, frequencies $\zeta \in \mathbb{R}^d$ and $\gamma > 0$, satisfies the block structure condition, then the incoming space $E^{in}(p, \zeta, \gamma)$ has a continuous extension to $\gamma = 0$. We apply this property to $G_k(a, \zeta - \mu n - i\gamma \nu)$, with $\nu \in \Gamma$ of length 1, considering $p = (a, \nu)$ as the parameters. This implies that the limit

$$E^{in}(a, \zeta, \nu) = \lim_{\gamma \to 0} E^{in}_k(a, \zeta - \mu n - i\gamma \nu)$$

exists and the convergence is locally uniform in $(a, \nu, \zeta)$. It remains to show that the limit is independent of $\nu$. This is clear from the proof in [MeZu2], since the limit is explicit in terms of $\partial_l e_j(a, \tilde{\xi})$ where the $e_j(a, \tilde{\xi})$ are analytic eigenvectors of diagonalization of $L$. \qed

**Corollary 2.14.** If the system $L$ belongs to the class $a\mathcal{M}$, the bundle $E^{in}(a, \zeta)$ has a continuous extension to $A \times (Z \setminus \{0\} - i\Gamma^\flat)$.

### 2.4 The Lopatinski condition

We consider boundary conditions (bdrycond) satisfying Assumption $L_3$. The invariant datum is the kernel of the boundary condition $\mathbb{K}(a) = \ker M(a) \subset E$ with $\text{dim } \mathbb{K} = N - N_+$. The Lopatinski determinant $D(a, \zeta)$ is the angle between $\mathbb{K}$ and $E^{in}(a, \zeta)$ or

$$D(a, \tilde{\xi}) = |\det(\mathbb{K}(a), E^{in}(a, \zeta))|$$

where the determinant is computed by taking orthonormal bases in each space. $D(a, \zeta)$ does not depend on the choice of these bases. It depends
only on the choice of a scalar product on $E$. The invariance property (2.13) shows that the natural domain of definition of $D$ is $\mathcal{Z}^\flat$, which is larger than $T^*\partial \Omega - i\Gamma^\flat$. In particular, we note that, for $\alpha \in \mathbb{C}\setminus\{0\}$, if $\zeta$ and $\alpha \zeta$ both belong to $T^*\partial \Omega - i\Gamma^\flat$ then

\begin{equation}
D(a, \zeta) = D(a, \alpha \zeta).
\end{equation}

Given a time direction $\nu \in \Gamma$, the weak Lopatinski condition is that $E^\text{in}(a, \zeta - i\gamma \nu) \cap \mathbb{K}(a) = \{0\}$, or equivalently that $D(a, \zeta - i\gamma \nu^\flat) \neq 0$, for all $(a, \zeta) \in \mathcal{A} \times T^*\partial \Omega$ and $\gamma > 0$. The strong Lopatinski condition is that there is a constant $C$ such that

\begin{equation}
\forall (a, \zeta) \in \mathcal{A} \times T^*\partial \Omega, \forall \gamma > 0 \forall u \in E^\text{in}(a, \zeta - i\gamma \nu^\flat), \quad |u| \leq C|M(a)u|.
\end{equation}

This is equivalent to the condition that there is a constant $c > 0$ such that

\begin{equation}
\forall (a, \zeta) \in \mathcal{A} \times T^*\partial \Omega, \forall \gamma > 0, \quad D(a, \zeta - i\gamma \nu^\flat) > c.
\end{equation}

Locally there are holomorphic versions of $D$:

**Lemma 2.15.** For all $(a, \zeta) \in \mathcal{A} \times \mathcal{Z}^\flat$, there are neighborhoods of $a$ and $\zeta$, there is a function $\ell(a, \zeta)$ continuous in $a$ and holomorphic in $\zeta$ and there is a constant $C > 1$ such that on $\omega$

\begin{equation}
\frac{1}{C} |\ell(a, \zeta)| \leq D(a, \zeta) \leq C|\ell(a, \zeta)|.
\end{equation}

**Proof.** One can fix an orthonormal basis $\{e_j\}$ of $E^\text{in}(a, \zeta)$. For $(a, \zeta)$ in a neighborhood of $(a, \zeta)$, the image of this basis by $\Pi(a, \zeta)$ is a basis of $E^\text{in}(a, \zeta)$. Together with a continuous basis $\{f_k\}$ of $\mathbb{K}(a)$, we can form the determinant

\[ \ell(a, \zeta) = \det (\Pi(a, \zeta)e_1, \ldots, f_1, \ldots), \]

which is holomorphic in $\zeta$ and $D(a, \zeta) = \sigma(a, \zeta)|\ell(a, \zeta)|$ where $\sigma(a, \zeta) = 1$.

**Remark 2.16.** The function $\ell$ can be globalized using analytic continuation and the property that $T^*\partial \Omega - i\Gamma^\flat$ is contractible. However, when dealing with the uniform Lopatinski condition, we think that the geometric definition (2.18) is more adapted. For instance, if $L$ is of class $\mathcal{M}$, $D$ has a continuous extension to $T^*\partial \Omega \setminus \{0\}$, while the holomorphic version $\ell$ may have no.

Theorem 1.5 is a consequence of Corollary 2.14 and of the next result:
Theorem 2.17. If the bundle \( E^{i\nu}(a, \zeta) \) has a continuous extension to \( A \times (T^*\partial \Omega \setminus \{0\} - i\Gamma^0) \) and if the uniform Lopatinski condition is satisfied in the direction \( \nu \), then for any closed subcone \( \Gamma_1 \) contained in the interior of \( \Gamma \), there exists a constant \( c > 0 \) such that

\[
\forall (a, \zeta) \in A \times (T^*\partial \Omega - i\Gamma_1^0), \quad D(a, \zeta) > c.
\]

In particular, the uniform Lopatinski condition is satisfied in all direction \( \nu \in \Gamma_1 \).

Proof. Again, it is convenient to use a parametrization \( Z \) of \( T^*\partial \Omega \). A consequence of the assumption is that \( D \) has a continuous extension to \( A \times (T^*\partial \Omega \setminus \{0\} - i\Gamma^0) \) and this extension is bounded from below by \( c \). For \( a \in A \), \( \zeta \in Z \) with \( |\zeta| = 1 \) and \( \nu \in \Gamma_1 \) with \( |\nu| = 1 \), consider the function

\[
f_{a,\zeta,\nu}(s) = D(a, \zeta - is\nu)
\]

which is defined for \( \Re s \geq 0 \).

a) We show that there is \( R > 0 \) such that for all \((a, \zeta, \nu)\)

\[
|s| \geq R \quad \Rightarrow \quad f_{a,\zeta,\nu}(s) \geq c/2.
\]

Indeed, \( |\Im \zeta/s| \leq 1/R \) and therefore if \( R \) is large \( \Im (s^{-1}\zeta - i\nu) \in -\Gamma \). Then for such \( s \), the invariance property (\ref{homD2.19}) implies that

\[
f_{a,\zeta,\nu}(s) = D(a, s^{-1}\zeta - i\nu).
\]

The uniform Lopatinski condition implies that \( D(a, \zeta) \geq c \) for real frequencies \( \zeta \in \Gamma^0 \). For such \( \zeta \), the invariance property (\ref{homD2.19}), which can be extended by continuity, implies that \( D(a, -i\zeta) = D(a, \zeta) \geq c \). Hence, by continuity and compactness, \( D(a, \zeta - i\nu) \geq c/2 \) when \( \nu \in \Gamma_1^0 \) and \( |\zeta| \) is so small enough. With (\ref{homD2.25}), this implies (\ref{homD2.24}).

b) The assumption implies that for real frequencies with \( |\zeta| = 1 \), \( D(a, \zeta) \geq c \). Hence, by continuity, there is \( \varepsilon > 0 \) such that

\[
|s| \leq R, \quad \Re s \leq \varepsilon, \quad \Rightarrow \quad f_{a,\zeta,\nu}(s) \geq c/2.
\]

c) On the compact domain \( \{|s| \leq R, \Re s \geq \varepsilon\} \), \( f_{a,\zeta,\nu} \) is proportional to an holomorphic function, which is bounded from below on the boundary. Thus the number of zeros in this domain is independent of the parameters. When \( \nu = \nu \), the assumption is that there are no roots, so that \( f_{a,\zeta,\nu} \) never vanishes on this domain. By compactness, it is uniformly bounded from below and the theorem follows. \( \square \)
3 Tangential symmetrizers

The goal of this section is to prove Theorem 1.6. The main step is to construct "Kreiss symmetrizers". We first we review their general construction. The time direction is fixed and we use the notations \((\tau, \xi) = (\tau', \xi_d)\) of the introduction. We denote by \(\zeta = (\tau, \xi')\) the tangential frequencies. We consider

\[
G(a, \zeta) = A_d(a)^{-1}(\tau A_0(a) + \sum_{j=1}^{d-1} \xi_j A_j(a)).
\]

The parameter \(a\) varies in \(\mathcal{A}\) and by homogeneity we can assume that \(\zeta \in S_d^- = \{(\tau, \xi') \in \mathbb{C} \times \mathbb{R}^{d-1}, |\tau|^2 + |\xi'|^2 = 1, \text{Im } \tau < 0\}\). The incoming space \(E_{in}(a, \zeta)\) is defined for \(\text{Im } \tau < 0\) and the uniform Lopatinski conditions is satisfied if and only if there is a constant \(C_0\) such that

\[
\forall (a, \zeta) \in \mathcal{A} \times S_d^+, \forall u \in E_{in}(a, \zeta) \quad |u| \leq C_0 |M(a)u|.
\]

**Definition 3.1.** A bounded symmetrizer on \(\Omega = \omega \times U, U \subset S_d^-\), is a smooth matrix \(S(a, \zeta)\) on \(\Omega\), such that there are \(C, c > 0\) such that for all \((a, \zeta) \in \Omega,

\begin{align*}
(3.3) & \quad S(a, \zeta) = S^*(a, \zeta), \\
(3.4) & \quad |S(a, \zeta)| \leq C, \\
(3.5) & \quad \text{Im } S(a, \zeta) G(p, \zeta) \geq c|\text{Im } \tau| \text{Id},
\end{align*}

It is a Kreiss symmetrizer for the boundary condition \(M\) if in addition, there are positive constants \(C_1\) and \(c_1\) such that

\[
(3.6) \quad S(a, \zeta) + C_1 M^*(a) M(a) \geq c_1 \text{Id}.
\]

The symmetrizer is continuous [smooth], if it extends continuously [smoothly] to \(\omega \times \overline{U} \subset \omega \times S_d^-\).

**Remark 3.2.** Changing the constants, one can replace \((3.6)\) by

\[
(3.7) \quad S(a, \zeta) \geq c_1 \text{Id} \quad \text{on } \ker M(a).
\]

Theorem 1.6 is a consequence of the following two results:

**Theorem 3.3.** Under the assumptions of Theorem 1.6, there is a smooth Kreiss symmetrizer.
Theorem 3.4. If there is a smooth Kreiss symmetrizer, the maximal estimates (2.3) are satisfied.

The remaining part of this section is devoted to the proof of the first theorem. The second is proved in [Kre, Maj, Chi] when the coefficients are smooth in \((t,x)\) and for instance in [Me5] when the coefficients are Lipschitz.

3.1 The general strategy

The holomorphic regularity in \(\tau\) is forgotten. In [Kre], O.Kreiss constructs near each point \((a,\zeta)\in A\times S^d\), families of symmetrizers \(S^\kappa\) which are independent of the boundary conditions, such that the negative cone of \(S^\kappa\) is an arbitrarily small conic neighborhood of \(E^{in}\). Next, he uses the uniform Lopatinski condition to chooses the parameter: because \(\ker M\) does not intersect \(E^{in}\), it is contained in the positive cone of \(S^\kappa\) if \(\kappa\) large enough, implying (6.6). The construction of the \(S^\kappa\) is performed locally, and we sum up the main intermediate step in the following definition:

**Definition 3.5.** Let \((a,\zeta)\in A\times S^d\). Consider a family of symmetrizers \(S^\kappa\) on \(\omega^\kappa \times U^\kappa\) where the \(\omega^\kappa\) are neighborhoods of \(a\) and \(U^\kappa = U^{\kappa} \cap S^d\) where the \(U^\kappa\) are neighborhoods of \(\zeta\) in \(S^d\). It is called a K-family near \((a,\zeta)\) if there is a space \(\tilde{E}\) of dimension \(\tilde{N}^+\) and a projector \(\Pi\) on \(\tilde{E}\) such that for all \((a,\zeta)\in \omega^\kappa \times U^\kappa\) and for all \(\kappa\),

\[
S^\kappa(a,\zeta) \geq m(\kappa)\Pi^{\kappa}\Pi' - \Pi'\Pi
\]

where \(\Pi' = \Id - \Pi\) and \(m(\kappa) \to +\infty\) as \(\kappa \to +\infty\).

Note that the constants \(C\) and \(c\) in (5.4) (5.5) may (and do in general) depend on \(\kappa\).

**Remark 3.6.** If \(S\) is continuous at \((a,\zeta)\), or has a continuous extension at this point when \(\Im \tau = 0\), shrinking the neighborhoods if necessary and changing the parameters, it is sufficient to verify (5.8) at \((a,\zeta)\).

**Remark 3.7.** The choice of the projector \(\Pi\) is arbitrary, if one accepts to modify the \(S^\kappa\). If \(\tilde{\Pi}\) is another projector on \(\tilde{E}\), then \(\tilde{\Pi}\Pi = \Pi\), \(\tilde{\Pi}'\Pi = 0\) and \(\tilde{\Pi}' = \tilde{\Pi}'\Pi'\). Hence,

\[
|\tilde{\Pi}'u| = |\tilde{\Pi}'\Pi'u| \leq \tilde{C}|\Pi'u|, \quad |\Pi'u| \leq C(|\tilde{\Pi}'u| + |\tilde{\Pi}u|),
\]

with \(C = |\Pi|\) and \(\tilde{C} = |\tilde{\Pi}|\). Thus,

\[
m(\kappa)|\Pi'u|^2 - |\Pi'u|^2 \geq (m(\kappa)/\tilde{C}^2 - 2C^2)|\tilde{\Pi}'u|^2 - 2C^2|\tilde{\Pi}u|^2.
\]
Therefore, changing $S^\kappa$ to $\tilde{S}^\kappa = \frac{1}{2} C^{-2} S^\kappa$ we see that (6.8) for $S^\kappa$ and $\Pi$ implies similar estimates for $\tilde{S}^\kappa$ and $\bar{\Pi}$, with $\tilde{m}(\kappa) = m(\kappa)/2C^2C_2 - 1$. In particular, we can always choose $\bar{\Pi}$ to be the orthogonal projector on $\mathbb{E}$ for a given scalar product in $\mathbb{C}^N$.

**Remark 3.8** (see [MeZu2]). Any symmetrizer $S(a, \zeta)$ is necessarily negative definite on $\mathbb{E}^{\text{in}}(a, \zeta)$ for $\text{Im} \zeta < 0$ implying that for $(a, \zeta) \in \omega^\kappa \times U^\kappa$ and $u \in \mathbb{E}^{\text{in}}(a, \zeta)$

$$m(\kappa)|\Pi' u|^2 \leq |\Pi u|^2.$$ 

Therefore, the space $\mathbb{E}^{\text{in}}(a, \zeta)$ has a limit as $(a, \zeta) \rightarrow (a, \zeta)$ in $\mathbb{A} \times S^d$ and this limit is $\bar{\mathbb{E}}$:

$$\mathbb{E} = \lim_{(a, \zeta, \gamma) \rightarrow (a, \zeta)} \mathbb{E}^{\text{in}}(a, \zeta).$$

This shows that $\mathbb{E}$ is unique. Denoting by $\mathbb{E}^{\text{in}}(a, \zeta)$ this limit when $\zeta \in \partial S^d = S^{d-1}$ is real, the same analysis shows that the family $\mathbb{E}^{\text{in}}(a, \zeta - i\gamma \nu)$ is a Cauchy sequence for the uniform convergence on $\mathbb{A} \times S^{d-1}$ implying that the following limit is uniform in $(a, \zeta) \in \mathbb{A} \times S^{d-1}$

$$\mathbb{E}^{\text{in}}(a, \zeta) = \lim_{\gamma \rightarrow 0} \mathbb{E}^{\text{in}}(a, \zeta - i\gamma \nu).$$

**Lemma 3.9.** Suppose that $S^\kappa$ is a $K$-family of symmetrizers on $\omega^\kappa \times U^\kappa$. Then for any boundary condition $M$ which satisfies the uniform Lopatinski condition, $S^\kappa$ is a Kreiss symmetrizer for $\kappa$ large enough.

**Proof.** The Lopatinski condition and Remark 3.8 imply that there is a constant $C_0$ such that

$$|\Pi u| \leq C_0 |M \Pi u| \leq C_0 |M u| + C_0 |M| |\Pi' u|. $$

Thus,

$$|u|^2 \leq 2|\Pi u|^2 + 2|\Pi' h|^2 \leq 6C_0^2 |M u|^2 + 6C_0^2 |M|^2 |\Pi' u|^2 - |\Pi u|^2.$$

and, for $m(\kappa) \geq 6C_0 |M|^2$, (6.8) follows, with $C_1 = 6C_0^2$ and $c_1 = 1$. 

**Proposition 3.10.** Suppose that for all $(a, \zeta) \in \mathbb{W} \times \mathbb{S}^d_-$, there are neighborhoods $\omega^\kappa \times U^\kappa$ of $(a, \zeta)$ and a $K$-family of bounded [resp. smooth] symmetrizers $S^\kappa(p, \zeta)$ on $\omega^\kappa \times U^\kappa$. Then for any boundary condition $M$ which satisfies the uniform Lopatinski condition, there is a bounded [resp. smooth] Kreiss symmetrizer for the boundary value problem $(L, M)$. 

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Proof. By Lemma \(\text{lemKS} 3.9\), all \( (a, \zeta) \in \omega \times S^d_\omega \), has a neighborhood \( \omega \times U \) such that there is a bounded [resp. smooth] symmetrizer \( S \) on \( \omega \times U \). Therefore there is a finite covering of \( \mathcal{A} \omega \times S^d_\omega \) by open sets \( \omega_j \times U_j \) and Kreiss symmetrizers \( S_j \) on \( \omega_j \times U_j, \). Consider a a partition of unity \( 1 = \sum \chi_j \) with \( \chi^\kappa_j \) supported in \( \omega_j \times U_j \). Then \( \Sigma = \sum \chi_j S_j \) is a Kreiss symmetrizer, which is bounded [resp. smooth] on \( \mathcal{A} \times S^d_\omega \).

### 3.2 Construction of K-symmetrizers

Let \( (a, \zeta) \in A \times S^d_\omega \). To construct symmetrizer, we first perform the smooth diagonal block reduction \( G \) on a neighborhood \( \omega \times U \) of \( (a, \zeta) \):

For \( \text{Im} \, \tau < 0 \), we denote by \( E^{in}(a, \zeta) \) the invariant subspace of \( G \) associated to eigenvalues in \( \{ \text{Im} \mu < 0 \} \).

\[
E^{in}(a, \zeta) = W^{-1}(a, \zeta) \left( \bigoplus_k E^{in}_k(a, \zeta) \right).
\]

It is sufficient to construct K-families for each block separately:

\begin{equation}
\text{blokKS Lemma 3.11.} \quad \text{Suppose that for all } k, S^\kappa_k \text{ is a K-family of bounded [smooth] symmetrizers for } G_k \text{ near } (a, \zeta). \text{ There are } K\text{-families of bounded [smooth] symmetrizers } S^\kappa \text{ for } G \text{ near } (a, \zeta). \end{equation}

Proof. Taking intersection, we can find common neighborhoods \( \omega^\kappa \times U^\kappa \) for the different \( G_k \). Relabeling the families \( S^\kappa_k \), we can also assume that they satisfy (B.8) with the same \( m(\kappa) \) and by Remark B.7 that the projectors \( \Pi_k \) are the orthogonal projectors on \( E_k \).

Then \( S^\kappa = W^* \text{diag}(S^\kappa_k)W \) is a family of bounded [smooth] symmetrizers for \( G \) and for \( u = W^{-1}(u_1, \ldots, u_k)^t \), there holds

\[
(S^\kappa u, u) = \sum (S^\kappa_k u_k, u_k) \geq m(\kappa) \sum |\Pi'_{k} u_k|^2 - \sum |\Pi_k u_k|^2.
\]

Let \( \Pi = W^{-1} \text{diag}(\Pi_k)W \). It is a projector on \( E \) and

\[
|W|^{-2} (S^\kappa u, u) \geq \tilde{m}(\kappa) |\Pi' u|^2 - |\Pi u|^2
\]

with \( \tilde{m} = m/(|W^{-1}|^2|W|^2) \). Therefore, \( |W|^{-2} S^\kappa \) is a K-family near \( (a, \zeta) \).

The construction of \( S^\kappa_k \) for blocks \( G_k \) is already made in several cases (see [Kre, BeSe, Me5]).
- First, when the spectrum of $G_k(a, \zeta)$ does not intersect the real line and this is always the case when $\text{Im} \tau \neq 0$;
- When $\tau$ is real and the spectrum of $G_k(a, \zeta)$ contains real eigenvalues, we can split further the blocks to consider only the case where this spectrum is limited to a single eigenvalue $\mu_k$. In this case, $\tilde{\xi} = (\zeta, -\mu_k) \in \mathbb{R}^{1+d}\backslash \{0\}$ is characteristic for $L(a, \cdot)$. If the characteristic manifold is smooth near $(a, \tilde{\xi})$ or more generally if $L$ is smoothly diagonalizable near this point, then $\tilde{G}_k$ satisfies the block structure condition and Kreiss construction applies (see [Maj, MeZu2, Me5]).

Therefore, to finish the proof of Theorem 3.3, we only have to construct $K$ families when the block $G_k$ is associated to an hyperbolic characteristic point $\tilde{\xi}$ in the sense of Definition 2.10.

### 3.3 Symmetrizers for hyperbolic blocks

Consider $(a, \zeta) \in A \times \mathbb{R}^d\backslash \{0\}$ and an invariant block $G_k$ near this point, such that the spectrum of $G_k(a, \zeta)$ is $\{\mu_k\}$. Denote by $\tilde{\xi} = (\zeta, -\mu_k) \in \mathbb{C}_a$.

**Theorem 3.12.** Assume that the system $L$ admits a smooth symmetrizer $S(a, \xi)$. If $\tilde{\xi}$ is an hyperbolic point in the sense of Definition 2.10, then there are families of Kreiss symmetrizers for the block $G_k$.

The main part of the construction is made in the following

**Lemma 3.13.** Suppose that $\tilde{\xi}$ is incoming [resp. outgoing]. There is a smooth symmetrizer $S(a, \zeta)$ such that

\begin{align}
S &= S^* \gg 0, \quad S G_k = (S G_k)^*,
\end{align}

and

\begin{align}
\text{Re } S G_k(a, \nu) &\gg 0; \quad [\text{resp. } \text{Re } S G_k(a, \nu) \ll 0 ]
\end{align}

**Proof. a)** Because $n$ is not characteristic for the linearized symbol $L(a, \tilde{\xi})$,
\[ \text{ker } L(a, \tilde{\xi}) \cap \text{range } L(a, n)^{-1} L(a, \tilde{\xi}) = \{0\} \]

implying that $\mu_k$ is a semi simple eigenvalue of $G(a, \zeta)$, with multiplicity $m_k = \dim \text{ker } L(a, \tilde{\xi})$.

By Proposition 2.4, for $\nu \in \Gamma$ of length 1 one has for $a$ in a neighborhood of $\tilde{a}$, $(\zeta, \xi_d)$ in a neighborhood of $\tilde{\xi}$, $\gamma \geq 0$ and $\text{Re } s > 0$ small:

\begin{align}
(s + \gamma)|u| &\leq C|G(a, \zeta - i\gamma \nu) + (\xi_d - i \text{Id})u|
\end{align}
This remains true for the block \( G_k \). Moreover, Proposition \( \text{propmicro} \) also asserts that \( G(a, \zeta - i\gamma \nu) + \xi_a \text{Id} \) has \( m_k \) eigenvalues close to 0, which are real when \( \gamma = 0 \). They must be the eigenvalues of \( G_k \), and therefore, for \((a, \zeta)\) in a neighborhood of \((a, \zeta)\), \( G_k(a, \zeta) \) has only real eigenvalues. The estimate \((\text{b.12})\) implies that they are semi simple and that the eigenprojectors are uniformly bounded.

b) The existence of a smooth symmetrizer implies that there is a smooth full symmetrizer \( \tilde{S}(a, \tilde{\xi}) \) (see \([\text{FrLa1, FrLa2 Met3}]\)). It is a smooth matrix \( \tilde{S} \) such that

\[
\tilde{S}L = (\tilde{S}L)^*, \quad \text{Re} \tilde{S}(a, \tilde{\xi})L(a, \nu) \gg 0 \text{ on } \ker L(a, \tilde{\xi}).
\]

Thus \( \mathcal{S}(a, \tilde{\xi}) = \tilde{S}(a, \zeta)L(a, n) \) is a full symmetrizer for \( G(a, \tilde{\xi}) \) and this can be transported in the block decomposition \((\text{c.17})\). Therefore, for \((a, \zeta, \xi_a)\) in a neighborhood of \((a, \tilde{\xi})\) there is a smooth full symmetrizer for \( G_k(a, \zeta) + \xi_a \text{Id} \).

With a), we are now in position to apply Theorem 6.5 of \([\text{Me3}]\) to conclude that there is a smooth symmetrizer \( S_k(a, \zeta) \) for \( G_k(a, \zeta) \), satisfying \((\text{symGk}3.10)\). Moreover, the construction in \([\text{Me3}]\) implies that \( \tilde{S} = S(a, \zeta) = \tilde{S}_1(a, \tilde{\xi}) \).

c) It is sufficient to prove the third property \((\text{symGk2}3.11)\) for \( a = a \). It is also proved in \([\text{Me3}]\) that \( S(a, \tilde{\xi}) \) is a Friedrichs symmetrizer for the localized operator \( L_{a, \tilde{\xi}} \). A version of the localized operator is

\[
L'(\zeta', \xi'_a) = \zeta' \cdot \nabla \zeta G_k(a, \zeta) + \xi'_a \text{Id}
\]

and \( S_k = S_k(a, \zeta) \) is a Friedrichs symmetrizer for \( L' \). In particular, \( S_k L'(\theta) \) is definite positive for all direction \( \theta \) in the cone of hyperbolicity of \( L' \) containing \( n \). In particular this is true for \( \nu \in \Gamma \) in the incoming case and for \( \nu \in -\Gamma \) in the outgoing case and \((\text{b.11})\) follows. This finishes the proof of the lemma.

\[\square\]

**Proof of Theorem** \((\text{b.12})\). When the mode is incoming, we choose \( S_k^\kappa = -\rho S_k \) for some \( \rho > 0 \) such that the property \((\text{b.8})\) is satisfied. and \( E = E_k(a, \zeta) \). 5 Because \( G(a, \zeta - i\gamma \nu) = G(a, \zeta) - i\gamma \partial_r G(a, \zeta) + O(\gamma^2) \), we see that

\[
\text{Im} S_k^\kappa G = \gamma \rho \text{Re} S_k \partial_r G + O(\gamma^2)
\]

therefore the property \((\text{b.143})\) is satisfied if \( \gamma \) is small enough.

When the mode is outgoing, we choose \( S_k^\kappa = \kappa S_k \) and \( E = \{0\} \). Again, \((\text{b.5})\) is satisfied for \( \gamma \) small and \((\text{b.8})\) is satisfied. \[\square\]
4 Para-differential estimates

To prove Theorem 1.6 we use different pseudo or para-differential calculi. In this section we present the technical results which will be needed. On the one hand, we consider tangential operators, with symbols \( a(t, x, \tau, \xi') \) such as Kreiss symmetrizers. On the other hand, we deal with spatial operators with symbols \( a(t, x, \xi', \xi_d) \) such as symmetrizers for \( L \). Combining these two approaches is one of the major technical difficulty in the analysis of non symmetric initial boundary value problem. In this section, we gather several estimates which will be used in the proof of Theorem 5.8.

4.1 Paradifferential calculi

We give here some definitions and notations and we refer for instance to [Met2] for details.

The spatial para-differential operators we consider are associated to symbols belonging to classes denoted by \( \Gamma_0^m \) and \( \Gamma_1^m \). A symbol \( a(t, x, \xi) \) defined on \( I \times \mathbb{R}^d \times \mathbb{R}^d \) belongs to \( \Gamma_0^m \) if it is \( C^\infty \) in \( \xi \) and for all \( \alpha \in \mathbb{N}^d \)

\[
|\partial_\alpha^\xi a(t, x, \xi)| \leq C_\alpha (1 + |\xi|)^{m-|\alpha|}.
\]

It belongs to \( \Gamma_1^m \) if in addition the first derivatives \( \partial_{t,x} a \) belong to \( \Gamma_0^m \). Next, \( \Sigma^m_k \) is the set of symbols \( \sigma(t, x, \xi) \in \Gamma_0^m \) which satisfies the spectral condition that their Fourier transform with respect to the \( x \)-variables, \( \hat{\sigma}(t, \eta, \xi) \), is supported in \( \{ |\eta| \leq \varepsilon(1 + |\xi|) \} \) for some \( \varepsilon < 1 \).

The para-differential operator \( T_a \) is by definition the pseudodifferential operator

\[
T_a = \sigma_a(t, x, D_x)
\]

with symbol

\[
\sigma_a(t, x, \xi) = \int G(x - y, \xi) a(t, y, \xi) dy
\]

and

\[
G(y, \xi) = (2\pi)^{-d} \int e^{i\eta \cdot \xi} \chi(\eta, \xi) d\eta
\]

where \( \chi \) is a \( C^\infty \) function supported in \( \{ |\eta| \leq \varepsilon(1 + |\xi|) \} \), equal to 1 on \( \{ |\eta| \leq \varepsilon_1(1 + |\xi|) \} \), for some \( 0 < \varepsilon_1 < \varepsilon < 1 \) and such that

\[
|\partial_\eta^\beta \partial_\xi^\alpha \chi(\eta, \xi)| \leq C_\alpha (1 + |\xi|)^{-|\alpha|-|\beta|}.
\]
The symbol $\sigma_a$ and the quantization $T_a$ depend on the choice of the cut-off function $\chi$, but if $\chi_1$ and $\chi_2$ satisfy the spectral condition, the difference between the two symbols $\sigma^1_a$ and $\sigma^2_a$ belong to $\Sigma^{m-1}_0$ if $a \in \Gamma^m_0$ so that the two operators of order $m$ $T^1_a$ and $T^2_a$ differ by an operator of order $m - 1$ (see [Me2]). All the results below do not depend on the choice of the cutoff function $\chi$.

The tangential quantization is defined similarly, permuting the role of $t$ and $x_d$. In this case, the symbols are functions of $(t,x,\zeta)$ where $\zeta = (\tau,\xi')$.

Accordingly, we use the notation $z = (t,x')$ for the tangential variables.

To avoid confusion, when necessary, we will note $T_{tg}$ the corresponding quantization.

Remark 4.1. The Kreiss symmetrizers are functions of $(t,x,\tau,\xi',\gamma)$ and the proof of the energy estimates (4.3) relies on a pseudo or para-differential calculus, see [Kre, ChPi, MeZu1]. We do not give details here, as no new result is used for this calculus.

4.2 A microlocal Cauchy problem

We first quote a para-differential version of the classical hyperbolic Cauchy problem.

**Proposition 4.2.** Consider a matrix of symbols $G \in \Gamma^1_1$. Assume that there is a matrix $S \in \Gamma^0_1$ such that $S = S^*$ is uniformly definite positive and $SG = (SG)^*$. Then, for $u_0 \in L^2$ and $f \in L^2([0,T] \times \mathbb{R}^d)$ the Cauchy problem

$$\partial_t u + iT_G u = f, \quad u|_{t=0} = u_0$$

has a solution $u \in C^0([0,T]; L^2(\mathbb{R}^d))$

**Sketch of proof.** First, we modify the symbol $S$ into

$$\tilde{S}(t,x,\xi) = \theta(\xi)S(t,x,\xi) + \lambda(1 + |\xi|^2)^{-1} \in \Gamma^0_1$$

with $1 - \theta$ compactly supported and $\theta = 0$ near the origin, and $\lambda$ large enough so that the operator $S = \text{Re} T_{\tilde{S}}$ is definite positive in $L^2$. In this case the symbolic calculus implies the following estimate (see e.g. [Kre, MeZu2])

**Lemma 4.3.** For $u \in C^0([0,T]; L^2(\mathbb{R}^d))$ satisfying $f := (\partial_t + iT_G)u \in L^2$, one has

$$\|u(t)\|_{L^2(\mathbb{R}^d)} \lesssim \|u(0)\|_{L^2(\mathbb{R}^d)} + \int_0^t \|u(t')\|_{L^2(\mathbb{R}^d)} dt' + \int_0^t \text{Re} (Sf(t'), u(t'))_{L^2(\mathbb{R}^d)} dt'$$

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The adjoint of $T_G$ is $T_G^* + R$ where $R(t)$ is bounded in $L^2$, uniformly in $t$. $(S^*)^{-1}$ is a symmetrizer for $G^*$, and therefore there are similar estimates for the backward Cauchy problem for $-i(T_G)^*$. By duality, this implies the existence of a solution $u \in L^2([0,T] \times \mathbb{R}^d)$ of (1.3). A variant of Friedrichs' lemma, still using the symbolic para-differential calculus, implies that this solution is strong, thus belongs to $C^0([0,T]; L^2\mathbb{R}^d)$ and satisfies (1.5).

4.3 Estimates of traces

Operators of the form

\[
P_a u = (T_a u)|_{x_d=0}
\]

will occur in the analysis.

For fixed $t$, $T_a$ is bounded from $L^2(\mathbb{R}^d)$ to $H^1(\mathbb{R}^d)$ when $a \in \Gamma_0^{-1}$ and from $H^1(\mathbb{R}^d)$ to $H^1(\mathbb{R}^d)$ when $a \in \Gamma_0^0$, $T_a$. Hence,

Lemma 4.4. i) If $a \in \Gamma_0^{-1}$, then $P_a$ is bounded from $L^2$ to $L^2([0,T]; H^{1\frac{1}{2}})$.

ii) If $a \in \Gamma_0^0$, $P_a$ is bounded from $L^2([0,T]; H^1)$ to $L^2$.

When $a$ is of degree 0 conditions must be imposed to be able to define the trace of $T_a u$ when $u \in L^2$. We will assume that

\[
a(t, x, (0, \ldots, 0, \xi_d)) = 0.
\]

We will show that, up to a bounded operator in $L^2$, one can replace $a$ by $a^0 = a(t, x', 0, \xi)$ and $T_a$ by $T_a^0 = \sigma_0^a(t, x', \partial_x)$ where

\[
\sigma_0^a(t, x', \xi) = \int G^0(x' - y', \xi) a(t, y', 0, \xi) dy'.
\]

with the modified mollification kernel

\[
G^0(y', \xi') = (2\pi)^{-d} \int e^{iy' \cdot \eta}(\eta', 0, (\xi', 0)) dy'.
\]

It acts only in the variables $x'$ and on the frequency side, the cut off is made at $|\eta'| \leq \varepsilon|\xi'|$, independently of $\xi_d$. We denote by $P_0^a u$ the trace $(T_0^a u)|_{x_d=0}$.

Introduce the operator $\langle D_{x'} \rangle$ with symbol $(1 + |\xi'|^2)^{\frac{1}{2}}$

Proposition 4.5. If $a \in \Gamma_0^0$ satisfies (1.7) then $P_a$ is bounded from $L^2$ to $\langle D_{x'} \rangle^\frac{1}{2} L^2$. Moreover, $P_a - P_0^a$ is bounded from $L^2$ to $L^2$. 

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If \( a \in \Gamma^0_1 \) satisfies (4.7), there are symbols \( a_j \in \Gamma^{-1}_1 \) such that

\[
a(t, x, \xi) = \sum_{j=1}^{d-1} a_j(t, x, \xi) \xi_j + a_0(t, x, \xi).
\]

To simplify notations, we omit \( t \) in the proof below since it appears as a parameter and the \( L^2 \) integrability in time follows from the uniformity of the estimates at each fixed \( t \). Using Lemma 4.4 and that \( T_{a_j} \xi_j - \frac{1}{i} \partial_x T_{a_j} \) is bounded from \( L^2 \) to \( H^1 \), we see that for \( u \in L^2 \) the trace of \( T_a u \) belongs to \( H^{-\frac{1}{2}} \) and the first statement of the proposition is proved.

Next, we compare \( P_a \) and \( P_0^a \). First, we note that

\[
a(t, x, \xi) = a_0(t, x', \xi) + x_d b(t, x, \xi), \quad b \in \Gamma^0_0.
\]

Lemma 4.6. If \( b \in \Gamma^0_0 \), then \( T_{x_d b} - x_d T_b \) is bounded from \( L^2 \) to \( L^2([0, T]; H^1(\mathbb{R}^d)) \). In particular, the trace operator \( P_{x_d b} \) is bounded from \( L^2 \) to \( L^2 \).

Proof. Again, we omit \( t \). From (4.11) we see that \( \rho = \sigma_{x_d b} - x_d \sigma_b \) is given by

\[
\rho(t, x, \xi) = \int (x_d - y_d) G(x - y, \xi) b(y, \xi) dy.
\]

Note that

\[
y_d G(y, \xi) = i \int e^{i \eta y} \partial_{y_d} \chi(\eta, \xi) d\eta
\]

and \( \partial_{y_d} \chi \) is a symbol of degree \(-1\) supported in a cone \(|\eta| \leq \varepsilon(1 + |\xi|)\). Therefore it follows that \( \rho \in \Gamma^{-1}_0 \). Moreover, \( \sigma_{x_d b} \) and \( x_d \sigma_b \) and hence \( \rho \) satisfy the spectral condition. Therefore \( \rho \in \Sigma^{-1}_0 \) and \( \rho(x, D_x) \) is bounded from \( L^2 \) to \( H^1 \).

Hence, to prove the Proposition, it remains to study \( P_0^a - P_0^0 \). We first compare the two symbols \( \sigma_0^a \) and \( \sigma_{a_0^0} \).

Lemma 4.7. Suppose that \( a = b \xi_j \) with \( b \in \Gamma^{-1}_1 \) and \( 1 \leq j \leq d - 1 \). Then the symbol \( \rho(t, x', \xi) = \sigma_0^a - \sigma_{a_0^0} \) satisfies the estimates

\[
|\partial^\alpha_{\xi} \partial^\beta_{\xi'} \rho(t, x', \xi)| \lesssim (1 + |\xi|)^{-1+|\alpha|} (1 + |\xi'|)^{-|\beta|}.
\]

Proof. We still drop \( t \) from the notations. Because \( a_0^0 \) does not depend on \( x_d \),

\[
\sigma_{a_0^0}(x, \xi) = \int G(x' - y', x_d - y_d, \xi) a(y', 0, \xi) dy = \int G^1(x' - y', 0, \xi) dy = \sigma_{a_0^0}(x', 0, \xi)
\]
where
\[ G^1(y', \xi) = (2\pi)^{-d} \int e^{iy' \cdot \eta'} \chi(\eta', 0, \xi) d\eta'. \]

Therefore,
\[ \rho(t, x', \xi) = \xi_j \int H(x' - y', \xi) b(y', 0, \xi) dy', \]

where
\[ H(y', \xi) = (2\pi)^{-d} \int e^{iy' \cdot \eta} \theta(\eta', \xi) d\eta', \quad \theta(\eta', \xi) = \chi(\eta', 0, \xi) - \chi(\eta', 0, \xi', 0). \]

The cut off function \( \theta \) is supported in \( \{ \varepsilon (1 + |\xi'|) \leq |\eta'| \leq \varepsilon (1 + |\xi|) \} \). For all fixed \( \xi_d \), consider \( \langle \xi \rangle b(x', \xi', \xi_d) \) as a symbol in \( (x', \xi') \). They are uniformly bounded in \( \Gamma_0^1(\mathbb{R}^{d-1}) \). Since \( \theta \) is supported in \( \{ \varepsilon (1 + |\xi'|) \leq |\eta'| \} \), the Lipschitz smoothness of \( b \) in \( x' \) allows to absorb one degree in \( \xi' \) implying that
\[ |\rho(t, x', \xi)| \lesssim (1 + |\xi|)^{-1}. \]

Differentiating in \( \xi' \) hits \( \xi_j b \) and \( \theta \), and therefore
\[ |\partial_{\xi'}^\beta \rho(t, x', \xi)| \lesssim (1 + |\xi|)^{-1} (1 + |\xi'|)^{-|\beta|}. \]

Since \( \theta \) is supported in \( \{ |\eta'| \leq \varepsilon (1 + |\xi|) \} \) this implies bounds for the \( x' \) derivative and the lemma is proved.

Combining (4.10) and the lemmas above, the next result finishes the proof of Proposition 4.5.

**Lemma 4.8.** Suppose that \( \rho(x', \xi) \) satisfies (4.12). Then the mapping \( u \mapsto v = (\rho(x', \partial_x)u)|_{x_d=0} \) is bounded from \( L^2(\mathbb{R}^d) \) to \( L^2(\mathbb{R}^{d-1}) \).

**Proof.** Denote by \( \tilde{u}(x', \xi_d) \) the partial Fourier transform of \( u \) with respect to the variable \( x_d \). Then
\[ v(x') = \int_\mathbb{R} \rho(x', D_x, \xi_d) \tilde{u}(x', \xi_d) d\xi_d. \]

Next use a dyadic partition of unity in the \( \xi' \) variables, which yields a decomposition
\[ \tilde{u}(x', \xi_d) = \sum \tilde{u}_k(\xi', \xi_d) \]
so that
\[ v = \sum v_k, \quad v_k(x') = \int_\mathbb{R} w_k(x', \xi_d) d\xi_d \]

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\[ w_k(x', \xi_d) = \rho(x', D_x', \xi_d) \tilde{u}_k(x', \xi_d) = \rho_k(x', D_x', \xi_d) \tilde{u}_k(x', \xi_d) \]

where the \( \rho_k \) are localized in \( |\xi'| \approx 2^k \) and satisfy, uniformly on \( k \), for all \( |\alpha| \leq 1 \) and all \( \beta \):

\[
\left| \partial_{x'}^{\alpha} \partial_{\xi}^{\beta} \rho_k(x', \xi) \right| \lesssim (|\xi_d| + 2^k)^{|\alpha|-1 2^{-|\beta|}}. \tag{4.15}
\]

The estimate for \( \alpha = 0 \) implies that

\[
\|w_k(\cdot, \xi_d)\|_{L^2(\mathbb{R}^{d-1})} \lesssim (2^k + |\xi_d|)^{-1} \varepsilon_k(\xi_d) \tag{4.16}
\]

with \( \varepsilon_k(\xi_d) = \|\tilde{u}_k(\cdot, \xi_d)\|_{L^2} \) (see e.g. [Me2]). Similarly, applying the same estimates with \( \rho \) replaced by \((i \xi_j + \partial_{x_j})\rho\), implies that

\[
\|w_k(\cdot, \xi_d)\|_{H^1(\mathbb{R}^{d-1})} \lesssim \varepsilon_k(\xi_d). \tag{4.17}
\]

Consider dyadic partition of unity \( Q_j(D_{x'}) \) so that

\[ v = \sum_j Q_j v, \quad Q_j v = \sum_k \int Q_j w_k(x', \xi_d) dx_d. \]

Then

\[
\|Q_j w_k(\cdot, \xi_d)\|_{L^2} \lesssim \min \left( \|w_k(\cdot, \xi_d)\|_{L^2}, 2^{-j} \|w_k(\cdot, \xi_d)\|_{H^1(\mathbb{R}^{d-1})} \right)
\]

so that

\[
\|Q_j w_k(\cdot, \xi_d)\|_{L^2} \lesssim (2^j + 2^k + |\xi_d|)^{-1} \varepsilon_k(\xi_d). 
\]

Hence

\[
\|Q_j v\|_{L^2} \leq \sum_k \int \|P_j w_k(\cdot, \xi_d)\|_{L^2} d\xi_d \leq \sum_k \int \frac{\varepsilon_k(\xi_d) d\xi_d}{2^j + 2^k + |\xi_d|} 
\]

and

\[
\|Q_j v\|_{L^2}^2 \leq \left( \sum_k \int \varepsilon_k^2(\xi_d) d\xi_d \right) \left( \sum_k \int \frac{d\xi_d}{(2^j + 2^k + |\xi_d|)^2} \right) 
\]

\[
\lesssim \|\tilde{u}\|_{L^2}^2 \sum_k (2^j + 2^k)^{-1} \lesssim j 2^{-j} \|\tilde{u}\|_{L^2}^2. 
\]

Therefore,

\[
\|v\|_{L^2}^2 \lesssim \sum_j \|Q_j v\|_{L^2}^2 \lesssim \|\tilde{u}\|_{L^2}^2 = \|u\|_{L^2}^2
\]

and the lemma is proved. \( \square \)
Lemma 4.9. If \( a \in \Gamma_1^0 \) satisfies (4.19) then \( P_0^a \), and hence \( P_a \), are bounded from \( L^2 \) to \( \langle D_x' \rangle^{\frac{1}{2}} L^2 \) and from \( \langle D_x' \rangle^{\frac{1}{2}} L^2 \) to \( L^2 \). In particular, there is a constant \( C \) such that for all \( u \in L_2([0,T] \times \mathbb{R}^d) \)

\[
\| P_0^a u \|_{L^2([0,T] \times \mathbb{R}^d)} \leq C \| \langle D_x' \rangle^{\frac{1}{2}} u \|_{L^2([0,T] \times \mathbb{R}^d)}.
\]

Proof. Again, we omit the time variable in the proof. Because of (4.10), the symbol \( \sigma_a^0 \) satisfies

\[
|\partial_x^\beta \sigma_a^0(x', \xi', \xi_d)| \lesssim |\xi'|(|\langle \xi' \rangle + |\xi_d|)^{-1} |\langle \xi' \rangle|^{-|\beta|}.
\]

It also satisfies the para-differential spectral localization in the tangent variables \( x' \). This means that \( \rho(\cdot, \cdot, \xi_d) \) and \( \xi_d \rho(\cdot, \cdot, \xi_d) \) are bounded families of para-differential symbols in the class \( \Sigma_0^0 \) and \( \Sigma_1^1 \) respectively on \( \mathbb{R}^{d-1} \). Therefore for all \( v \) in \( L^2(\mathbb{R}^{d-1}) \) and all \( \xi_d \in \mathbb{R} \)

\[
\| \sigma_a^0(x', D_x', \xi_d) v \|_{L^2(\mathbb{R}^{d-1})} \lesssim \| v \|_{L^2(\mathbb{R}^{d-1})},
\]

\[
|\xi_d| \| \sigma_a^0(x', D_x', \xi_d) v \|_{H^{-1}(\mathbb{R}^{d-1})} \lesssim \| v \|_{L^2(\mathbb{R}^{d-1})}.
\]

Introduce a dyadic partition of unity in \( \mathbb{R}^{d-1} \) so that

\[
\tilde{u}(x', \xi_d) = \sum_j Q_j(D_x') \tilde{u}(x', \xi_d)
\]

where \( \tilde{u} \) denotes the partial Fourier transform of \( u \) in the variable \( x_d \). The spectrum in \( \xi' \) of \( Q_j \tilde{u} \) is contained in \( |\xi'| \approx 2^j \). Then \( f_j = \sigma_a^0(x', D_x', \xi_d) Q_j \tilde{u} \) has also its spectrum in a larger annulus but still of order \( |\xi'| \approx 2^j \). The estimates (4.19) imply that

\[
\| f_j(\cdot, \xi_d) \|_{L^2} \lesssim (1 + 2^{-j} |\xi_d|)^{-1} \varepsilon_j(\xi_d)
\]

where \( \varepsilon_j(\xi_d) = \| Q_j \tilde{u}(\cdot, \xi_d) \|_{L^2} \). Hence

\[
v = P_0^a u = \sum_j v_j(x'), \quad v_j(x') = \int f_j(x', \xi_d) d\xi_d
\]

and

\[
\| v_j \|_{L^2} \lesssim \int (1 + 2^{-j} |\xi_d|)^{-1} \varepsilon_j(\xi_d) d\xi_d \lesssim 2^{j/2} \| v \|_{L^2(\mathbb{R})}.
\]

Because the \( v_j \) are spectrally supported in annuli \( |\xi'| \approx 2^j \) one has

\[
\| v \|_{L^2}^2 \lesssim \sum_j \| v_j \|_{L^2}^2 \lesssim \sum_j 2^{j} \| v_j \|_{L^2(\mathbb{R})}^2 \approx \| \langle D_x' \rangle^{\frac{1}{2}} w \|_{L^2(\mathbb{R}^d)}^2
\]

and the lemma is proved. \( \Box \)
4.4 Energy balance for the IBVP

Consider a system \( L = \partial_t + \sum A_j \partial x_j \) with Lipschitz coefficients. We assume that it admits a symmetrizer \( S(t, x, \xi) \), which is Lipschitz continuous in \((t, \xi)\), homogeneous of degree 0 and \( C^\infty \) in \( \xi \neq 0 \). In particular the estimate \((\ref{en:bnd})\) applies to solutions \( u \in C^0([0, T]; L^2(\mathbb{R}^d)) \) of \( Lu = f \in L^2([0, T] \times \mathbb{R}^d) \) since \( \|Lu - (\partial_t + iT)u\|_{L^2} \lesssim \|u\|_{L^2} \) where \( A \) denotes here the symbol \( \sum \xi_j A_j(t, x) \).

Following ideas from \cite{FriLa2}, we want to obtain an estimate similar to \((\ref{en:bnd})\) for smooth functions defined on the domain \( \{x_d \geq 0\} \). But in contrast with \cite{FriLa2}, we assume that we already have a control of \( \|u|_{x_d=0}\|_{L^2} \), for instance given by a preliminary use of Kreiss symmetrizers. Boundary term occurs and we analyze them below. An important data is the value of \( S \) on the conormal to the boundary and we introduce

\[
S_\infty(t, x) = S(t, x, (0, \ldots, 0, 1)).
\]

**Proposition 4.10.** There is a constant \( C \) such that for \( u \in C^\infty_0([0, T] \times \mathbb{R}^d_+) \) one has

\[
\|u(t)\|_{L^2(\mathbb{R}^d_+)}^2 \lesssim \|u(0)\|_{L^2(\mathbb{R}^d_+)}^2 + \|u\|_{L^2([0, T] \times \mathbb{R}^d_+)}^2 + \|f\|_{L^2([0, T] \times \mathbb{R}^d_+)}^2
\]

\[
+ \|g\|_{L^2([0, T] \times \mathbb{R}^d-1)}^2 + \|P_{S_1} \tilde{u}\|_{L^2([0, T] \times \mathbb{R}^d-1)}^2
\]

where \( f = Lu, \ g = u|_{x_d=0} \) and \( P_{S_1} \) is the trace operator associated to the symbol \( S_1 = \theta(S - S_\infty) \) and \( \tilde{u} \) denotes the extension of \( u \) by 0 for \( x_d < 0 \).

**Proof.** Consider \( u \in C^\infty_0([0, T] \times \mathbb{R}^d_+) \) and introduce \( f = Lu \in C^\infty_0([0, T] \times \mathbb{R}^d_+) \) and \( g = u|_{x_d=0} \in C^\infty_0([0, T] \times \mathbb{R}^{d-1}) \). Let \( \chi \in C^\infty(\mathbb{R}) \) with support in \([0, \infty[\) and equal to 1 on \([1, \infty[\). Let \( \chi_\varepsilon(x_d) = \chi(x_d/\varepsilon) \) and \( u_\varepsilon = \chi_\varepsilon u \). Then

\[
Lu_\varepsilon = f_\varepsilon + f_\varepsilon^1, \quad f_\varepsilon = \chi_\varepsilon f, \quad f_\varepsilon^1 = \chi_\varepsilon Adu = \varepsilon^{-1} \chi'(x_d/\varepsilon) Adu.
\]

We apply the energy estimate \((\ref{en:bnd})\) to \( u_\varepsilon \) and pass to the limit. The difficulty is concentrated in the term

\[
I_\varepsilon = I_S(f_\varepsilon^1, u_\varepsilon) = \int_0^t \langle Sf_\varepsilon^1(t'), u_\varepsilon(t') \rangle_{L^2(\mathbb{R}^d)} dt'.
\]

The proposition will follow from the estimate

\[
\limsup_{\varepsilon \to 0} |I_\varepsilon| \lesssim \|u\|_{L^2(\mathbb{R}^d)}^2 + \|f\|_{L^2([0, T] \times \mathbb{R}^d)}^2 + \|g\|_{L^2([0, T] \times \mathbb{R}^d-1)}^2 + \|P_{S_1} \tilde{u}\|_{L^2([0, T] \times \mathbb{R}^d-1)}^2.
\]
a) For \( u \) smooth, \( f^1_\varepsilon = g_1 \chi'_\varepsilon + h^1_\varepsilon \) with \( g_1 = A_d u|_{x=0} \) and \( \|h^1_\varepsilon\|_{L^2} \lesssim \varepsilon^{1/2} (\|u(t)\|_{L^2} + \|\partial_x u(t)\|_{L^2}) \). Therefore is sufficient to prove (4.23) for \( I_S(h_\varepsilon, u_\varepsilon) \) with \( h_\varepsilon := g_1 \chi'_\varepsilon \).

b) The spatial Fourier transform of \( h_\varepsilon \) is \( \hat{h}_\varepsilon(t) = \hat{\chi}'(\varepsilon \xi) \hat{g}_1(t, \xi) \). Since \( \chi' \in S(\mathbb{R}) \),

\[
\|h_\varepsilon\|_{L^2([0,T]; H^{-1}(\mathbb{R}^d))} \lesssim \|g\|_{L^2}.
\]

Recalling the definition (4.24), note that \( T_{\delta S} - (T_{\delta S})^* \) and hence \( S - (T_{\delta S})^* \) are of degree \(-1\). This implies that

\[
\|S h_\varepsilon - T^*_{\delta S} h_\varepsilon\|_{L^2} \lesssim \|g\|_{L^2}.
\]

Therefore we are reduced to prove (4.23) with

\[
I_\varepsilon = \tilde{I}_S(h_\varepsilon, u_\varepsilon) = \int_0^t (h_\varepsilon, T_{\delta S} u_\varepsilon)_{L^2(\mathbb{R}^d)} dt'.
\]

We split \( S \) into \( S_\infty + (S - S_\infty) \) and we study each term separately.

c) The operator \( T_{\delta S_\infty} - S_\infty \) is of degree \(-1\), therefore \( T_{\delta S_\infty} u_\varepsilon - S_\infty u_\varepsilon \) is bounded in \( L^2([0,T]; H^1) \). Therefore

\[
\tilde{I}_{S_\infty}(h_\varepsilon, u_\varepsilon) = \int_0^t (g_1 \chi'_\varepsilon, S_\infty \chi_\varepsilon u) dt' + O(\|g\|_{L^2} \|u\|_{L^2})
\]

For \( u \) smooth,

\[
\lim_{\varepsilon \to 0} \int_0^t (g_1 \chi'_\varepsilon, S_\infty \chi_\varepsilon u) dt' = \frac{1}{2} (g_1, S_\infty|_{x=0} g)_{L^2([0,T] \times \mathbb{R}^{d-1})}
\]

and therefore

\[
\text{(4.26)} \quad \limsup_{\varepsilon \to 0} \tilde{I}_{S_\infty}(h_\varepsilon, u_\varepsilon) \lesssim \|g\|^2 + \|u\|^2.
\]

d) We now show, using the notation (4.27), that

\[
\lim_{\varepsilon \to 0} \tilde{I}_{S - S_\infty}(h_\varepsilon, u_\varepsilon) = (g_1, P_{S_\infty} u)_{L^2([0,T] \times \mathbb{R}^{d-1})}
\]

and this will finish the proof of the proposition.

Using that \( S \) is homogeneous of degree 0, we can write

\[
\text{(4.28)} \quad S_1(t, x, \xi) = \theta(\xi)(S - S_\infty)(t, x, \xi) = \sum_{j=1}^{d-1} S_{1,j}(t, x, \xi) \xi_j
\]
where the $S_{1,j}$ are of degree $-1$. Hence,

$$\|T_{S_1} \chi_{\xi} u\|_{L^2([0,T], H^1)} \lesssim \sum_{j=1}^{d-1} \|\partial_{x_j} u\|_{L^2}.$$  

Moreover, $T_{S_1} \chi_{\xi} u \to T_{S_1} \tilde{u}$ in $L^2([0,T], H^1)$ and the trace on $\{x_d = 0\}$ is well defined. Using (4.24) and the convergence $g_1 \chi'_{\xi} \to g_1 \otimes \delta_{x_d = 0}$ in $L^2([0,T], H^{-1})$, this implies that

$$(g_1 \chi'_{\xi}, T_{S_1} u_{\xi}) \to (g_1, (T_{S_1} \tilde{u})|_{x_d = 0})$$

that is (4.28) and the proposition is proved.

**Proposition 4.11.** If $S(t, x, \xi)$ admits a bounded holomorphic extension in the cone $\{|\text{Im} \xi| \leq \delta |\xi|\}$ for some $\delta > 0$. Then, there are constant $C$ and $\delta_1 > 0$ such that for all $\delta' \in [0, \delta_1]$ and all $u \in C_0^\infty([0,T] \times \mathbb{R}^d)$

$$\|P_{S_1} \tilde{u}\|_{L^2([0,T] \times \mathbb{R}^{d-1})} \leq C \left( \|\langle D_{x'}\rangle^{\frac{1}{2}} e^{-\beta x_d (D_{x'})} u\|_{L^2([0,T] \times \mathbb{R}^d)} + \|u\|_{L^2([0,T] \times \mathbb{R}^d)} \right).$$

**Proof.** To simplify notations, we omit the variable $t$ which is just a parameter. Note that when $\delta' = 0$, (4.29) is simply (4.18).

First, we remark that one can replace $S_1$ by

$$S_2 = \theta(\xi', 0)(S - S_{\infty})$$

because $S_1 - S_2$ is supported in $|\xi'| < 2$ and thus of order $-1$ by (4.28).

According to Proposition 4.5, we can replace $P_{S_2}$ by $P_{S_2}^0$ and $v = P_{S_1} \tilde{u}$ is given by

$$v(x') = \int e^{ix' \cdot \xi'} \sigma(x', \xi', d) \hat{u}(\xi) d\xi$$

where $\hat{u}$ is the spatial Fourier transform of $\tilde{u}$ and

$$\sigma(t, x', \xi', d) = \int G^0(x' - y', \xi') S_2(y', 0, \xi, d) dy'.$$

By assumption, $S - S_{\infty}$, and thus $S_2$, have holomorphic extensions in $\xi_d$ to the domain $\{|\text{Im} \xi_d| \leq \delta |\xi|\}$ and this extension vanishes when $\xi' = 0$.

Hence, by homogeneity, it

$$|(S - S_{\infty})(x, \xi)| \lesssim \frac{|\xi'|}{|\xi|}.$$
Therefore, since $\sigma$ vanishes for $\xi'$ small, $\sigma$, has an holomorphic extension in $\xi_d$ to a domain $\{\text{Im} \xi_d \leq \delta_1(\langle \xi' \rangle)\}$. Moreover, because $u$ is supported in $\{x_d \geq 0\}$, its Fourier transform $\hat{u}$ is holomorphic in $\{\text{Im} \xi_d < 0\}$. Therefore, one can shift the integration path in $\xi_d$ to $\mathbb{R} - i\delta \langle \xi' \rangle$

\[ v(x') = \int e^{ix' \cdot \xi'} \sigma(x', \xi', \xi_d - i\delta \langle \xi' \rangle) \hat{u}(\xi', \xi_d - i\delta \langle \xi' \rangle) d\xi \]

\[ = \int e^{ix' \cdot \xi'} \rho(x', \xi', \xi_d) \hat{w}(\xi', \xi_d) d\xi \]

\[ = \int \rho(x', D_{x'} \xi_d) \tilde{w}(x', \xi_d) d\xi_d \]

with $\rho(x', \xi', \xi_d) = \sigma(x', \xi', \xi_d - i\delta \langle \xi' \rangle)$, $w = e^{-\delta' x_d (D_{x'})} u$, and $\tilde{w}$ denoting its partial Fourier transform in the variable $x_d$. We conclude by applying (H.18) to $\rho$ and $w$.

4.5 Elliptic estimates

Consider a system

(4.31) \[ \partial_{x_d} u + iT^{45} u = f \]

where $A$ is a matrix with coefficients in $\Gamma^1$. 

**Proposition 4.12.** Suppose that the spectrum of $A(t, x, \zeta)$ is contained in $|\text{Im} \lambda| \geq c|\zeta|$. Then there is a constant $C$ such that

\[ \| \langle D_z \rangle^{\frac{1}{2}} u \|_{L^2} \leq C (\| u \|_{L^2} + \| f \|_{L^2} + \| u|_{x_d=0} \|_{L^2}) \]

**Proof.** This is a special case of the tangential analysis (see e.g. [Kre, ChPi, Met5]). The assumption implies that the matrix $A$ is conjugated to a block diagonal matrix with blocks $A_\pm$ having their spectrum in $\{ \pm \text{Im} \lambda \geq c|\zeta| \}$. Each block has a symmetrizer, and there is a symmetrizer $S = S^*$ such that $\text{Im} SA \geq c'|\zeta|$. 

We will use the following extension of this estimate.

**Proposition 4.13.** Suppose that the spectrum of $A(t, x, \zeta)$ is contained in $|\text{Im} \lambda + \delta|\zeta| \geq c|\zeta|$ for some $\delta \in [0, 1]$. Then there is a constant $C$ such that

\[ \| \langle D_z \rangle^{\frac{1}{2}} e^{-\delta (D_z) x_d} u \|_{L^2} \leq C (\| u \|_{L^2} + \| f \|_{L^2} + \| u|_{x_d=0} \|_{L^2}) \]
Proof. The symbols in \( \zeta \mapsto e^{-\delta x_d \langle \zeta \rangle} \) form a bounded family of symbols of degree zero. Therefore, the commutator \([e^{-\delta (D_z)x_d}, T_A]\) are uniformly bounded in \( L^2 \). This shows that \( v = e^{-\delta (D_z)x_d} u \) satisfies
\[
\| \partial_v + (T_{tg} + i \delta \langle D_z \rangle)v \|_{L^2} \leq C(\| u \|_{L^2} + \| f \|_{L^2}).
\]
The symbol of \( T_{tg} + i \delta \langle D_z \rangle \) is \( A(t) + i \delta \langle \zeta \rangle \) and its spectrum is contained in \( |\text{Im } \lambda| \geq c' |\zeta| \). Hence one can apply Proposition 4.12 to \( v \) and the estimate follows since \( v|_{x_d=0} = u|_{x_d=0} \).

5 Semi group estimates and the IBVP in \( L^2 \)

The goal of this section is to solve the initial boundary value problem (1.1) and prove Theorem 1.6. We first review the analysis of the boundary value problem and next show what has to be added to treat initial data in \( L^2 \).

5.1 The main steps

Consider an hyperbolic system \( L \) on \( \{ x_d \geq 0 \} \)
\[
L = A_0(t,x) \partial_t + \sum_{j=1}^d A_j(t,x) \partial_{x_j} + B(t,x)
\]
together with boundary conditions \( M \) on \( \{ x_d = 0 \} \). The adjoint operator \( L^* \) is
\[
L^* = -(A_0(t,x))^* \partial_t - \sum_{j=1}^d A_j^*(t,x) \partial_{x_j} + B_1(t,x)
\]
where \( B_1 = -B^* + \partial_t A_0 + \sum \partial_j A_j^* \). There are adjoint boundary conditions \( M' \) for \( L^* \) such that for all smooth enough functions \( u \) and \( v \) on \( [a,b] \times \mathbb{R}^d_+ \):
\[
(Lu,v) - (u,L^*v) = (Mu|_{x_d=0}, M_1 v|_{x_d=0}) - (M_1' u|_{x_d=0}, M' v|_{x_d=0})
+ (u|_{t=b}, v|_{t=b}) - (u|_{t=a}, v|_{t=a}).
\]
for some boundary matrices \( M_1 \) and \( M_1' \). Here \((\cdot,\cdot)\) denotes the \( L^2 \) scalar products on the appropriate domains. The formula extends to unbounded time intervals. The matrices \( M_1, M', M_1' \) are not unique but the invariant key property is that
\[
\ker M' = (A_d \ker M)^\perp.
\]
Introduce the spaces \( L^2 = e^{\gamma t} L^2 \) with norms \( \|u\|_{L^2} = \|e^{-\gamma t}u\|_{L^2} \) and note that, in the identity (5.3), \((\cdot, \cdot)\) can be understood as well as the duality \( L^2 \times L^2_{-\gamma} \). We suppose here that the Kreiss estimates have already been proven and take them as an assumption.

**Assumption 5.1.** The following a priori estimates are valid : for \( \gamma \geq \gamma_0 \)

\[
\gamma \|u\|_{L^2}^2 + \|u|_{x=d=0}\|_{L^2}^2 \lesssim \gamma^{-1} \|Lu\|_{L^2}^2 + \|Mu|_{x=d=0}\|_{L^2}^2
\]

and

\[
\gamma \|v\|_{L^2_{-\gamma}}^2 + \|v|_{x=d=0}\|_{L^2_{-\gamma}}^2 \lesssim \gamma^{-1} \|L^*u\|_{L^2_{-\gamma}}^2 + \|M'v|_{x=d=0}\|_{L^2_{-\gamma}}^2.
\]

In particular, this applies to systems in the class \( sM \).

**Proposition 5.2** ([Kre, ChPi]). Under Assumption 5.1 the boundary value problem

\[
Lu = f, \quad Bu|_{x=d=0} = g
\]

is well posed in \( L^2_{\gamma} \) for \( \gamma \geq \gamma_0 \).

Indeed, (5.6) implies that (5.7) has a weak solution in \( L^2_{\gamma} \). By tangential smoothing and Friedrichs Lemma, this solution is a strong solution and therefore satisfies (5.5). In particular, this implies uniqueness of weak solution. Moreover, the causality principle is satisfied: if \( f \) and \( g \) vanish for \( t \leq t_0 \), then \( u \) also vanishes for \( t \leq t_0 \).

We now consider the initial-boundary value problem

\[
Lu = f, \quad Mu|_{x=d=0} = g, \quad u|_{t=0} = u_0.
\]

**Proposition 5.3.** The problem (5.8) is well posed in \( L^2_{\gamma_0} \) when \( u_0 = 0 \).

Proof. Existence is obtained by extending \( f \) and \( g \) by 0 for \( t < 0 \). Then there is a solution \( u \in L^2_{\gamma_0} \) and the causality principle implies that it vanishes when \( t < 0 \). Therefore, its trace \( u|_{t=0} \) also vanishes. Note that the trace is well defined in \( H^{1/2}_{loc} \) since the equation is non characteristic in time. Uniqueness follows in the same way : if \( f = 0 \) and \( g = 0 \), the extension \( \tilde{u} \) of \( u \) by 0 in the past is a weak solution of \( L\tilde{u} = 0, Mu = 0 \) and therefore vanishes. \( \square \)
This result settles the question of uniqueness of solutions for $L_{bvp}$. The existence part is easy when the data vanish on a neighborhood of the edge $\{t = x_d = 0\}$.

**Proposition 5.4.** If $u_0 \in L^2$, $f \in L^2_{\gamma_0}$ and $g \in L^2_{\gamma_0}$ vanish on a neighborhood of $\{t = x_d = 0\}$, then $(5.8)$ has a unique solution $u \in L^2_{\gamma_0}$.

Moreover, if in addition $u_0$, $f$ and $g$ belong to $H^1$, the solution $u$ also belongs to $H^1$.

**Proof.** Extend $u_0$ and $f$ by 0 for $x_d \leq 0$ and solve the Cauchy problem $Lw = \tilde{f}$, $v|_{t=0} = \tilde{u}_0$. Then there is a unique solution $v \in C^0([0,1]; L^2(\mathbb{R}^d))$, which by finite speed of propagation vanishes for $x_d \leq 0$ and $t \leq t_0$ for some $t_0 > 0$. We solve the problem for $w = u - \chi v$, where $\chi(t)$ is supported in $[0, t_0]$ and $\chi(0) = 1$:

$$Lw = (1 - \chi)f - A_0 \partial_t \chi v, \quad Mw|_{x_d = 0} = g, \quad w|_{t = 0} = 0.$$  

Indeed, by Proposition 5.3 there is a solution $w \in L^2_{\gamma_0}$.

The $H^1$ smoothness is proved similarly taking $H^1$ extensions of $u_0$ and $f$, which vanish near the edge and an $H^1$ extension of $g$ which vanish in the past.

The difficult part of the proof is now to prove estimates for $u$ independent of the neighborhood where the data vanish. We prove them under the following assumption:

**Assumption 5.5.** $L$ admits a symmetrizer $S(t, x, \xi)$ which is Lipschitz continuous in $(t, x)$, and real analytic in $\xi$.

**Theorem 5.6.** Under Assumptions 5.1 and 5.5, there is a constant $C$ such that for all smooth $u_0$, $f$ and $g$ which vanish on a neighborhood of $\{t = x_d = 0\}$, the unique $H^1$ solution of $(5.8)$ satisfies

$$\|u(t)\|_{L^2} + \|u|_{x_d = 0}\|_{L^2([0, t] \times \mathbb{R}^{d-1})} \leq C \left( \|u_0\|_{L^2} + \|g\|_{L^2([0, t] \times \mathbb{R}^{d-1})} \int_0^t \|f(s)\|_{L^2(\mathbb{R}^d)} ds \right).$$

By density-continuity, the mapping $(u_0, f, g) \mapsto u$ uniquely extends to $u_0 \in L^2$, $f \in L^1([0, T]; L^2)$ and $g \in L^2$. Then $u \in C^0([0, T]; L^2)$, is a weak solution of $(5.8)$. Since uniqueness is already known, the theorem above implies the next corollary and hence Theorem 5.6.

**Corollary 5.7.** Under Assumptions 5.1 and 5.5, for all $u_0 \in L^2(\mathbb{R}^d_+)$, $f \in L^1([0, T]; L^2)$ and $g \in L^2([0, T] \times \mathbb{R}^{d-1})$, there is a unique $u \in C^0([0, T]; L^2(\mathbb{R}^d_+))$ solution of $(5.8)$ on $[0, T] \times \mathbb{R}^d_+$. Moreover, $u$ satisfies $(6.9)$.  

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5.2 The main estimate

We reduce the proof of the estimate (5.9) to a simpler one:

**Theorem 5.8.** Under Assumption 5.5, for \( u \in H^1(\mathbb{R}^{1+d}) \) with support in \( t \in [0, 2] \) one has

\[
\left\| u(t) \right\|_{L^2(\mathbb{R}^d_+)} \lesssim \| Lu \|_{L^2} + \| u \|_{L^2} + \| u|_{x_d=0} \|_{L^2}.
\]

Taking this theorem for granted, we show that it implies Theorem 5.6. Because \((SA_0)^{-1}(A_0^*)^{-1}\) is a symmetrizer for \(L^*\), one has similar estimates for the adjoint problem:

\[
\left\| v(t) \right\|_{L^2(\mathbb{R}^d_+)} \lesssim \| L^* v \|_{L^2} + \| v \|_{L^2} + \| v|_{x_d=0} \|_{L^2}.
\]

**Corollary 5.9.** Consider the backward initial boundary value problem for \( t \leq 1 \)

\[
L^* \Phi = \varphi, \quad M' \varphi|_{x_d=0} = \psi, \quad \Phi|_{t=1} = 0.
\]

Then

\[
\left\| \Phi|_{t=0} \right\|_{L^2} + \left\| \Phi \right\|_{L^2([0,1]×\mathbb{R}^d_+)} + \left\| \Phi|_{x_d=0} \right\|_{L^2} \lesssim \left\| \varphi \right\|_{L^2([0,1]×\mathbb{R}^d_+)} + \left\| \psi \right\|_{L^2}.
\]

By duality, this implies the following estimate for the direct problem:

**Proposition 5.10.** There is a constant \( C \) such that for data vanishing on a neighborhood of the edge, the solutions of (5.8) satisfy

\[
\| u|_{t=1} \|_{L^2} + \| u \|_{L^2([0,1]×\mathbb{R}^d_+)} + \| u|_{x_d=0} \|_{L^2([0,1]×\mathbb{R}^{d-1})} \lesssim \| u_0 \|_{L^2(\mathbb{R}^d_+)} + \| f \|_{L^2([0,1]×\mathbb{R}^d_+)} + \| g \|_{L^2([0,1]×\mathbb{R}^{d-1})}.
\]

**Proof.** By duality, the corollary implies that

\[
\left\| u \right\|_{L^2([0,1]×\mathbb{R}^d_+)} + \left\| u|_{x_d=0} \right\|_{L^2} \lesssim \left\| u_0 \right\|_{L^2} + \| f \|_{L^2([0,1]×\mathbb{R}^d_+)} + \| g \|_{L^2}.
\]

To get the missing term, that is the \( L^2 \) norm of \( u|_{t=1} \), it is now sufficient to apply the direct estimate of Theorem 5.8 to \( v = tu \), since \( Lv = tLu + A_0u \) is now controlled in \( L^2 \). \( \square \)
We pass from the estimate for time 1 to estimates for all time \( t \in [0, 1] \) by scaling. For \( \varepsilon \in [0, 1] \) consider the systems

\[
L_{\varepsilon}(t, x, \partial_t, \partial_x) = A_0(\varepsilon t, \varepsilon x)\partial_t + \sum A_j(\varepsilon t, \varepsilon x)\partial_{x_j} + \varepsilon E(\varepsilon t, \varepsilon x)
\]

and the boundary conditions

\[
M_{\varepsilon}(t, x) = M(\varepsilon t, \varepsilon x).
\]

**Lemma 5.11.** If \( u \) is a solution of (5.8), then \( u_{\varepsilon}(t, x) = u(\varepsilon t, \varepsilon x) \) satisfies

\[
L_{\varepsilon} u_{\varepsilon} = \varepsilon f_{\varepsilon}, \quad B_{\varepsilon} u_{\varepsilon | x_d=0} = g_{\varepsilon}, \quad u_{\varepsilon | t=0} = u_{0, \varepsilon}
\]

where \( f_{\varepsilon}, g_{\varepsilon} \) and \( u_{0, \varepsilon} \) are deduced from \( f, g \) and \( u_0 \) by the scaling.

**Proof.** One has

\[
A_j(\varepsilon t, \varepsilon x)(\partial_{x_j} u_{\varepsilon})(t, x) = \varepsilon (A_j \partial_{x_j} u)(\varepsilon t, \varepsilon x)
\]

and similar formulas for the traces. \( \square \)

We note that the Assumptions 5.1 are satisfied for all \( \varepsilon \in [0, 1] \), with uniform constants:

**Lemma 5.12.** The boundary value problems \((L_{\varepsilon}, B_{\varepsilon})\) satisfy the estimates (5.3) and (5.6) with constants independent of \( \varepsilon \in [0, 1] \), for \( \gamma \geq \varepsilon \gamma_0 \).

**Proof.** With \( \gamma' = \gamma/\varepsilon \geq \gamma_0 \), the direct estimates are immediate consequences of the scaling identities

\[
\gamma \|u_{\varepsilon}\|_{L^2_{\gamma'}}^2 = \varepsilon^{-d} \gamma' \|u\|_{L^2_{\gamma'}}^2, \quad \gamma^{-1} \|\varepsilon f_{\varepsilon}\|_{L^2_{\gamma'}}^2 = \varepsilon^{-d} (\gamma')^{-1} \|f\|_{L^2_{\gamma'}}^2
\]

and

\[
\|u_{\varepsilon | x_d=0}\|_{L^2_{\gamma'}}^2 = \varepsilon^{-d} \|u_{| x_d=0}\|_{L^2_{\gamma'}}^2, \quad \|g_{\varepsilon}\|_{L^2_{\gamma'}}^2 = \varepsilon^{-d} \|g\|_{L^2_{\gamma'}}^2
\]

The adjoint operator \((L_{\varepsilon})^*\) is the scaled operator \((L^*)_{\varepsilon}\) deduces from \( L^* \) as in (5.16). Similarly, \( B'_{\varepsilon} = B'(\varepsilon t, \varepsilon x) \) are dual boundary conditions so that the estimates for \( L^*_\varepsilon \) follow by the same scaling argument. \( \square \)

In the same vein, considering the symmetrizers \( S_{\varepsilon}(t, x, \xi) = S(\varepsilon t, \varepsilon x, \xi) \) implies that the Assumption 5.5 is satisfied for \( L_{\varepsilon} \), with uniform constants.
Proposition 5.13. There is a constant $C$ such that for data vanishing on a neighborhood of the edge, the solutions of (5.8) are continuous in time with values in $L^2$ satisfy for $t \in [0,1]$

$$\|u(t)\|_{L^2} + \|u_{x_d=0}\|_{L^2([0,t]\times\mathbb{R}^{d-1})} \lesssim \|u_0\|_{L^2(\mathbb{R}^d_+)} + \|f\|_{L^2([0,t]\times\mathbb{R}^d_+)} + \|g\|_{L^2([0,t]\times\mathbb{R}^{d-1})}.$$

Proof. The estimates at time $t_1$ follows from (5.14) applied to $u_\varepsilon$ with $\varepsilon = t_1$. When the data are $H^1$, the solution is $H^1$ and therefore continuous in time with values in $L^2$. Therefore, by density the solutions belong to $C_0([0,1];L^2)$.

This is almost the desired estimate (5.9), except for the norm of $f$. It remains to replace the $L^2$ norm above by an $L^1([0,1],L^2)$ norm. For that we split the problem into two pieces:

**hominvp** (5.20) \hspace{1cm} Lu = f, \quad Mu|_{x_d=0} = 0, \quad u|_{t=0} = 0.

and

**hominvp** (5.21) \hspace{1cm} Lu = 0, \quad Mu|_{x_d=0} = g, \quad u|_{t=0} = u_0.

By linearity, it is sufficient to prove (5.9) for the solution of each problem separately. For the second equation, this follows directly from Proposition 5.13 and it remains to prove (5.9) for the solution of (5.20). We show that it follows from (5.19) using Duhamel’s principle.

Proposition 5.14. There is a family of bounded operators $\mathcal{E}(t,s)$ from $L^2(\mathbb{R}^d_+)$ to $L^2(\mathbb{R}^d_+)$, for $0 \leq s \leq t \leq 1$, such that for all $s \in [0,1]$, $u(t) = \mathcal{E}(t,s)u_0$ is the unique solution in $C^0([s,1],L^2(\mathbb{R}^d_+))$ of

$$Lu = 0, \quad Mu|_{x_d=0} = g, \quad u|_{t=0} = u_0.$$

In particular, for all $u_0 \in L^2(\mathbb{R}^d_+)$, $t \mapsto \mathcal{E}(t,s)u_0$ belongs to $C^0([s,1],L^2(\mathbb{R}^d_+))$.

Moreover, for all $u_0 \in L^2(\mathbb{R}^d_+)$, $s \mapsto \mathcal{E}(t,s)u_0$ belongs to $C^0([0,t],L^2_w(\mathbb{R}^d_+))$ where $L^2_w(\mathbb{R}^d_+)$ denotes the space $L^2$ equipped with the weak topology.

Proof. Clearly, what we have done before for the initial time $t = 0$ is true for all initial time $t = s$. Thus, Proposition 5.13 implies that when $u_0 \in L^2$ vanishes near the boundary, there is a unique solution $u \in C^0([s,1];L^2)$ which satisfies

$$\|u(t)\|_{L^2} \leq C\|u_0\|_{L^2}$$
The operator \( u_0 \mapsto u \) extends \( b \) density to \( u_0 \in L^2 \) implying the first part of the proposition. The second follows by duality: the corresponding operator \( \mathcal{F}(t, s') \) for the backward transposed problem is defined for \( 0 \leq t \leq s' \leq 1 \) and \( v = \mathcal{F}(\cdot, s') v_0 \) solves
\[
L^* v = 0, \quad M'u_{|x_d=0} = 0, \quad u_{|t=s'} = v_0.
\]
In particular, \( t \mapsto \mathcal{F}(t, s') v \) is continuous from \([0, s']\) to \( L^2(\mathbb{R}_+^d) \). The duality relation (5.3) shows that \( \mathcal{E}(t, s) = \mathcal{F}(s, t)^* \) and therefore \( s \mapsto (\mathcal{E}(t, s) u_0, v_0) \) is continuous for all \( u_0 \) and \( v_0 \) in \( L^2 \).

**Lemma 5.15.** For \( f \) smooth, vanishing in a neighborhood of the edge, the solution of (5.20) is given by Duhamel’s principle:

\[
(5.22) \quad u(t) = \int_0^t \mathcal{E}(t, s) f(s) ds.
\]

**Proof.** Note that for \( f \in \mathcal{C}^0([a, 1]; L^2(\mathbb{R}_+^d)) \), \( s \mapsto \mathcal{E}(t, s) f(s) \) is continuous from \([0, t]\) to \( L^2_w \) so that the integral (5.22) makes sense. Denote it by \( \tilde{u}(t) \).

For \( \psi \in H^1(\mathbb{R}_+^d) \) vanishing near \( x_d = 0 \), let \( \Psi(\cdot) = \mathcal{F}(\cdot, t) \psi \) which is a \( H^1 \) solution on \([0, t] \times \mathbb{R}_+^d\) of
\[
L^* \Psi = 0, \quad B'\Psi_{|x_d=0} = 0, \quad \Psi(t) = \psi.
\]
Then
\[
(\tilde{u}(t), \psi) = \int_0^t (\mathcal{E}(t, s) f(s), \psi) ds = \int_0^t (f(s), \mathcal{F}(s, t) \psi) ds = (Lu, \Psi)_{L^2([0,t] \times \mathbb{R}_+^d)} = (u(t), \psi)
\]
where the last equality follows from (5.3), which is satisfied since \( u \) is \( H^1 \). Hence \( \tilde{u}(t) = u(t) \) and the lemma is proved.

Using the estimates of Proposition 5.13 for \( \mathcal{E}(\cdot, s) f(s) \) and integrating them in \( s \) implies

**Corollary 5.16.** For \( f \) smooth, vanishing in a neighborhood of the edge, the solution of (5.20) satisfies

\[
(5.23) \quad \|u(t)\|_{L^2} + \|u_{|x_d=0}\|_{L^2([0,t] \times \mathbb{R}_+^{d-1})} \lesssim \int_0^t \|f(s)\|_{L^2(\mathbb{R}_+^d)} ds
\]

This finishes the proof of Theorem 5.6.
5.3 Proof of Theorem 5.8

Recall that we are considering a function \( u \in H^1(\mathbb{R}^{1+d}_+) \), supported in \( t \in [0,2] \). We can paralinearize the operator \( A^{-1}_d L = \partial_{x_d} + G \) and write

\[
\partial_{x_d} u + T_G u = f, \tag{5.24}
\]

where \( T_G \) denotes the tangential paradifferential operator of symbol \( iG(t,x,\zeta) \) with \( \zeta = (\tau,\xi') \in \mathbb{R} \times \mathbb{R}^{d-1} \), and \( f \) satisfies

\[
\| f \|_{L^2} \lesssim \| Lu \|_{L^2} + \| u \|_{L^2}. \tag{5.25}
\]

Consider a finite microlocal partition of unity

\[
1 = \chi_0(\zeta) + \sum_{k=1} \chi_k(\zeta) \tag{5.26}
\]

with \( \chi_0 \) is supported in \( \{ |\zeta'| \leq c |\tau| \} \) while the \( \chi_k \) for \( k \geq 1 \) are supported in \( \{ |\tau| \leq 2c^{-1} |\zeta'| \} \). Let \( u_k = \chi_k(D_z) u \), where \( z = (t,x') \). We will estimate the \( L^2 \) norm of each \( u_k(t) \) separately, using different methods according to \( k = 0 \) or \( k \geq 1 \). Note that

\[
\partial_{x_d} u_k + T_G \tilde{\chi}_k u_k = f_k, \tag{5.27}
\]

where \( f_k \) satisfies (5.25) and \( \tilde{\chi}_k \) is equal to one on the support of \( \chi_k \). Note also that \( u_k \) is not any more supported in \( t \in [0,2] \), but \( u_k \) has an \( H^1 \) norm for \( t \notin [-1,3] \) controlled by the \( L^2 \) norm of \( u \). In particular

\[
\| u_k|_{t=-1} \|_{L^2} + \| u_k|_{t=3} \|_{L^2} \lesssim \| u \|_{L^2}. \tag{5.28}
\]

We prove that one can choose the partition so that the \( u_k \) satisfy

\[
\| u_k(t) \|_{L^2} \lesssim \| u_k \|_{L^2} + \| f_k \|_{L^2} + \| u_k|_{x_d=0} \|_{L^2} + \| u_k(-1) \|_{L^2}. \tag{5.29}
\]

Theorem 5.8 is a consequence of the following two results:

**Proposition 5.17.** One can choose \( c > 0 \) such that if \( \chi_0 \) is supported in \( \{ |\zeta'| \leq c |\tau| \} \) then \( u_0 = \chi(D_z) u \) satisfies (5.29).

**Proposition 5.18.** For all \( z = (t,x') \) and all \( \zeta = (\tau,\xi') \in \mathbb{R}^d \) with \( |\zeta'| = 1 \), there is a conical neighborhood of \((z,\zeta)\) such that if \( \chi_k \) is supported in this neighborhood, \( u_k = \chi_k(D_z) u \) satisfies (5.29).

Indeed, by compactness, one can choose a partition of unity (5.26) such that the estimate (5.29) is satisfied for all \( u_k = \chi_k(D_z) u \).
5.3.1 The case $|\xi'| \leq c|\tau|$

The proof of Proposition 5.17 is based on an extension of $u_0$ to $\{x_d < 0\}$. We first make a block reduction of the symbol $G(t,x,\zeta)$ for large $\tau$. If $c$ is small enough, the cone $\{|\xi'| \leq c|\tau|\}$ is contained in the interior of the cone $\Gamma^\delta$ and therefore, by Proposition 2.4:

**Lemma 5.19.** If $c$ is small enough, the eigenvalues of $G(t,x,\zeta)$ in $\{|\xi'| \leq c|\tau|\}$ are real and split in two groups, located in $\{\pm \lambda \geq c|\tau|\}$ respectively.

As a corollary, there is a smooth microlocal block reduction

$$\tilde{\chi}_0(\zeta)G(t,x,\zeta) = V^{-1}G_1V, \quad G_1 = \begin{pmatrix} G_+ & 0 \\ 0 & G_- \end{pmatrix}$$

where the eigenvalues of $G_{\pm}$ are located in $\{ |\lambda| \geq c|\tau| \}$ with $\pm \lambda \tau > 0$.

If $\chi(D_x)$ is supported in $\Gamma^\delta$ and $u_0 = \chi(D_x)u$, on can split

$$v := T_Vu_0 = \begin{pmatrix} v_+ \\ v_- \end{pmatrix}$$

satisfies

$$\partial_{x_d}v_{\pm} + T_iG_{\pm}v_{\pm} = f_{\pm}$$

with

$$\|f_{\pm}\|_{L^2} + \|v_{\pm}\|_{L^2} + \|v_{\pm}|_{x_d = 0}\|_{L^2} \lesssim \|u_0\|_{L^2} + \|f_0\|_{L^2} + \|u_0|_{x_d = 0}\|_{L^2}.$$  

Moreover, the blocks $G_{\pm}$ are strongly hyperbolic and admit smooth symmetrizers. Therefore, by Proposition 4.2 applied with $x_d$ as time variable, one can extend $f_{\pm}$ and $v_{\pm}$ to $\{x_d < 0\}$ so that (5.31) and (5.32) remain satisfied. Denoting by $\tilde{v}$ the extension of $v$, we see that $\tilde{u}_0 = T_{V^{-1}}\tilde{v}$ satisfies

$$\|L\tilde{u}_0\|_{L^2} + \|\tilde{u}_0\|_{L^2} + \|v_{\pm}|_{x_d = 0}\|_{L^2} \lesssim \|u_0\|_{L^2} + \|f_0\|_{L^2} + \|u_0|_{x_d = 0}\|_{L^2}.$$  

Moreover, on $\{x_d > 0\}$, $\tilde{u}_0|_{x_d > 0} - u_0 = (T_{V^{-1}}T_V - \text{Id})u_0$ and by the symbolic calculus

$$\int_0^\infty \|\tilde{u}_0|_{x_d > 0}(\cdot, x_d) - u_0(\cdot, x_d)\|_{H^1_{\dot{L}^{\infty}}}^2 dx_d \lesssim \|u_0\|_{L^2}^2$$

and therefore

$$\sup_t \|\tilde{u}_0|_{x_d > 0}(t, \cdot) - u_0(t, \cdot)\|_{L^2_{\dot{L}^{\infty}}}^2 \lesssim \|\tilde{u}_0\|_{L^2}^2.$$
Using the energy estimate for the Cauchy problem implies that for $t \in [0,T]$

\[
\|\tilde{u}_0(t)\|_{L^2} \leq \|L\tilde{u}_0\|_{L^2([-1,T] \times \mathbb{R}^d)} + \|\tilde{u}_0\|_{L^2([-1,T] \times \mathbb{R}^d)}
\]

and Proposition 5.14 follows from (5.33) and (5.34).

5.3.2 The case $|\tau| \leq C|\xi'|$

Let $c$ be chosen as in Proposition 5.17. We now consider the remaining frequencies $|\tau| \leq C|\xi'|$ with $C = 1/c$. The idea is to use Propositions 4.10 and 4.11 to estimate $\|u_k(t)\|_{L^2}$. Let $\delta_1 > 0$ be given by Proposition 4.11 and let $\alpha_1 = \delta_1/(2 + C)$, so that

\[
|\tau| \leq (1 + C)|\xi'| \quad \Rightarrow \quad \alpha_1|\zeta| \leq \delta_1|\xi'|.
\]

We fix a point $\tilde{x} = (t, x)$ and $\tilde{\zeta} = (\tau, \xi')$ with $|\tau| \leq C|\xi'|$. We assume as we may that $|\zeta| = 1$. The spectrum of $G(\tilde{x}, \tilde{\zeta})$ is made of at most $N$ isolated eigenvalues. Denote by $\mu_k$ the distinct values of their imaginary part. Then, there is $\alpha \in [0, \alpha_1]$ such that $\inf |\mu_k + \alpha| \geq \alpha_1/N$. Therefore, with $c = \frac{1}{2}\alpha_1/N$, there is a conical neighborhood of $(\tilde{x}, \tilde{\zeta})$ such that for $(\tilde{x}, \tilde{\zeta})$ in this neighborhood, the spectrum of $G(\tilde{x}, \tilde{\zeta})$ is contained in $|\text{Im} \lambda + \alpha| \geq c|\zeta|$.

We choose $\chi$ supported in this neighborhood and set $v = \chi(\tilde{x}, D_x)u$. Then

\[
\|\partial_x v + iT\tilde{G}v\|_{L^2} \lesssim \|u\|_{L^2} + \|f\|_{L^2},
\]

where $\tilde{G}$ is an extension of $G$ outside the given neighborhood such that $\tilde{G}$ satisfies the spectral property everywhere. Hence, the elliptic estimate of Proposition 4.13 implies that

\[
\|\langle D_x \rangle \frac{1}{2} e^{-\alpha \langle D_x \rangle x_d} v\|_{L^2} \lesssim \|u\|_{L^2} + \|f\|_{L^2} + \|u|_{x_d=0}\|_{L^2}.
\]

Using (5.36), this implies that

\[
\|\langle D_{x'} \rangle \frac{1}{2} e^{-\delta x_d(D_{x'})} v\|_{L^2} \lesssim \|u\|_{L^2} + \|f\|_{L^2} + \|u|_{x_d=0}\|_{L^2}.
\]

with $\delta = (2 + C)\alpha \in [0, \delta_1]$. Hence the estimate (5.29) for $v$ follows from the energy estimates of Propositions 4.10 and 4.11 and the proof of Theorem 5.8 is now complete.
6 Counterexamples

In this section we prove Theorem 1.4. We give an example of an ill posed initial boundary value problem for a $6 \times 6$ symmetric system with boundary conditions which satisfy the uniform Lopatinski condition. This example can be seen as well as a transmission problem for a symmetric $3 \times 3$ system.

The example is in dimension $d = 3$, the space variables are denoted by $(x, y, z)$ and the boundary is $\{ x = 0 \}$. The dual variables are $(\xi, \eta, \zeta)$. The eigenvalues have variable multiplicities on the manifold $\xi = \eta = y = 0, \zeta \neq 0$.

Consider in $\mathbb{R}^{1+3}$

\[(6.1) \quad L_\varepsilon = \begin{pmatrix} \partial_t - \varepsilon \partial_x & \partial_y & y \partial_z \\ \partial_y & \partial_t + \varepsilon \partial_x & 0 \\ y \partial_z & 0 & \partial_t + \varepsilon \partial_x \end{pmatrix} = \text{Id} \partial_t + \varepsilon J \partial_x + A \partial_y + y B \partial_z \]

With $\varepsilon_1 = 1$ and $\varepsilon_2 = -1$, consider on $\{x > 0\}$ the doubled system

\[(6.2) \quad L_{\varepsilon_1} U_1 = 0, \quad L_{\varepsilon_2} U_2 = 0 \]

together with boundary conditions on $\{x = 0\}$ of the form

\[(6.3) \quad BU := \begin{pmatrix} u_2 \\ v_1 \\ w_1 \\ u_2 \\ v_2 \\ w_2 \end{pmatrix} - M \begin{pmatrix} u_1 \\ v_1 \\ w_1 \end{pmatrix} = 0, \quad \text{where} \quad U_j = \begin{pmatrix} u_j \\ v_j \\ w_j \end{pmatrix}. \]

We choose $M$ of the form

\[(6.4) \quad M = \begin{pmatrix} c & 0 & 0 \\ 0 & \frac{1}{2} & \frac{i}{2} \\ 0 & -\frac{i}{2} & \frac{1}{2} \end{pmatrix} \]

The system is symmetric. The form (6.3) is well adapted to the diagonal 1-D system $L(\partial_t, \partial_x, 0, 0)$ and when $\| M \| < 1$, the system is maximal strictly dissipative. In particular, if $M$ is of the form (6.4), the boundary condition is dissipative if and only if $|c| \leq 1$. The uniform Lopatinski condition is satisfied on a wider range of $c$:

\[\text{Proposition 6.1. When } |c| < 2, \text{ the boundary conditions (6.3), (6.4) satisfy the uniform Lopatinski condition for the system (6.2).} \]

Therefore, Theorem 6 follows from the next result:
Theorem 6.2. When \( c \in ]-2,-1[ \), the initial boundary value problem is strongly ill posed in the sense that there are families \( U_\lambda \) of smooth solutions of \( LU = 0 \) on \( \{ t \geq 0, x \geq 0 \} \), \( BU = 0 \) on \( \{ t \geq 0, x = 0 \} \) such that

i) the \( U_\lambda(0, \cdot) \) are bounded in \( H^s(\mathbb{R}_+^3) \) for all \( s \),

ii) for all time \( t > 0 \), the \( U_\lambda(t, \cdot) \) are not bounded in \( L^2(\mathbb{R}_+^3) \).

Remark 6.3. Since the \( U_\lambda \) are smooth up to the boundary on the initial surface, the compatibility conditions are satisfied at infinite order on the edge \( \{ t = x = 0 \} \). We do not make them explicit nor comment more on this point here.

To prove the theorem, we first construct exact solutions of \( L_\varepsilon U_\varepsilon = 0 \).

Lemma 6.4. Let \( \mu > 0 \) and \( \gamma > 0 \) satisfy \( \mu^2 - \gamma^2 = 1 \) and for \( \varepsilon \in \{-1,+1\} \) let

\[
\delta = \varepsilon \mu - \gamma = \frac{1}{\gamma + \varepsilon \mu}.
\]

For all \( \zeta > 0 \), introduce

\[
\Phi(t,x,y,z,\zeta) = \sqrt{\zeta} (\gamma t - \mu x) + i \zeta z - \frac{1}{2} \zeta y^2
\]

Then

\[
U_\varepsilon(t,x,y,z,\zeta) = e^{\Phi} (\sqrt{\zeta} y e_1 + \delta e_2)
\]

with

\[
e_1 = \begin{pmatrix} 0 \\ -1 \\ i \end{pmatrix}, \quad e_2 = \begin{pmatrix} 1 \\ 0 \\ 0 \end{pmatrix}, \quad e_3 = \begin{pmatrix} 0 \\ 1 \\ i \end{pmatrix}.
\]

is an exact solution of \( L_\varepsilon U_\varepsilon = 0 \).

Proof. One has

\[
\zeta^{-\frac{1}{2}} e^{-\Phi} L_\varepsilon U_\varepsilon = \zeta y^2 F_2 + \zeta^{\frac{1}{2}} y F_1 + F_0
\]

with

\[
F_2 = (-A + iB)e_1 = 0,
\]

\[
F_1 = (-A + iB)\delta e_2 + (\gamma \text{Id} - \varepsilon \mu J)e_1 = (\delta + \gamma - \varepsilon \mu)e_1,
\]

\[
F_0 = Ae_1 + (\gamma \text{Id} - \varepsilon \mu J)\delta e_2 = ((\gamma + \varepsilon \mu)\delta - 1)e_2.
\]
where we have used that
\[ (-A + iB)e_1 = 0, \quad (-A + iB)e_2 = e_1, \quad Ae_1 = -e_2, \]
\[ Je_1 = e_1, \quad Je_2 = -e_2. \]

Therefore, the conditions on the parameters imply that \( L_\varepsilon U_\varepsilon = 0 \)

**Lemma 6.5.** Let \( \mu > 0 \) and \( \gamma > 0 \) satisfy \( \mu^2 - \gamma^2 = 1 \). Let \( U_1 \) and \( U_2 \) be defined by (6.1) with \( \delta_j = \varepsilon_j \mu - \gamma \), with \( \varepsilon_1 = 1 \) and \( \varepsilon_2 = -1 \) as above. Then \( L_\varepsilon U_j = 0 \) and the boundary condition (6.3) (6.4) is satisfied if \( c = \delta_2/\delta_1 \).

**Proof.** On the boundary
\[ U_1 = e^\Phi \begin{pmatrix} \delta_1 \\ -Y \\ iY \end{pmatrix}, \quad U_2 = e^\Phi \begin{pmatrix} \delta_2 \\ -Y \\ iY \end{pmatrix}, \]
with \( Y = y\sqrt{\zeta} \). Therefore,
\[ e^{-\Phi} B U = \begin{pmatrix} \delta_2 \\ -Y \\ iY \end{pmatrix} - \begin{pmatrix} c & 0 & 0 \\ 0 & \frac{1}{2} & \frac{i}{2} \\ 0 & -\frac{i}{2} & \frac{1}{2} \end{pmatrix} \begin{pmatrix} \delta_1 \\ -Y \\ iY \end{pmatrix} = 0 \]
when \( c = \delta_2/\delta_1 \).

**Corollary 6.6.** Let \( \mu > 0 \) and \( \gamma > 0 \) satisfy \( \mu^2 - \gamma^2 = 1 \) and let \( c = \delta_2/\delta_1 \). Then initial boundary value problem for (6.2) (6.3) (6.4) is strongly ill posed.

**Proof.** Consider for \( \lambda \) large the function
\[ U_\lambda(t, x, y, z) = e^{-\lambda \rho} \int U(t, x, y, z, \zeta) \varphi(\zeta - \lambda) d\zeta \]
where \( \varphi \in C^\infty_0([1, \infty[) \) and \( \rho < \frac{1}{2} \). They satisfy \( LU_\lambda = 0 \) and the boundary condition \( BU_\lambda = 0 \), for all time. In particular, the compatibility conditions at the edge \( \{ t = x = 0 \} \) are satisfied.

At \( t = 0 \) the phase \( \Phi \) is \( -\sqrt{\zeta} \mu x - \frac{1}{2} \zeta y^2 + i\zeta z \) and for all \( s \),
\[ \|U_\lambda(t, \cdot)\|_{H^s} = O(1) \]
and similar estimates are true for \( t < 0 \). On the other hand, for \( t > 0 \) the phase has the amplification factor \( \gamma \sqrt{\zeta} t \), and
\[ \|U_\lambda(t, \cdot)\|_{L^2} \gtrsim e^{\gamma \sqrt{\zeta} t/2}. \]
Proof of Theorem 6.2. For \( c \in [2, -1] \), one can choose \( \mu > 0 \) and \( \gamma > 0 \) such that \( \mu^2 = 1 + \gamma^2 \) and

\[
c = \frac{\delta_2}{\delta_1} = -\frac{\mu + \gamma}{\mu - \gamma}
\]

and the theorem follows.

Proof of Proposition 6.1. For the symbolic analysis, \( y \) is a parameter independent of \( \eta \) and we can replace \( y \zeta \) by \( \zeta \). Clearly, this is where the symbolic analysis diverges from the exact computations with differential operator.

a) For \( L_\varepsilon \), the equations for the eigenvectors are

\[
\begin{align*}
(-\varepsilon \xi + \tau)u + \eta v + \zeta w &= 0 \\
(\varepsilon \xi + \tau)v + \eta u &= 0 \\
(\varepsilon \xi + \tau)w + \zeta u &= 0
\end{align*}
\]

(6.9)

Introduce polar coordinates for \((\eta, \zeta)\):

\[
\eta = \rho \cos \theta, \quad \zeta = \rho \sin \theta.
\]

The eigenvalues are

\[
\xi = -\varepsilon \tau, \quad \xi = \pm \sqrt{\tau^2 - \rho^2}
\]

When \( \rho > 0 \), they are distinct and simple. An eigenvector for \( \xi = -\varepsilon \tau \) is

\[
R_0 = \begin{pmatrix} 0 \\ -\sin \theta \\ \cos \theta \end{pmatrix}.
\]

Eigenvectors for the other two are

\[
R = \begin{pmatrix} \alpha \\ \cos \theta \\ \sin \theta \end{pmatrix} \quad \text{or} \quad R' = \begin{pmatrix} 1 \\ \beta \cos \theta \\ b \sin \theta \end{pmatrix},
\]

with

\[
\alpha = \frac{-\rho}{-\varepsilon \xi + \tau}, \quad \beta = \frac{-\rho}{\varepsilon \xi + \tau}.
\]
Recall that in this analysis, $\text{Im} \tau < 0$ and the incoming space is the invariant space associated to eigenvalues in $\text{Im} \xi > 0$. When $\varepsilon = -1$, $\text{Im} (\varepsilon \xi + \tau) < 0$ and thus $\dim \mathbb{E}^\text{in}_\varepsilon$ has dimension one and is generated by $R'$:

$$
\mathbb{E}^\text{in}_\varepsilon = \mathbb{C} \begin{pmatrix}
1 \\
\beta \cos \theta \\
\beta \sin \theta
\end{pmatrix}
$$

When $\varepsilon = +1$, $\text{Im} (-\varepsilon \xi + \tau) < 0$, $\dim \mathbb{E}^\text{in}_\varepsilon$ has dimension two and is generated by $R_0$ and $R$:

$$
\mathbb{E}^\text{in}_\varepsilon = \mathbb{C} \begin{pmatrix}
0 \\
-\sin \theta \\
\cos \theta
\end{pmatrix} \oplus \mathbb{C} \begin{pmatrix}
\alpha \\
\cos \theta \\
\sin \theta
\end{pmatrix}.
$$

Combining the two cases, we conclude that for the symbol of the doubled system (6.2) $\dim \mathbb{E}^\text{in}$ has dimension three and is generated by

$$
E_0 = \begin{pmatrix} R_0 \\ 0 \end{pmatrix}, \quad E_1 = \begin{pmatrix} R_1 \\ 0 \end{pmatrix}, \quad E_2 = \begin{pmatrix} 0 \\ R_2 \end{pmatrix}
$$

with

$$
R_0 = \begin{pmatrix}
0 \\
-\sin \theta \\
\cos \theta
\end{pmatrix}, \quad R_1 = \begin{pmatrix}
a \\
\cos \theta \\
\sin \theta
\end{pmatrix}, \quad R_2 = \begin{pmatrix}
1 \\
a \cos \theta \\
a \sin \theta
\end{pmatrix}.
$$

with

$$
a = \frac{\rho}{\tau - \xi}, \quad \text{where} \quad \xi^2 = \tau^2 - \rho^2, \quad \text{Im} \xi > 0.
$$

This extends to the case $\rho = 0$. Recall the following elementary result,

**Lemma 6.7.** The image of $\{\text{Im} \tau < 0, \rho \in \mathbb{R}\}$ by the mapping (6.11) is $D := \{|a| < 1\}$.

b) Applying the boundary conditions $B$ to the basis of $\mathbb{E}^\text{in}$ yields

$$
BE_0 = \begin{pmatrix}
0 \\
-\sin \theta \\
\cos \theta
\end{pmatrix},
$$

$$
BE_1 = \begin{pmatrix}
0 \\
\cos \theta \\
\sin \theta
\end{pmatrix} - M \begin{pmatrix} a \\ 0 \end{pmatrix} = \begin{pmatrix}
-ca \\
\cos \theta \\
\sin \theta
\end{pmatrix}
$$

50
\[
BE_2 = \begin{pmatrix} 1 \\ 0 \\ 0 \end{pmatrix} - M \begin{pmatrix} \cos \theta \\ a \sin \theta \\ a \cos \theta \end{pmatrix} = \begin{pmatrix} 1 \\ -\frac{a}{2}e^{i\theta} \cos \theta \\ -\frac{a}{2}e^{i\theta} \sin \theta \end{pmatrix}.
\]

Because the three vectors in (6.10) are uniformly independent, the uniform Lopatinski condition is satisfied if and only if the modulus of the Lopatinski determinant \( \Delta = \det(BE_0, BE_1, BE_2) \) is bounded from below by a positive constant. One has

\[
\Delta = \det \begin{pmatrix} 0 & -ca & 1 \\ -\sin \theta & \cos \theta & -\frac{a}{2}e^{i\theta} \\ \cos \theta & \sin \theta & \frac{i}{2}ae^{i\theta} \end{pmatrix}
\]

and

\[
\Delta = -1 + \frac{1}{2}ca^2 e^{i\theta} (\cos \theta - i \sin \theta) = -1 + \frac{1}{2}ca^2.
\]

Since \(|a| < 1\), we see that \(|\Delta| \geq 1 - \frac{1}{2}|c|\) and the proposition is proved. \(\square\)

**References**


