# Some remarks on the Cauchy problem 

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## 1. Introduction

Consider the first order partial differential equation

$$
\begin{equation*}
M[u]=\frac{\partial}{\partial t} u-\sum_{k=1}^{l} A_{k}(x, t) \frac{\partial}{\partial x_{k}} u-B(x, t) u=0, \tag{1.1}
\end{equation*}
$$

where $u$ is a vector-valued function with $N$ components; $A_{k}$ and $B$ are matrices of order $N$, infinitely differentiable with respect to $t$ and $x=\left(x_{1}, x_{2}, \cdots, x_{l}\right)$. Consider the Cauchy problem for this equation with given initial value at $t=0$. We say that the Cauchy problem for (1.1) is well posed for the future in the space $\mathcal{E}$, if i) for any given initial value $u(x, 0) \in \mathcal{E}_{x}$, there exists a unique solution $u(x, t) \in \mathcal{E}_{x, t}, t \geqslant 0$, which takes the given initial value at $t=0$, ii) this linear mapping $u(x, 0) \rightarrow u(x, t)$ is continuous from $\mathscr{E}_{x}$ into $\mathscr{E}_{x, t}$. Here $\mathscr{E}$ is the space of all infinitely differentiable functions with customary topology, (see L. Schwartz [8]) ${ }^{1)}: f_{j} \rightarrow 0$ in $\mathcal{E}$,

1) In this paper, especially in section 1, we followed the notations of Schwartz's treatise [8]. We used the following linear topological spaces: $\mathcal{E}, \mathcal{E}^{m}(\Omega)$, $\mathcal{B}, \mathscr{G}_{L^{2}}^{\infty}, \mathcal{E}_{L^{2}}^{m}(\Omega), \mathcal{E}_{L^{2}(\mathrm{loc})}^{m}(\Omega), \Omega$ being an open set in $R^{n}$, and $\mathcal{E}_{-}\left(\mathscr{D}_{L^{2}}^{\infty}\right)$. We shall explain briefly. $\mathcal{E}^{m}(\Omega)$ is the space of all functions $f(x), m$-times continuously differentiable with the topology : $f_{j} \rightarrow 0$, if $D^{\nu} f_{j}(x),|\nu| \leqslant m$, converge to 0 uniformly on any compact in $\Omega$. $\mathcal{E}_{L_{2}}^{m}(\Omega)$ is the Hilbert space of all functions $f(x)$ belonging to $L^{2}(\Omega)$ with their derivatives (in distribution sense) up to order $m$, with the inner product: $(f, g)=\sum_{i v \mid \leqslant_{m}}\left(D^{\nu} f, D^{\nu} g\right)_{L^{2}} . \quad \mathcal{E}_{L_{2(\text { loc })}^{2 m}(\Omega)}(\Omega)$ the space of all functions $f(x)$ such that $D^{\nu} f(x) \in L^{2}{ }_{\text {loc }}(\Omega)$, for $|\nu| \leqslant m$, with the topology : $f_{j} \rightarrow 0$, if for any compact $K$ of $\Omega$, $D^{\nu} f_{j}(x) \rightarrow 0$ in $L^{2}(K)$, for $|\nu| \leqslant m$. Finally, $f(x, t) \in \mathcal{E}^{\mathcal{~}}\left(G_{L^{2}}^{\infty}\right)$, if the mapping $t \rightarrow f(x, t)$ $\in \mathscr{D}_{L^{2}}^{\infty}$ is infinitely differentiable for $t \leqslant 0$.
if for any compact $K$ and any integer $\nu, D^{\nu} f_{j}$ converge uniformly to 0 on $K$.

In the study of hyperbolic systems, it is always assumed that (H) $\sum A_{k}(x, t) \xi_{k}=A \cdot \xi$ has only real eigenvalues for any $(x, t ; \xi)$, where $\xi$ is a non-zero real vector.

Our purpose is to prove the following
Theorem 1.1. Assume that, for some $\xi^{0}(\neq 0)$ real, the matrix $\sum A_{k}(0,0) \xi_{k}^{0}$ has a non real eigenvalue, then (1.1) is not well posed in any small neighborhood of the origin.

From this it follows the following interesting
Corollary. Assume the coefficients $A_{k}(x, t)$ and $B(x, t)$ to be analytic functions of $x$ and $t$. If we make the same assumption as that of Theorem 1.1, then there exists a function $\psi_{0}(x) \in \mathcal{E}$, defined in a neighborhood of $x=0$, such that there exists no solution $u(x, t) \in \mathcal{E}$ of (1.1) satisfying $u(x, 0)=\psi_{0}(x)$, in any small neighborhood of the origin.

Remark 1. This corollary is also true even if we restrict there the domain of existence of $u(x, t)$ to a half-space, that is, for instance for $t \geqslant 0$.

Remark 2. Theorem 1.1 was proved at first by Petrowsky [6] in the case where the coefficients depend only on $t$. In the case of variable coefficients, P. D. Lax proved this under the condition that this non real eigenvalue is simple [3]. Our proof is a direct extension of that of Petrowsky and quite different from that of Lax ${ }^{2)}$. The above corollary was proved by Lax in [3], however we present here our proof.

We shall prove Theorem 1.1 in section 3. Here we prove the corollary, assuming Theorem 1.1 to be true. This proof was borrowed fairly from Hörmander's paper [2], p. 135.

[^0]We assume now :
i) The coefficients are analytic functions;
ii) For each $\psi(x) \in \mathcal{E}(\Omega), \Omega$ being an arbitrary open set in $x$ space containing the origin ( $x=0$ ), once fixed for all, there exists a solution $u(x, t) \in \mathcal{E}$ in a neighborhood of the origin, which may depend on $\psi$.

From Holmgren's Theorem, it follows this: Define a familly of open sets

$$
\begin{equation*}
D_{\rho}=\left\{(x, t) ;|t|+|x|^{2}<\rho\right\}, \quad 0<\rho<\rho_{0} \quad(\text { small }), \tag{1.2}
\end{equation*}
$$

then any solution $u(x, t) \in \mathscr{E}^{1}$ of $M[u]=0$ with a given initial value is unique in $D_{\rho}$. Now we prove

Lemma 1.1. Under the assumptions i) and ii), there exists a $\varepsilon>0$ such that, for any function $\psi(x) \in \mathcal{E}(\Omega)$, there exists a unique solution in $\mathscr{E}^{1}\left(D_{\varepsilon}\right)$, satisfying $u(x, 0)=\psi(x)$.

Proof. Take a sequence $\varepsilon_{0}=\rho_{0}>\varepsilon_{1}>\varepsilon_{2}>\cdots>\varepsilon_{n} \cdots \rightarrow 0 ;$ Denote by $A_{n, m}$ the set of all functions $\psi(x)$ of $\mathcal{E}(\Omega)$ such that 1) the solution $u(x, t)$ having $\psi(x)$ as initial value has at least $D_{\mathrm{s}_{n}}$ as its existence domain: more precisely, $M[u]=0, u(x, t) \in E_{L^{2}}^{\left[\frac{2}{2}\right]+2}\left(D_{\varepsilon_{n}}\right) ; u(x, 0)$ $=\psi(x)$ for $x \in D_{\mathrm{s} n} \cap(t=0)$,
2) $\|u(x, t)\|_{\left[\frac{2}{2}\right]+2} \leqslant m$, where $\left\|\|_{k}\right.$ means the customary norm of the functions in $\mathbb{E}_{L^{2}}^{k}\left(D_{\Sigma_{n}}\right):\|f\|_{k}^{2}=\sum_{\| V \leqslant k}\left\|D^{\nu} f\right\|_{L^{2}}^{2}$.

The set $A_{n, m}$ is symmetric convex and closed. The first two properites are evident. We need only to show that $A_{n, m}$ is closed. Suppose that $\psi_{n} \in A_{n, m} \rightarrow \psi_{0} \in \xi(\Omega)$. Corresponding to $\psi_{n}$, we have the sequence $u_{n}(x, t)$. Since $u_{n}$ is bounded in $\mathcal{E}_{L^{2}}^{\left[\frac{2}{2}\right]+2}\left(D_{\varepsilon_{n}}\right)$, we can choose a convergent subsequence $u_{n p}$ in $\mathscr{E}_{L_{2}(\text { (loc })}^{\left[\frac{2}{2}\right]+1}$ (namely, they converge on any compact of $D_{\mathrm{e} n}$ in $\mathscr{E}_{L^{2}}^{\left[\frac{2}{2}\right]+1}{ }^{-1}$ topology). Furthermore, if necessary, by taking again its subsequence, we can assume this subsequence to be weakly convergent (as a Hilbert space) in $\mathcal{E}_{L^{2}}^{\left[\frac{l}{2}\right]+1}\left(D_{\varepsilon_{n}}\right)$. We denote this limit by $u_{0} \in \mathcal{E}_{t^{2}}^{\left[\frac{l}{2}\right]+2}\left(D_{\varepsilon_{n}}\right)$. Then, $u_{n p} \rightarrow u_{0}$
weakly implies that $u_{0}$ satisfies 2 ), and by Sobolev, $u_{0} \in \mathscr{E}^{1}\left(D_{\varepsilon_{n}}\right)$, and $M\left[u_{0}\right]=0 . \quad u_{n p} \rightarrow u_{0}$ in $\mathscr{E}_{L_{2}\left(l_{0 c}\right)}^{\left[\frac{2}{2}\right]+1}\left(D_{\varepsilon n}\right)$ implies that, since $u_{n p}$ are uniformly convergent on every compact in $D_{\text {ह } n} \cap(t=0)$, we have $u_{0}(x, 0)=\psi_{0}(x)$, which proves that $A_{n, m}$ is closed. Since $\bigcup_{n, m} A_{n, m}$ $=\mathcal{E}(\Omega)$, one of $A_{n, m}$, say $A_{n_{0}, m_{0}}$, is of second category, therefore it contains an open set of $\mathcal{E}(\Omega)$. Since $A_{n_{0}, m_{0}}$ is symmetric and convex, it contains a neighborhood of 0 in $\mathcal{E}(\Omega)$. This implies that, for any $\psi(x) \in \mathcal{E}(\Omega)$, there exists a unique solution $u(x, t) \in$ $\mathcal{E}_{L^{2}}^{\left[\frac{2}{2}\right]+2}\left(D_{\mathrm{\varepsilon} n_{0}}\right)$, a fortiori $\in \mathcal{E}^{1}\left(D_{\mathrm{\varepsilon} n_{0}}\right)$ satisfying $u(x, 0)=\psi(x)$. The proof is thus complete.

Lemma 1.2. The mapping $\psi(x) \in \mathcal{E}(\Omega) \rightarrow u(x, t) \in \mathcal{E}^{1}\left(D_{\varepsilon}\right), D_{\varepsilon} \quad$ being defined in the previous lemma, is continuous.

Proof. This is the closed graph theorem of Banach.
Proof of the corollary. Assume that. for every $\psi(x) \in \mathcal{E}(\Omega)$, there exists a solution $u(x, t) \in \mathscr{E}$ in a neighborhood of the origin. Then, by Lemma 1.1 and 1,2 , there exists $D_{\varepsilon}(\varepsilon>0)$ such that $\psi(x) \in$ $\mathcal{E}(\Omega) \rightarrow u(x, t) \in \mathscr{E}^{1}\left(D_{\varepsilon}\right)$ is continuous. However Theorem 1.1 says that, under the assumtion stated there. this mapping cannot be continuous (as we shall see in the proof of this theorem, we need not to assume $u(x, t) \in \mathcal{E}$, it is enough to assume $\left.u(x, t) \in \mathcal{E}^{1}\right)$. This contradiction proves the corollary.

We remarked (Remark 1) that the corollary is also true even if we assume that the existence domain of $u$ in $D_{\rho}^{\prime}=\{(x, t) ; t \geqslant 0$, $\left.t+|x|^{2}<\rho\right\}$, where $\rho$ may depend on the initial value $\psi(x)$. We restrict ourselves to point out how to modify the proof of Lemma 1.1. At first we extend all functions $u(x, t)$ defined in $D_{\varepsilon_{n}}^{\prime}$ to $\tilde{D}_{\mathrm{\varepsilon} n}$ as follows (see, Schwartz, [7]): For $t<0$, define $u_{1}(x, t)$ $=\sum_{\nu=1}^{n} a_{\nu} u(x,-\nu t)$, where $p=\left[\frac{l}{2}\right]+2, a_{\nu}$ being determined uniquely by the condition that this function has the same traces on the hyperplane $t=0$ as those of $u(x, t)$ up to order $p-1=\left[\frac{l}{2}\right]+1$; namely, $\sum_{v=1}^{p}(-\nu)^{k} a_{\nu}=1, k=0,1, \cdots, p-1$. Now $\tilde{u}(x, t)=u(x, t)+$ $u_{1}(x, t)$ is defined in $\widetilde{D}_{\varepsilon_{n}}$, and we have $D^{\nu} \tilde{u}(x, t)=D^{\nu} u(x, t)+D^{\nu} u_{1}(x, t)$
for $|\nu| \leqslant p=\left[\frac{l}{2}\right]+2$, the derivatives being taken in distribution sense, for both traces on $t=0$ of $u$ and $u_{1}$ up to order $p-1$ coincide. Moreover $\|\tilde{u}\|_{p}$ is bounded. Then we choose a subsequence of $u_{n}$ as follows : $u_{n p} \rightarrow u_{0}$ weakly in $\mathscr{E}_{L_{2}}^{p}\left(D_{\varepsilon_{n}}^{\prime}-(t=0)\right)$. Next consider the subsequence $u_{n_{p}}$ corresponding to $\left\{n_{p}\right\}$, and choose again a subsequence in such a way that this new subsequence converge in $\mathfrak{E}_{I_{2}(1 / c)}^{p-1}\left(\widetilde{D}_{\mathrm{\varepsilon}_{n}}\right)$.

Remark 3. It is desired that the corollary is proved in the following form: there exists a function $\psi_{0}(x) \in \mathcal{E}$, such that there exists no solution $u(x, t) \in \mathcal{E}^{1}$. We don't know whether this is true of not. However, as the proof shows, we can say that there exists no solution $u(x, t) \in \mathcal{E}^{\left[\frac{l}{2}\right]+2}, l$ being the dimension of the space.

Theorem 1.1. is of course true for higher order single equations, not only for kowalewskians but also for $p$ evolution equations. We say here that the linear single equation

$$
\begin{gather*}
M[u]=\frac{\partial^{m}}{\partial t^{m}} u+a_{1}\left(x, t ; \frac{\partial}{\partial x}\right) \frac{\partial^{m-1}}{\partial t^{m-1}} u+a_{2}\left(x, t ; \frac{\partial}{\partial x}\right) \frac{\partial^{m-2}}{\partial t^{m-2}} u+  \tag{1.3}\\
\cdots+a_{m}\left(x, t ; \frac{\partial}{\partial x}\right) u=0
\end{gather*}
$$

is $p$-evolution ( $p:$ positive integer), if
( $p$ ) $a_{i}(x, t ; \xi)$ is a polynomial in $\xi$ of degree $\leqslant i p$. We donote this highest homogeneous part (principal part as $p$-evolution) of degree $i p$ by $h_{i}(x, t: \xi)$.

There are infinitely many possible choices of $p$ for a given equation. Namely, if (1.3) is $p_{0}$-evolution, then it is of course $p$-volution for any $p>p_{0}$. However, in this case, the corresponding principal part becomes $h_{i} \equiv 0(i=1,2, \cdots, m)$ which is a trivial choice of $p$. Taking into account of this remark, we have

Theorem 1.2. Consider the characteristic equation (as p-evolution) :

$$
P(\lambda)=\lambda^{m}+h_{1}(x, t ; i \xi) \lambda^{m-1}+h_{2}(x, t ; i \xi) \lambda^{m-2}+\cdots+h_{m}(x, t ; i \xi)=0 .
$$

We denote the roots by $\lambda_{i}(x, t ; \xi)(\xi$ real $\neq 0)$.
If for some $x_{0}, \xi_{0}$, one of roots, say $\lambda_{1}$ has a positive real part:

$$
\begin{equation*}
\text { Real part } \lambda_{1}\left(x_{0}, 0 ; \xi_{0}\right)>0, \tag{1.4}
\end{equation*}
$$

then the forward Cauchy problem with initial time $t=0$ is not well posed in $\mathscr{D}_{L^{2}}^{\infty}$ in any small neighborhood of $t=0$.

The theorem is true even if we replace there $\mathscr{D}_{L^{2}}^{\infty}$ by $\mathscr{B}$ or $\mathscr{E}^{1]}$. However, we should remark the following fact: In the case of the equations with constant coefficients, if (1.3) is not kowalewskian, then the Cauchy problem is not well posed in $\mathcal{E}$, see Gårding [1].

We can consider the corollary corresponding to that of Theorem 1.1. We can say the following fact: For the parabolic equation of the form

$$
\begin{gather*}
M[u]=\frac{\partial}{\partial t} u-\sum_{i, j=1}^{i} a_{i j}(x, t) \frac{\partial^{2}}{\partial x_{i} \partial x_{j}} u-\sum a_{i}(x, t) \frac{\partial}{\partial x_{i}} u-  \tag{1.5}\\
c(x, t) u=0 ;
\end{gather*}
$$

where $\sum a_{i j}(x, t) \xi_{i} \xi_{j} \geqslant c|\xi|^{2}, c$ is a positive constant; the coefficients are in $\mathcal{B}$, consider the backward Cauchy problem in $\mathscr{D}_{1,2}^{\infty}$. Then there exists a $\psi(x) \in \mathscr{D}_{L_{2}^{2}}^{\infty}$ such that there exists no solution $u(x, t) \in \mathscr{E}_{-}\left(\mathscr{D}_{L^{2}}^{\infty}\right)^{1)}$ with the initial value $\psi(x)$ (at $t=0$ ) in any small neighborhood of $t=0$. In fact, we know that the backward Cauchy problem for this equation is unique in this space, see [4].

We shall give a brief proof of Theorem 1.2 in the last section.
Now we give a rough sketch of the proof of Theorem 1.1. At first, we remark that, if necessary by replacing $\xi^{0}$ by $-\xi^{0}$, we can assume

$$
\begin{equation*}
\text { Imaginary part } \lambda_{1}\left(0,0 ; \xi^{0}\right)<0, \quad\left|\xi^{0}\right|=1 \tag{1.5}
\end{equation*}
$$

Then, consider the convolution operator $M_{0}: M_{0}=\frac{\partial}{\partial t}-i A_{0} \xi^{0} \Lambda$, where $A_{0} \xi^{0}=\sum A_{k}(0,0) \xi_{k}^{0}$. We shall prove in section 2 that, so far as we restrict the sequence of the initial values in a convenient way, then this operator gives a good approximation to $M$. More explicitely, we define $u_{n}(x, 0)=\exp \left(i n \xi^{0} x\right) \psi(x)$, where $\psi(x)$ is an analytic function whose Fourier transform has its support located
only in a small neighborhood of $\xi^{0}$. Now we assume the continuity in $\mathcal{E}$, then

$$
\begin{equation*}
\max _{x \in \Omega}\left|u_{n}(x, t)\right| \leqslant 0\left(n^{h}\right) \quad \text { for } \quad 0 \leqslant t \leqslant T^{\prime} \tag{3.7}
\end{equation*}
$$

$\Omega$ being a compact chosen closely to the origin. Finally using the inequality obtained in section 2 , we show that (3.7) is a contradiction.

## 2. Operator $M_{0}$ approximating to $M$.

Take a function $\beta(x) \in C^{\infty}$ and $\alpha(\xi) \in C^{\infty}$ of small supports, which take the value 1 in a neighborhood of $x=0$ and $\xi^{0}$ respectively. We shall define these sizes later. Define

$$
\begin{equation*}
\alpha_{n}(\xi)=\alpha(\xi / n), \tag{2.1}
\end{equation*}
$$

and, denoting the Fourier inverse image of this function by $\alpha_{n}(x)$, define

$$
\begin{equation*}
\alpha_{n} f=\alpha_{n}(x) * f(x) \tag{2.2}
\end{equation*}
$$

Our aim is to show that the convolution operator

$$
\begin{equation*}
M_{0}=\frac{\partial}{\partial t}-i A_{0} \xi^{0} \Lambda \tag{2.3}
\end{equation*}
$$

is, in a certain sense, a first approximation of $M$.
At first, apply $\beta(x)$ to (1.1),

$$
\begin{equation*}
M[\beta(x) u]=\sum A_{k} \beta_{k}(x) u \tag{2.4}
\end{equation*}
$$

where $\beta_{k}(x)=\frac{\partial}{\partial x_{k}} \beta(x)$.
Now, since $M$ acts on $\beta(x) u$, it does not change if we modify the coefficients $A_{k}$ and $B$ outside the support of $\beta(x)$. Therefore we can write (2.4) in the form

$$
\begin{equation*}
\left(\frac{\partial}{\partial t}-\sum \tilde{A}_{k} \frac{\partial}{\partial x_{k}}-\tilde{B}\right)[\beta u]=\sum \widetilde{A}_{k} \beta_{k}(x) u \tag{2.5}
\end{equation*}
$$

where the oscillations of $\widehat{A}_{k}$ and $\widetilde{B}$ become small, when the size of Supp ( $\beta$ ) become small.

Next. we define $\exp \left(-\delta^{\prime} \Lambda t\right),\left(\delta^{\prime}>0\right)$, by

$$
\begin{equation*}
\left(\exp \left(-\delta^{\prime} \Lambda t\right) u\right)=\exp \left(-\delta^{\prime}|\xi| t\right) \hat{u}(\xi) \tag{2.6}
\end{equation*}
$$

This is the elementary solution of the evolution equation, for $t \geqslant 0$,

$$
\left(\frac{\partial}{\partial t}+\delta^{\prime} \Lambda\right) u=0 .
$$

Denote

$$
\begin{equation*}
\gamma_{n}(t)=\exp \left(-\delta^{\prime} \Lambda t\right) \alpha_{n} \tag{2.7}
\end{equation*}
$$

Apply this convolution operator to (2.5), then

$$
\begin{align*}
&\left(\frac{\partial}{\partial t}-\sum \widetilde{A}_{k} \frac{\partial}{\partial x_{k}}-\delta^{\prime} \Lambda-\tilde{B}\right)\left[\gamma_{n} \beta(x) u\right]  \tag{2.8}\\
&= \gamma_{n} \sum \widetilde{A}_{k} \beta_{k} u-\sum\left(\gamma_{n} \widetilde{A}_{k}-\widetilde{A}_{k} \gamma_{n}\right) \frac{\partial}{\partial x_{k}}[\beta(x) u] \\
& \quad-\left(\gamma_{n} \widetilde{B}-\widetilde{B} \gamma_{n}\right)(\beta(x) u) .
\end{align*}
$$

Now we consider the left-hand side. We denote this operator by $\tilde{M}-\delta^{\prime} \Lambda I$. Consider the operator $\sum \hat{A}_{k} \frac{\partial}{\partial x_{k}}=\mathscr{H} \Lambda$. Since $\mathscr{H} \Lambda$ acts on the functions of the form $\alpha_{n} v$, we can replace there $\mathscr{H}$ by $\tilde{\mathcal{H}}$. $\tilde{\mathcal{H}}$ is defined as follows: We call the support of $\alpha(\xi)$ on the unit sphere the projection from the origin. Then we change the symbol $i \sum A_{k}(x, t) \xi_{k} /|\xi|$ outside the support of $\alpha(\xi)$ on the sphere. Thus we obtain a singular integral operator $\tilde{\mathcal{H}}$. Here we can assume that the oscillation of $\sigma(\mathscr{H})$ becomes smaller and smaller if we choose the sizes of the supports of $\alpha$ and $\beta$ smaller and smaller.

Now we want to examine this situation more precisely. Consider the eigenvalues $\lambda_{1}^{0}, \cdots, \lambda_{N}^{0}$ of $i A_{0} \xi^{0}=i \sum A_{k}(0,0) \xi_{k}^{0}$. If eventually some of them are pure imaginary, we translate them by $-\varepsilon(\varepsilon>0$, small $)$, then we can assume that

$$
\begin{align*}
& \text { Real part } \quad \lambda_{1}^{0}-\varepsilon, \cdots, \lambda_{N_{1}}^{0}-\varepsilon \geqslant 6 \delta, \quad \delta>0, \quad N_{1} \geqslant 1  \tag{2.9}\\
& \text { Real part } \lambda_{N_{1}+1}^{0}-\varepsilon, \cdots, \lambda_{N}^{0}-\varepsilon \leqslant-6 \delta
\end{align*}
$$

Then, by a known theorem (see for instance, [6] p. 24, Lemma 5), we can find a non-singular matrix $N_{0}$ such that

$$
N_{0} \cdot i A_{0} \xi^{0} \cdot N_{0}^{-1}=\left|\begin{array}{cccc}
\lambda_{1}^{0} & & & \\
& \lambda_{1}^{0} & & 0 \\
& a_{i j}^{0} & \ddots & \\
& & & \lambda_{N}^{0}
\end{array}\right|=\mathscr{D}_{0},
$$

where $\left|a_{i,}^{0}\right|<\frac{\delta}{4 N}, \delta$ being defined in (2.9).
Denote

$$
\begin{equation*}
\sigma(\mathscr{H})=i A_{0} \xi^{0}+\sigma\left(\mathcal{H}_{\varepsilon}\right) . \quad \text { Then } \tag{2.11}
\end{equation*}
$$

$$
\begin{equation*}
N_{0} \mathscr{H}=\left(\mathscr{D}_{0}+N_{0} \mathcal{H}_{\mathrm{e}} N_{0}^{-1}\right) N_{0}=\left(\mathscr{D}_{0}+\mathscr{D}_{\varepsilon}\right) N_{0} . \tag{2.12}
\end{equation*}
$$

Now we define the sizes of the supports of $\beta$ and $\alpha$ and $T$ : If we take these sizes and $T$ small, we shall have, denoting the $(i, j)$-element of $\mathscr{D}_{\mathrm{z}}$ by $d_{i j}^{(\varepsilon)}$,

$$
\begin{equation*}
\left|\sigma\left(d_{i j}^{(\varepsilon)}\right)(x, t ; \xi)\right| \leqslant \frac{\delta}{4 N}, \quad \text { for } \quad 0 \leqslant t \leqslant T . \tag{2.13}
\end{equation*}
$$

Now we want to derive an inequality. At first we state a lemma on singular integral operators.

Lemma 2.1. Let $H$ be a singular integral operator (in the sense of Calderón and Zygmund).
i) If $\sigma=\inf _{x, \xi}$. Real part $\sigma(H)(x ; \xi)>0$, then

$$
\text { Real part }(H \Lambda u, u) \geqslant \frac{\sigma}{2}(\Lambda u, u)-C_{1}\|u\|^{2},
$$

where $C_{1}$ is a constant.
ii) Let $M=\sup _{x, \xi}|\sigma(H)(x ; \xi)|$, then

$$
|(H \Lambda u, u)| \leqslant 2 M(\Lambda u, u)+C_{2}\|u\|^{2},
$$

$C_{2}$ being a constant.
We don't give the proof. The reader will easily verify by consulting our previous paper [5].

Now we assume in (2.8), $\delta^{\prime}=\varepsilon+\frac{\delta}{2^{p}}, p=0$, positive integer, then, it can be written

$$
\begin{equation*}
\left\{\frac{\partial}{\partial t}-\left(\mathscr{D}_{0}+\mathscr{D}_{\varepsilon}-\delta^{\prime} I\right) \Lambda-B_{0}{ }_{j}^{\prime}\left[v^{(n)}\right]=N_{0} f,\right. \tag{2.14}
\end{equation*}
$$

where $f$ is the right-hand side of (2.8); $B_{0}(x, t)=N_{0} B(x, t) N_{0}^{-1}$, and

$$
\begin{equation*}
v^{(n)}=\gamma_{n} \beta(x) N_{0} u \tag{2.15}
\end{equation*}
$$

Hereafter we omit the suffix ( $n$ ) of $v^{(n)}$.
Consider now
(2.16) $S(t)=\sum_{i=1}^{N_{1}}\left\|v_{i}(t)\right\|^{2}-\sum_{j=N_{1}+1}^{N}\left\|v_{j}(t)\right\|^{2}$, where $v_{i}(t)=v_{i}(x, t)$ is the $i$-th component of $v$.
$S^{\prime}(t)=\sum_{i=1}^{N_{1}} 2 \operatorname{Re}\left(v_{i},\left(v_{i}\right)_{t}\right)-\sum_{j=N_{1}+1}^{N} 2 \operatorname{Re}\left(v_{j},\left(v_{j}\right)_{t}\right)$
$=\sum_{i} 2 \operatorname{Re}\left(v_{i},\left(\lambda_{i}^{0}-\delta^{\prime}+d_{i j}^{(\ell)}\right) \Lambda v_{i}\right)-\sum_{j} 2 \operatorname{Re}\left(v_{j},\left(\lambda_{j}^{0}-\delta^{\prime}+d_{i j}^{(\varepsilon)}\right) \Lambda v_{j}\right)$
$+\sum_{i} \sum_{k j} 2 \operatorname{Re}\left(v_{i},\left(a_{c k}^{0}+d_{i k}^{(\ell)}\right) \Lambda v_{k}\right)-\sum_{j} \sum_{k} 2 \operatorname{Re}\left(v_{j},\left(a_{j k}^{0}+d_{j k}^{(\ell)}\right) \Lambda v_{k}\right)$
$+\sum_{i} 2 \operatorname{Re}\left(v_{i}, f_{i}^{\prime}\right)-\sum_{j} 2 \operatorname{Re} e\left(v_{j}, f_{j}^{\prime}\right)-\sum_{i} 2 \operatorname{Re}\left(v_{i}, \sum_{k} b_{i k} v_{k}\right)-\sum_{j} 2 \operatorname{Re} e\left(v_{j}, \sum_{k} b_{j k} v_{k}\right)$,
where $f_{i}^{\prime}$ is the $i$ th component of $N_{\mathrm{o}} f$, and $B_{0}=\left(b_{i j}\right)$.
By the previous lemma, the first row of the right-hand side is greater than

$$
2 \delta\left(\sum_{i}\left(\Lambda v_{i}, v_{i}\right)+\sum_{j}\left(\Lambda v_{j}, v_{j}\right)\right)-C\|v\|^{23}
$$

In the same way, the second row is greater than

$$
-N \frac{2 \delta}{2 N}\left(\sum_{i}\left(\Lambda v_{i}, v_{i}\right)+\sum_{j}\left(\Lambda v_{j}, v_{j}\right)\right)-C\|v\|^{2}
$$

The last row is greater than

$$
\begin{aligned}
-C\|v\|^{2} & -\|v\|\left\|N_{0} f\right\| . \text { Hence } \\
S^{\prime}(t) & \geqslant \delta\left\{\sum_{i=1}^{N}\left(\Lambda v_{i}, v_{i}\right)\right\}-C\|v\|^{2}-2\|v\|\left\|N_{0} f\right\| \\
& \geqslant \delta\{\operatorname{distance}(0, \operatorname{Supp}(\hat{v}))-C\}\|v\|^{2}-2\|v\|\left\|N_{0} f\right\| .
\end{aligned}
$$

[^1]Assuming distance $(0, \operatorname{Supp}(\hat{v}))>4 C$, we have for $v^{(n)}$,

$$
\begin{equation*}
\frac{d}{d t} S\left(t ; v^{(n)}\right) \geqslant \frac{\delta}{2} n\left\|v^{(n)}\right\|^{2}-C\left\|v^{(n)}\right\|\|f\| . \tag{2.17}
\end{equation*}
$$

In fact, distance $\left(0, \operatorname{Supp}\left(\hat{v}^{(n)}\right)\right) \geqslant(3 / 4) n$,

## 3. Proof of Theorem 1.1.

We prove this theorem by contradiction. We assume therefore (1.1) to be well posed in $\mathscr{E}$.

At first we define a series of solutions $u_{n}(x, t)$ of (1.1). Namely we define their initial values. Let $\hat{\psi}(\xi)$ be a function whose support is located in a small neighborhood of $\xi^{0}$; On the support of $\hat{\psi}(\xi), \alpha(\xi)=1$; we assume $\int|\hat{\psi}(\xi)|^{2} d \xi=1$. We define then

$$
\begin{gather*}
\hat{\psi}_{n}(\xi)=\hat{\psi}\left(\xi-n \xi^{0}\right) \cdot \quad \text { Namely }  \tag{3.1}\\
\psi_{n}(x)=\exp \left(i n \xi^{0} x\right) \psi(x) \tag{3.2}
\end{gather*}
$$

Now we define $u_{n}(x, t)$ by

$$
M\left[u_{n}\right]=0, \quad u_{n}(x, 0)=N_{0}^{-1}\left(\begin{array}{c}
\psi_{n}(x)  \tag{3.2}\\
0 \\
\vdots \\
0
\end{array}\right), \quad n=1,2, \cdots
$$

To consider $\alpha_{n} u_{n}(x, t)$ means that we take only a high frequency part of $u_{n}(x, t)$ ( $t$ being parameter). Now we want to show that (3.3) $\left\|\alpha_{n} \beta(x) N_{0} u_{n}(x, 0)\right\|=c+0\left(\frac{1}{n}\right), c$ is a positive constant. $\alpha_{n} \beta(x) u_{n}(x, 0)=\alpha_{n} \beta(x) \psi_{n}(x)=\beta(x) \alpha_{n} \psi_{n}(x)+\left(\alpha_{n} \beta(x)-\beta(x) \alpha_{n}\right) \psi_{n}(x)$. Here $\alpha_{n} \psi_{n}(x)=\psi_{n}(x)$, for $\alpha_{n}(\xi)=1$ on the support of $\hat{\psi}_{n}(\xi)$. Hence the first term of the right-hand side is $\beta(x) \psi_{n}(x)=\exp \left(\right.$ in $\left.^{0} x\right) \times$ $\beta(x) \psi(x)$. Since $\psi(x)$ is analytic we have

$$
\begin{equation*}
\left\|\beta(x) \alpha_{n} \psi_{n}(x)\right\|=\left(\int|\beta(x) \psi(x)|^{2} d x\right)^{\frac{1}{2}}=c>0 . \tag{3.4}
\end{equation*}
$$

Now we look at the last term. We know that

$$
\left(\alpha_{n} \beta(x)-\beta(x) \alpha_{n}\right) u=\int \alpha_{n}(x-y)\{\beta(y)-\beta(x)\} u(y) d y
$$

By Taylor expansion : $\beta(y)-\beta(x)=\sum_{1 \leqslant|v| \leqslant p} \frac{(y-x)^{\nu}}{\nu!} \beta^{(\nu)}(x)+\sum_{|\nu|=p+1} \frac{(y-x)^{\nu}}{\nu!} \times$ $\beta_{\nu}(x, y)$, we see that the last integral is equal to

$$
\begin{gather*}
\sum_{1 \leqslant|v| \leqslant p} \frac{(-1)^{|\nu|}}{\nu!} \beta^{(\nu)}(x)\left(x^{\nu} \alpha_{n}\right) * u+\sum_{|v|=p+1} \frac{(-1)^{|\nu|}}{\nu!} \times  \tag{3.6}\\
\int \beta_{\nu}(x, y)(x-y)^{\nu} \alpha_{n}(x-y) u(y) d y
\end{gather*}
$$

When we put $u=\psi_{n}(x)$, the first terms corresponding to $\sum_{1 \leqslant i v \mid \leqslant p}$ are all zero, since $\left(x^{\nu} \alpha_{n}\right) * \psi_{n}(x) \underset{\nrightarrow}{ }$ const. $\alpha_{n}^{(\nu)}(\xi) \hat{\psi}_{n}(\xi)=0$.

Next, the last terms are majorized by $c\|u\| \int\left|\left(x^{\nu} \alpha_{n}\right)(x)\right| d x$. Now $\left|\left(x^{\nu} \alpha_{n}\right)(x)\right|$ is majorized by const. $\int\left|\alpha_{n}^{(\nu)}(\xi)\right| d \xi=$ const. $\left(\frac{1}{n}\right)^{|\nu|} \int_{\operatorname{Supp}\left(\alpha_{n}\right)}\left|\alpha^{(\nu)}\left(\frac{\xi}{n}\right)\right| d \xi \leqslant$ const. $\left(\frac{1}{n}\right)^{|\nu|-l}, l$ being the dimension of the space. On the other hand, for $|x| \geqslant 2$

$$
\begin{gathered}
\left(x^{\nu} \alpha_{n}\right)(x)=\frac{1}{|x|^{2 k}}\left(x^{\nu}|x|^{2 k} \alpha_{n}\right)(x) \leqslant \frac{\text { const. }}{|x|^{2 k}} \times \\
\int\left|\Delta^{k} \alpha_{n}^{(\nu)}(\xi)\right| d \xi \leqslant \frac{\text { const. }}{|x|^{2 k}}\left(\frac{1}{n}\right)^{|\nu|+2 k-l}
\end{gathered}
$$

This shows that, if we take $|x| \geqslant 1$, then

$$
\begin{equation*}
\int\left|\left(x^{\nu} \alpha_{n}(x)\right)(x)\right| d x=0\left(\frac{1}{n^{|\nu|-i}}\right) . \tag{3.6}
\end{equation*}
$$

Thus we proved (3.3).
Now we put $u=u_{n}(x, t)$ in (2.8) and $\delta^{\prime}=\varepsilon+\delta$. By hypothesis, there exist a positive integer $h$ and a neighborhood (in $x, t$ space) $\Omega$ of $(x, t)=0$ and small $T^{\prime}$ such that

$$
\begin{equation*}
\max _{(x, t) \in \Omega}\left|u_{n}(x, t)\right| \leqslant 0\left(n^{h}\right) \quad \text { for } \quad 0 \leqslant t \leqslant T^{\prime} \tag{3.7}
\end{equation*}
$$

By taking the support of $\beta(x)$ small, we can assume that the support of $\beta(x)$ is contained in $\Omega$, therefore we have

$$
\begin{equation*}
\left\|\beta(x) u_{n}(x, t)\right\| \leqslant 0\left(n^{h}\right), \quad \text { for } \quad 0 \leqslant t \leqslant T^{\prime} \tag{3.8}
\end{equation*}
$$

Now we consider the right-hand side of (2.8), we want to show that it is expressed, modulo bounded functions (in $L^{2}$-sense) with respect to $n$, as linear combination of

$$
\sqrt{n} \exp \left\{-\left(\varepsilon+\frac{\delta}{2}\right) \Lambda \dot{t}\right\} \sqrt{n} \bar{\gamma}^{\mid \gamma_{\mid}} \alpha_{n}^{(\gamma)} \times\left\{\begin{array}{l}
\beta(x) u_{n}  \tag{3.9}\\
\left(\frac{1}{\sqrt{n}}\right) \beta_{k}(x) u_{n}
\end{array}\right.
$$

whose coefficients being bounded operators in $L^{2}$ (bounded with respect to $n$ too). Namely

$$
\begin{align*}
& \sqrt{n} \sum C_{\gamma, \kappa}^{(\mathfrak{r})}(t) \exp \left\{-\left(\varepsilon+\frac{\delta}{2}\right) \Lambda t\right\} \times  \tag{3.10}\\
& \quad \sqrt{n^{|\gamma|} \alpha_{n}^{(\gamma)}}\left(\frac{1}{\sqrt{n}}\right)^{(\kappa)} \beta^{(\kappa)}(x) u_{n}(x, t)+0(1)
\end{align*}
$$

where $\left\|C_{\gamma_{,}, ~}^{(n)}(t)\right\|_{\mathcal{L}\left(L^{2} ; L^{2}\right)} \leqslant$ const. independent of $n ; 1 \leqslant|\gamma|+|\kappa| \leqslant$ $m=2(h+l),|\kappa| \leqslant 1$.

To see this, take the most delicate part of (2.8), $\left(\gamma_{n} \widetilde{A}_{k}-\widehat{A}_{k} \gamma_{n}\right) \times$ $\frac{\partial}{\partial x_{k}}\left[\beta(x) u_{n}\right]$. Denoting $\widetilde{A}_{k}$ merely by $A$, and using, as previously, the Taylor expansion, this can be expressed as

$$
\begin{align*}
& \sum_{|v|=1}^{m} \frac{(-1)^{|\nu|} \mid \nu}{\nu!} D^{\nu} A  \tag{3.11}\\
&\left(x^{\nu} \gamma_{n}\right) \frac{\partial}{\partial x_{k}}\left[\beta(x) u_{n}\right] \\
& \quad+\sum_{|\nu|=m+1} \int A_{\nu}(x, y)\left(x^{\nu} \gamma_{n}\right)(x-y) \frac{\partial}{\partial y_{k}}\left[\beta(y) u_{n}(y, t)\right] d y
\end{align*}
$$

where

$$
\begin{equation*}
m=2(h+l) \tag{3.12}
\end{equation*}
$$

where we chose $m$ so as to make the last terms of (3.11) bounded (with respect to $n$ ). Take at first a term of the first part: $\left(x^{\nu} \gamma_{n}\right)$ $\frac{\partial}{\partial x_{k}}\left[\beta(x) u_{n}\right]$. The convolution operator $\left(x^{\nu} \gamma_{n}\right) \frac{\partial}{\partial x_{k}}$ has its Fourier image const. $D_{\xi}^{\nu} \gamma_{n}(\xi) \xi_{k}=$ const. $D_{\xi}^{้}\left[\exp \{-(\varepsilon+\delta)|\xi| t\} \alpha_{n}(\xi)\right] \xi_{k} . \quad$ By Leibniz,

$$
\begin{align*}
& D_{\xi,}^{\nu}\left[\exp \{-(\varepsilon+\delta)|\xi| t\} \alpha_{n}(\xi)\right]  \tag{3.13}\\
& \quad=\sum_{\nu_{1}+\nu_{2}=\nu} C_{\nu_{1}}^{\nu} D_{\xi}^{\nu}[\exp \{-(\varepsilon+\delta)|\xi| t\}] D^{\nu_{2}} \alpha_{n}(\xi)
\end{align*}
$$

In general, we have

$$
D_{\xi}^{v} \exp \left\{-\delta^{\prime}|\xi| t\right\}=\psi_{\nu}(\xi, t) \exp \left\{-\delta^{\prime}|\xi| t\right\}, \quad\left(\delta^{\prime}=\varepsilon+\delta\right),
$$

where $\quad\left|\psi_{\nu}(\xi, t)\right| \leqslant C_{\nu}\left\{\sum_{k=1}^{|v|}(t|\xi|)^{k}\right\} /|\xi|^{|\nu|}, \quad$ for $\quad|\xi| \geqslant 1$.
Since $\left(\sum_{k=1}^{|| |}(t|\xi|)^{k}\right) \exp \left(-\frac{\delta}{2}|\xi| t\right)$ is bounded, we have finally

$$
\begin{equation*}
D_{\xi}^{\nu} \exp \{-(\varepsilon+\delta)|\xi| t\}=\mathcal{P}_{\vee}(\xi, t) \exp \left\{-\left(\varepsilon+\frac{\delta}{2}\right)|\xi| t\right\} \tag{3.14}
\end{equation*}
$$

where

$$
\left|\mathcal{P}_{\nu}(\xi, t)\right| \leqslant \frac{c_{\nu}}{|\xi|^{|v|}}, \quad \text { for } \quad|\xi| \geqslant 1 .
$$

Hence, there is no question as regards to the terms $\left|\nu_{1}\right| \geqslant 1$ in the right-hand side of (3.13). As regards to the terms $\nu_{1}=0$, (hence $\left|\nu_{2}\right| \geqslant 1$ ), we have $\alpha_{n}^{\left(\nu_{2}\right)}(\xi) \xi_{k}=n\left(\frac{\xi_{k}}{n}\right) \alpha_{n}^{\left(v_{2}\right)}(\xi)$. Since $\xi_{k} / n$ is uniformly bounded (with respect to $n$ ) on the support of $\alpha_{n}(\xi)$, we see that (3.10) is true for any of the first terms in (3.11). Therefore we need only to show that the last terms of (3.11) is $0(1)$. These terms are majorized by the form (by Young's Theorem)

$$
\begin{equation*}
c\left(\int\left|\frac{\partial}{\partial x_{k}}\left(x^{\nu} \gamma_{n}\right)(x)\right| d x\right)\left\|\beta(x) u_{n}\right\| . \tag{3.16}
\end{equation*}
$$

Now we want to estimate the above integral. Take Fourier transform of the integrand: $c^{\prime} \xi_{k} \gamma_{n}^{(\nu)}(\xi)$. This is majorized by $c^{\prime}|\xi|\left|\gamma_{n}^{(\nu)}(\xi)\right|$. Taking into account of $\left|\alpha_{n}^{(\nu)}(\xi)\right| \leqslant c\left(\frac{1}{n}\right)^{|\nu|}$, and of (3.14), we have

$$
\begin{equation*}
|\xi|\left|\gamma_{n}^{(\nu)}(\xi)\right| \leqslant c\left(\frac{1}{n}\right)^{|\nu|-1} . \tag{3.17}
\end{equation*}
$$

This implies


$$
=c\left(\frac{1}{n}\right)^{m-l}=c\left(\frac{1}{n}\right)^{n}\left(\frac{1}{\sqrt{n}}\right)^{m} .
$$

On ther other hand, for $|x| \geqslant 2$, putting

$$
\frac{\partial}{\partial x_{k}}\left(x^{\nu} \gamma_{n}\right)(x)=\frac{1}{|x|^{2 k^{\prime}}}|x|^{2 k^{\prime}} \frac{\partial}{\partial x_{k}}\left(x^{\nu} \gamma_{n}\right)(x) \text {, we see that this is }
$$ majorized by

$c \int\left|\Delta^{k^{\prime}}\left[\xi_{k} \gamma_{n}^{(\nu)}(\xi)\right]\right| d \xi \leqslant c\left(\frac{1}{n}\right)^{2 k^{\prime}+|\nu|-1} \int_{\operatorname{Supp}\left(\alpha_{n}\right)} d \xi \leqslant$ const. $\left(\frac{1}{n}\right)^{2 k^{\prime}+|v|-1-2 l}$, choosing $k^{\prime}=l$, we have

$$
\begin{equation*}
\left|\frac{\partial}{\partial x_{k}}\left(x^{\nu} \gamma_{n}\right)(x)\right| \leqslant c \frac{1}{|x|^{2 l}}\left(\frac{1}{n}\right)^{m}, \quad \text { for } \quad|x| \geqslant 2, \tag{3.19}
\end{equation*}
$$

where $c$ does not depend on $n$.
Finally we have

$$
\begin{equation*}
(3.16) \leqslant\left(\frac{1}{\sqrt{n}}\right)^{m} 0\left(\frac{1}{n^{n}}\right)\left\|\beta(x) u_{n}(x, t)\right\|=0(1)\left(\frac{1}{\sqrt{n}}\right)^{m} . \tag{3.20}
\end{equation*}
$$

We can see that the same reasoning on the rest terms of (2.8) will give (3.10).

In view of (3.10), we consider next all the functions appearing there :

$$
\exp \left\{-\left(\varepsilon+\frac{\delta}{2}\right) \Lambda t\right\} \sqrt{n}^{|\gamma|} \alpha_{n}^{(\gamma)} \sqrt{n}^{-|\kappa|} \beta^{(\kappa)}(x) u_{n}(x, t)
$$

where $\gamma$ and $\kappa$ satisfy the condition mentioned there. Namely, we make the same process for these functions as for $\gamma_{n} \beta(x) u$ : we replace in the above reasoning $\beta(x) \rightarrow\left\{\begin{array}{l}\beta(x) \\ \beta_{k}(x)\end{array}, \exp \{-(\varepsilon+\delta) \Lambda t\} \alpha_{n} \rightarrow\right.$ $\exp \left\{-\left(\varepsilon+\frac{\delta}{2}\right) \Lambda t\right\} \alpha_{n}^{(\gamma)}$. In total, we are led to consider all functions of the form:

$$
\begin{equation*}
\theta^{(n)}(p, \gamma, \kappa) u_{n}=\sqrt{n^{|\gamma|-|\kappa|}} \exp \left\{-\left(\varepsilon+\frac{\delta}{2^{p}}\right) \Lambda t\right\} \alpha_{2}^{(\gamma)} \beta^{\kappa \kappa}(x) u_{n} \tag{3.21}
\end{equation*}
$$

where $p \leqslant|\gamma|+|\kappa| \leqslant m, p=0,1, \cdots, m$.
We shall then have the equations analogous to (2.8) for all these functions. Namely, we have the equations of the form (2.14), where i) $v^{(n)}=\theta^{(n)}(p, \gamma, \kappa) N_{0} u_{n}$,
ii) the right-hand side $N_{0} f$ is expressed, modulo bounded functions, as a linear combination of the functions (3.21), whose
coefficients are $\sqrt{n} \times$ (uniformly bounded operators in $L^{2}$ ).
Now we return to section 2. For each $\theta^{(n)}(p, \gamma, \kappa) N_{0} u_{n}(x, t)$, consider $S(t)$, denoted by $S\left(t ; \theta^{(n)}(p, \gamma, \kappa) N_{0} u_{n}\right)$. Finally define

$$
\begin{align*}
& S_{n}(t)=\sum S\left(t ; \theta^{(n)}(p, \gamma, \kappa) N_{0} u_{n}(x, t)\right) ;  \tag{3.32}\\
& \sigma_{n}(t)=\sum\left\|\theta^{(n)}(p, \gamma, \kappa) N_{0} u_{n}(x, t)\right\|^{2},
\end{align*}
$$

where the summation is extended over all functions in (3.21). Taking into account of the fact that the support of $\left(\theta^{(n)}(p, \gamma, \kappa)\right.$ $\left.N_{0} u_{n}\right)^{\wedge}(\xi, t)$ is contained in that of $\alpha_{n}(\xi)$, if we apply the inequality (2.17), we shall have

$$
\begin{equation*}
S_{n}^{\prime}(t) \geqslant \frac{1}{2} \delta n \sigma_{n}(t)-C \sqrt{n} \sigma_{n}(t)-0(1), \tag{3.33}
\end{equation*}
$$

where $C$ is a constant. Since for $n$ large, $\frac{1}{2} \delta n-C \sqrt{n} \geqslant \frac{\delta}{4} n$,

$$
\begin{equation*}
\left(\exp \left(-\frac{\delta}{4} n t\right) S_{n}(t)\right)_{t}^{\prime} \geqslant-0(1) \exp \left(-\frac{\delta}{4} n t\right) \tag{3.34}
\end{equation*}
$$

Now we know that, since at $t=0$ all functions $\theta^{(n)}(p, r, \kappa) N_{0} u_{n}$ have their components 0 except the first components, we have $S_{n}(0) \geqslant S\left(0 ; \gamma_{n} \beta(x) N_{0} u_{n}\right)=\left\|\alpha_{n} \beta(x) \psi_{n}(x)\right\|^{2}, \quad$ and this is equal to $c^{2}+0\left(\frac{1}{n}\right)$, by (3.3). Hence

$$
\begin{equation*}
S_{n}(0) \geqslant c^{2} / 2 \quad \text { for } n \text { large } \tag{3.35}
\end{equation*}
$$

By integration of (3.34), we have

$$
\begin{equation*}
S_{n}(t) \geqslant \frac{c^{2}}{2} \exp \left(\frac{\delta}{4} n t\right)-0\left(\frac{1}{n}\right) \quad \text { for } n \text { large } \quad 0 \leqslant t \leqslant T^{\prime} \tag{3.36}
\end{equation*}
$$

On the other hand, taking into account of (3.21) and (3.7), we see that $\left\|\theta^{(n)}(p, \gamma, \kappa) N_{0} u_{n}\right\|^{2} \leqslant$ const. $n^{|\gamma|-|\kappa|}\left\|\beta^{\kappa \kappa)}(x) u_{n}(x, t)\right\|^{2}$ $=0\left(n^{m+2 h}\right)=0\left(n^{4 h+2 l}\right)$.
Since $\quad S_{n}(t) \leqslant \sigma_{n}(t)$, it follows

$$
\begin{equation*}
S_{n}(t) \leqslant 0\left(n^{4 h+2 l}\right) \quad \text { for } \quad 0 \leqslant t \leqslant T \tag{3.37}
\end{equation*}
$$

(3.36) and (3.37) cannot be compatible unless $t=0$. The proof of Theorem 1.1 is thus complete.

## 4. Proof of Theorem 1.2.

In principle, this proof is the same as that of Theorem 1.1. Therefore, we want to explain how to reduce this proof to the previous one. We treat here the kowalewskians $(p=1)$.

At first, denoting $u_{1}=u, u_{2}=\frac{\partial}{\partial t} u, \cdots, u_{m}=\frac{\partial^{m-1}}{\partial t^{m-1}} u$, we consider an equivalent system to (1.3):
(4.1) $L[u]=\frac{\partial}{\partial t} u-(A+J)[u]=0$, where

$$
A=\left\lvert\, \begin{array}{cc} 
\\
0 \\
-a_{m}-a_{m-1} \cdots-a_{1}
\end{array}\right., \quad, \quad J=\left(\begin{array}{cccc}
0 & 1 & & \\
& 0 & 1 & \\
& & \ddots & \ddots \\
& & \ddots & 1 \\
& & & 0
\end{array}\right), \quad u=\left(\begin{array}{c}
u_{1} \\
u_{2} \\
\vdots \\
\vdots \\
u_{m}
\end{array}\right) .
$$

Apply $\beta(x)$ to $L[u]$, then

$$
\begin{gather*}
L[\beta(x) u]+\sum_{1 \leqslant|v| \leqslant m} C_{\nu}\left(x, t ; \frac{\partial}{\partial x}\right)\left[\beta^{v \nu}(x) u\right]=0,  \tag{4.2}\\
\text { where, denoting } C_{\nu}=\left(\begin{array}{c} 
\\
0 \\
c_{1, \nu} c_{2, v} \cdots c_{m, \nu}
\end{array}\right),
\end{gather*}
$$

(4.3) $c_{i, \nu}(x, t ; \xi)$ is of order (in $\left.\xi\right) \leqslant m+1-i-|\nu|$.

Now we apply $\gamma_{n}(t)$ to (4.2) just in the same way as in the section 2 , we then have
(4.4) $L\left[\gamma_{n} \beta(x) u\right]+\left(\gamma_{n} L-L \gamma_{n}\right)[\beta(x) u]+\gamma_{n} \sum C_{\nu}\left[\beta^{(\nu)}(x) u\right]=0$.

Taking into account of (4.1), we have here

$$
\begin{equation*}
\gamma_{n} L-L \gamma_{n}=\gamma_{n} A-A \gamma_{n} \tag{4.5}
\end{equation*}
$$

Now we insert a process which was not done in the case of the first order system. Denote

$$
E_{m}(\Lambda)=\left(\begin{array}{lllll}
(\Lambda+1)^{m-1} & & &  \tag{3.6}\\
& (\Lambda+1)^{m-2} & & \\
& & \ddots & \\
& & & 1
\end{array}\right)
$$

Instead of $u$ itself, we consider $E_{m}(\Lambda) u$. Then it follows from (4.2),
(4. 7) $\quad\left(E_{m} L E_{m}^{-1}\right)\left[E_{m} \gamma_{n} \beta(x) u\right]+\left(\gamma_{n} A-A \gamma_{n}\right) E_{m}^{-1}\left[E_{m} \beta(x) u\right]$

$$
+\gamma_{n} \sum_{1 \leqslant|v| \leqslant m} C_{\nu}\left(x, t ; \frac{\partial}{\partial x}\right) E_{m}^{-1}\left[E_{m} \beta(x) u\right]=0 .
$$

Now consider the first term. We can express it by singular integral operators :

$$
\begin{equation*}
\left(\frac{\partial}{\partial t}-\mathscr{H} \Lambda-\mathcal{B}\right)\left[\gamma_{n} E_{m} \beta(x) u\right], \tag{4.8}
\end{equation*}
$$

where
i) $\mathscr{B}$ is a bounded operator in $L^{2}$;
(4. 9)

$$
\left(\begin{array}{ccccc}
0 & 1 & & & \\
& 0 & 1 & & \\
& & & & \\
& & & \ddots & \\
& & & \ddots & \\
& & & & \\
-h_{m} & -h_{m-1} & & \\
& -h_{1}
\end{array}\right) ; \quad h_{i}=h_{i}(x, t ; i \xi /|\xi|) .
$$

Next, we see that $\left(\gamma_{n} A-A \gamma_{n}\right) E_{m}^{-1}$ has just the same property as $\sum\left(\gamma_{n} A_{k}-A_{k} \gamma_{n}\right) \frac{\partial}{\partial x_{k}}$ in (2.8). Finally, denoting $C_{\nu} E_{m}^{-1}=C_{\nu}^{\prime}$, we see that the term $c_{i, \nu}^{\prime}$ corresponding to $c_{i, \nu}$ has the order of differentiation $\leqslant(m+1-i-|\nu|)-(m-i)=1-|\nu|$. This shows that $\gamma_{n} \sum C_{v}^{\prime}\left[\beta^{(\nu)}(x) u\right]$ has the same property as $\gamma_{n} \sum A_{k} \beta_{k} u$ in (2.8).

Finally, we define $u_{n}(x, 0)$ by

$$
N_{0} E_{m}(\Lambda) u_{n}(x, 0)=\left(\begin{array}{c}
\psi_{n}(x)  \tag{4.10}\\
0 \\
\vdots \\
0
\end{array}\right),
$$

where $N_{0}$ is defined by $\sigma(\mathscr{H})$ in the same way as in (2.10). (4.10) can be written
$(4.10)^{\prime}$

$$
u_{n}(x, 0)=E_{m}(\Lambda)^{-1} N_{0}^{-1}\left(\begin{array}{c}
\psi_{n}(x) \\
0 \\
\vdots \\
0
\end{array}\right) .
$$

Obviously, $u_{n}(x, 0) \in \mathscr{D}_{1,2}^{\infty} \cap \mathcal{B}$. Finally we remark that
(4.11)

$$
\begin{aligned}
& \left\|\alpha_{n} N_{0} E_{m} \beta(x) u_{n}(x, 0)\right\|_{L^{2}} \geqslant\left\|\beta(x) \alpha_{n} N_{0} E_{m} u_{n}(x, 0)\right\| \\
& \quad-\left\|\left[\beta(x)\left(\alpha_{n} N_{0} E_{m}\right)-\left(\alpha_{n} N_{0} E_{m}\right) \beta(x)\right] u_{n}(x, 0)\right\| \\
& =\left\|\beta(x) \psi_{n}(x)\right\|-0\left(\frac{1}{n}\right), \text { by the same reason as in (3.6). }
\end{aligned}
$$

Thus we can verify easily that, replacing $\alpha_{n}^{(\gamma)} \beta^{(\kappa)}(x) N_{0} u_{n}$ in the previous section by $\left(N_{0} E_{m}(\Lambda)\right) \alpha_{n}^{(\gamma)} \beta^{\kappa \kappa)}(x) u_{n}$ the analogous reasoning can be carried out.

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[^0]:    2) We should mention that our proof is based on the theory of the singular integral operators (in the sense of Calderón and Zygmund). More precisely, we used the localisation of singular integral operators both in $x$ and $\xi$ space. We showed its utility in our previous paper: Le problème de Cauchy pour les systèmes hyperboliques et paraboliques. Mem. Coll. Sci. Univ. Kyoto, 32, 181-212 (1959).
[^1]:    3) Throughout this paper we may use the symbol $C$ in order to re- present positive constants. Sometimes it expresses a positive constant which can be chosen independently of $n$.
