# 50. Cauchy Problem for Degenerate Parabolic Equations 

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1. Introduction. We consider the Cauchy problem for the equation

$$
\begin{align*}
& \partial_{t} u-\sum_{j, k=1}^{n} \partial_{x_{j}}\left(a_{j k}(x, t) \partial_{x_{k}} u\right)-\sum_{j=1}^{n} b_{j}(x, t) \partial_{x_{j}} u-c(x, t) u  \tag{1.1}\\
& \quad=\partial_{t} u-A u=f
\end{align*}
$$

$(x, t)$ in $\boldsymbol{R}^{n} \times[0, \infty)$ with the initial-value

$$
\begin{equation*}
u(x, 0)=u_{0}(x), \tag{1.2}
\end{equation*}
$$

where $a_{j_{k}}(x, t), b_{j}(x, t), c(x, t)$ are real-valued smooth functions. We assume that $\left(a_{j k}\right)_{1 \leq j \leq n, 1 \leq k \leq n}$ is symmetric and satisfies the condition: for any $(x, t) \in \boldsymbol{R}^{n} \times[0, \infty)$

$$
\begin{equation*}
\sum_{j, k=1}^{n} a_{j k}(x, t) \xi_{j} \xi_{k} \geq 0 \quad \text { for all } \xi \in \boldsymbol{R}^{n} \tag{1.3}
\end{equation*}
$$

O. A. Oleǐnik has treated this problem (see [3] and [4]). Her method consists of the following procedure (elliptic regularization): Instead of (1.1), the following equations (depending on a positive parameter $\varepsilon$ ) in $G=R^{n} \times[0, T]$

$$
\begin{equation*}
\partial_{t} u-\varepsilon \Delta u-A u=f \tag{1.4}
\end{equation*}
$$

are considered. Let $u_{s}$ be the solution of (1.4) with the given initialvalue $u_{0}(x) \in L^{2}\left(\boldsymbol{R}^{n}\right)$ and $f(x, t) \in L^{2}(G)$. Then it is shown that $\left\{u_{s}(x, t)\right\}$ is bounded in $L^{2}(G)$. Then a weak limit of them, as $\varepsilon \rightarrow+0$, gives the desired solution $u(x, t) \in L^{2}(G)$. The uniqueness of the solution is proved. She also proved the smoothness of $u$, assuming the smoothness of $u_{0}$ and $f$.

Contrary to the above point of view, we regard (1.1) as evolution equation. More precisely, we want to show the existence of the unique solution $u(x, t) \in \mathcal{E}_{t}^{0}\left(L^{2}\right) \cap \mathcal{E}_{t}^{1}\left(\mathscr{D}_{L^{2}}^{\prime 2}\right)$ of (1.1)-(1.2) for any $f(x, t) \in \mathcal{E}_{t}^{0}\left(L^{2}\right)$ and any initial-value $u_{0}(x) \in L^{2}$.*)

Our approach is based on the semi-group theory. Instead of elliptic regularization, we use Friedrichs' mollifier. Its property (see

[^0]Lemma) gives immediately the desired result (see Theorem 1). The smoothness can be obtained in the following form: when $u_{0}(x) \in \mathscr{D}_{L^{2}}^{m}$ and $f(x, t) \in \mathcal{E}_{t}^{0}\left(\mathscr{D}_{L^{2}}^{m}\right)$, the solution $u(x, t)$ belongs to $\in \mathcal{E}_{t}^{0}\left(\mathscr{D}_{L^{2}}^{m}\right) \cap \mathcal{E}_{t}^{1}\left(\mathscr{D}_{L^{2}}^{m-2}\right)$.

It seems to us that our method is more natural than the one relying on elliptic regularization and will be useful to other problems. A forthcoming paper will give the detailed proof including some other results.
2. Statement of results. Let $a_{j k}(x, t) \in \mathcal{E}_{t}^{0}\left(\mathcal{B}^{2}\right) ; b_{j}(x, t) \in \mathcal{E}_{t}^{0}\left(\mathcal{B}^{1}\right)$; $c(x, t) \in \mathcal{E}_{t}^{0}\left(\mathcal{B}^{0}\right)$. We assume the condition (1.3). Then we have the following theorem.

Theorem 1. For any initial-value $u_{0}(x) \in L^{2}$ and any $f(x, t) \in \mathcal{E}_{t}^{0}\left(L^{2}\right)$, there exists a unique solution $u(x, t) \in \mathcal{E}_{t}^{0}\left(L^{2}\right) \cap \mathcal{E}_{t}^{1}\left(\mathscr{D}_{L^{2}}^{\prime 2}\right)$ of the Cauchy problem (1.1)-(1.2).

To prove this, following propositions are essential. The first one is the energy inequality. The second one shows that Hille-Yosida's theorem is applicable.

Proposition 1. Let $f(x, t) \in \mathcal{E}_{t}^{0}\left(L^{2}\right)$ and $u(x, t) \in \mathcal{E}_{t}^{0}\left(L^{2}\right) \cap \mathcal{E}_{t}^{1}\left(\mathscr{D}_{L^{2}}^{\prime 2}\right)$ be the solution of (1.1). Then it holds for any $t(0 \leq t \leq T)$

$$
\begin{equation*}
\|u(t)\| \leq e^{r t}\|u(0)\|+\int_{0}^{t} e^{r(t-s)}\|f(s)\| d s \tag{2.1}
\end{equation*}
$$

where $\gamma$ is a constant which may depend on $T$ but does not depend on $u$ and $f$.

Now we assume coefficients be functions of only $x$. Then we can obtain the following proposition.

Proposition 2. Take the domain of definition $\mathscr{D}(A)$ of $A$ as follows:

$$
\begin{equation*}
\mathscr{D}(A)=\left\{u ; u \in L^{2}, A u \in L^{2}\right\} . \tag{2.2}
\end{equation*}
$$

Then, for large $\lambda,(\lambda-A)$ defines a one-to-one surjective mapping of $\mathscr{D}(A)$ onto $L^{2}$. Moreover there exists a constant $\beta$ such that

$$
\begin{equation*}
\left\|(\lambda-A)^{-1}\right\|_{\mathcal{L}\left(L^{2}, L^{2}\right)} \leq \frac{1}{\lambda-\beta} \quad \text { for any } \lambda>\beta \tag{2.3}
\end{equation*}
$$

If we use the following lemma, these propositions can be proved in the same way as hyperbolic equation (see [2], $\S \S 2,4$ in Chapter 6).

Lemma. Let $\rho_{\varepsilon} *$ be Friedrichs' mollifier, where we assume $\rho(x)$ even function. Let $a(x) \in \mathscr{B}^{2}$ be real-valued function, and let $u(x) \in L^{2}$. Then it holds for any $\nu(|\nu| \leq 2)$

$$
\left|\operatorname{Re}\left(u_{s},\left[\rho_{s} *, a(x)\right] \partial_{x}^{\nu} u\right)\right| \leq C\|u\|,
$$

2) $\quad \operatorname{Re}\left(u_{s},\left[\rho_{s} *, a(x)\right] \partial_{x}^{\nu} u\right) \rightarrow 0 \quad$ as $\varepsilon \rightarrow+0$,
where $u_{s}$ stands for $\rho_{\epsilon} * u,\left[\rho_{\varepsilon} *, a(x)\right] \partial_{x}^{\nu} u=\rho_{s} *\left\{a(x) \partial_{x}^{\nu} u\right\}-a(x) \rho_{\varepsilon} * \partial_{x}^{\nu} u$, and $C$ is a constant independent of $u$ and $\varepsilon$.

Proof of Lemma. Consider only the case of $|\nu|=2$, because, in the case of $|\nu| \leq 1,1)$ and 2) are clear by Friedrichs' lemma. We denote $\partial_{x}^{\nu}$ by $\partial_{j} \partial_{k}$. By Taylor expansion

$$
\begin{align*}
{\left[\rho_{s} *, a(x)\right] \partial_{j} \partial_{k} u=} & -\sum_{i=1}^{n} a^{(i)}(x)\left(x_{i} \rho_{s}\right) * \partial_{j} \partial_{k} u \\
& +\sum_{|\mu|=2} \frac{1}{\mu!} \int a_{\mu}(x, y)(x-y)^{\mu} \rho_{s}(x-y) \partial_{y_{j}} \partial_{y_{k}} u(y) d y \tag{2.4}
\end{align*}
$$

At first take the 1-st term of (2.4).

$$
\begin{align*}
& -2 \operatorname{Re}\left(u_{s}, a^{(i)}(x)\left(x_{i} \rho_{\mathrm{s}}\right) * \partial_{j} \partial_{k} u\right)=\left(a^{(i, j)}(x)\left(x_{i} \rho_{\mathrm{s}}\right) * \partial_{k} u, u_{s}\right) \\
& \quad+\left(u_{s},\left[\left(x_{i} \rho_{\mathrm{s}}\right) *, a^{(i)}(x)\right] \partial_{j} \partial_{k} u\right)+\left(u_{s},\left(x_{i} \rho_{\mathrm{s}}\right) *\left\{a^{(i, k)}(x) \partial_{j} u\right\}\right)  \tag{2.5}\\
& \quad+\left(\left(x_{i} \rho_{\mathrm{s}}\right) * \partial_{k} u,\left[a^{(i)}(x), \rho_{\mathrm{s}} *\right] \partial_{j} u\right),
\end{align*}
$$

where we used the relations that $\left(\left(x_{i} \rho_{s}\right) * u, v\right)=-\left(u,\left(x_{i} \rho_{s}\right) * v\right)$ and that $\left(\rho_{e} * u, v\right)=\left(u, \rho_{c} * v\right)$. These all terms in the right-hand side of (2.5) can be majorized by

$$
\sum_{i, j=1}^{n}\|u\|\left\|\left(x_{i} \rho_{\mathrm{s}}\right) * \partial_{j} u\right\|+\sum_{|\nu| \leq 2} \sum_{i=1}^{n}\|u\|\left\|\left[a^{(i)}(x),\left(x_{i} \rho_{\mathrm{c}}\right) *\right] \partial_{x}^{\nu} u\right\| .
$$

In the same way as Friedrichs' lemma, we can show for $\nu(|\nu| \leq 2)$

$$
\begin{gathered}
\left\|\left(x_{i} \rho_{\mathrm{s}}\right) * \partial_{j} u\right\|,\left\|\left[a^{(i)}(x),\left(x_{i} \rho_{\mathrm{s}}\right) *\right] \partial_{x}^{\nu} u\right\| \leq \text { const. }\|u\|, \\
\left\|\left(x_{i} \rho_{\mathrm{s}}\right) * \partial_{j} u\right\|,\left\|\left[a^{(i)}(x),\left(x_{i} \rho_{\mathrm{s}}\right) *\right] \partial_{x}^{\nu} u\right\| \rightarrow 0 \quad \text { as } \varepsilon \rightarrow+0 .
\end{gathered}
$$

Next we consider the 2 -nd term of (2.4). Denote it by $R_{s} u$.

$$
\begin{align*}
R_{\epsilon} u & =\sum_{|\mu|=2} \frac{1}{\mu!} \int \partial_{y_{k}} \partial_{y_{j}}\left\{a_{\mu}(x, y)(x-y)^{\mu} \rho_{s}(x-y)\right\} u(y) d y \\
& =\sum_{|\mu|=2} \frac{1}{\mu!} \int \partial_{y_{k}} \partial_{y_{j}}\left\{a_{\mu}(x, y)(x-y)^{\mu} \rho_{s}(x-y)\right\}\{u(y)-u(x)\} d y . \tag{2.6}
\end{align*}
$$

If we note that $\sum_{|| | \leq \leq 2} \sum_{|\mu|=2} \int\left|\left(x^{\mu} \rho_{c}\right)^{(\nu)}(x)\right| d x<$ const. (independent of $\varepsilon$ ), the same reasoning as in the proof of Friedrichs’ lemma gives

$$
\begin{aligned}
& \left\|R_{\mathrm{s}} u\right\| \leq \text { const. }\|u\|, \\
& R_{s} u \rightarrow \infty \quad \text { as } \varepsilon \rightarrow \infty \text {. }
\end{aligned}
$$

Thus the proof is completed.
At the end we state the theorem concerning the smoothness of the solution. Let $a_{j k}(x, t) \in \mathcal{E}_{t}^{0}\left(\mathscr{B}^{m+2}\right) ; b_{j}(x, t) \in \mathcal{E}_{t}^{0}\left(\mathcal{B}^{m+1}\right) ; c(x, t) \in \mathcal{E}_{t}^{0}\left(\mathscr{B}^{m}\right)$, where $m=0,1,2, \cdots$. We assume the condition (1.3). Then we have the following theorem.

Theorem 2. For any initial-value $u_{0}(x) \in \mathscr{D}_{L^{2}}^{m}$ and any $f(x, t) \in \mathcal{E}_{t}^{0}\left(\mathscr{D}_{L^{2}}^{m}\right)$, there exists a unique solution $u(x, t) \in \mathcal{E}_{t}^{0}\left(\mathscr{D}_{L^{2}}^{m}\right) \cap \mathcal{E}_{t}^{1}\left(\mathscr{D}_{L^{2}}^{m-2}\right)$ of the Cauchy problem (1.1)-(1.2).

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## References

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[3] O. A. Oleǐnik: On the smoothness of the solutions of degenerate elliptic and parabolic equations. Sov. M. Dokl., 6, 972-976 (1965).
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[^0]:    *) Throughout this paper, we use the following notation: $x=\left(x_{1}, \cdots, x_{n}\right)$. $\partial_{t}$ $=\partial / \partial t, \partial_{j}=\partial x_{j}=\partial / \partial x_{j}, \partial_{x}^{\nu}=\partial_{1}^{\nu_{1}} \cdots \partial_{n}^{\nu n}$, where $\nu=\left(\nu_{1}, \cdots, \nu_{n}\right) . L^{2}=L^{2}\left(\boldsymbol{R}^{n}\right) . u(x) \in \mathscr{D}_{L^{2}}^{m}$ means that its derivatives (in the sense of distribution) $\partial_{x}^{\nu} u$ up to order $m$ belong to $L^{2} . \mathscr{D}_{L^{2}}^{\prime m}$ is the dual space of $\mathscr{D}_{L^{2}}^{m}$ and sometimes we denote it by $\mathscr{D}_{L^{2}}^{-m} . \varphi(x) \in \mathscr{G}^{m}$ means that its derivatives $\partial_{x}^{v} \varphi$ up to order $m$ are continuous and bounded in $\boldsymbol{R}^{n}$. $f(t) \in \mathcal{E}_{t}^{k}\left(\mathscr{D}_{L^{2}}^{m}\left(\right.\right.$ or $\left.\mathscr{B}^{m}\right)$ ) means that $t \rightarrow t(t) \in \mathscr{D}_{L^{2}}^{m}\left(\right.$ or $\left.\mathscr{B}^{m}\right)$ is continuously differentiable up to order $k$.

