# A Necessary Condition for Hypoellipticity of Degenerate Elliptic-Parabolic Operators 

Kazuo AMANO<br>Tokyo Metropolitan University<br>(Communicated by T. Sirao)

## Introduction

The aim of this paper is to study hypoellipticity of degenerate elliptic-parabolic operators from the view point of the control theory. Hörmander and Oleǐnik-Radkevič proved (see [4]) that the degenerate elliptic-parabolic operator

$$
\begin{equation*}
L=\sum_{i, j=1}^{d} a_{i j}(x) \frac{\partial^{2}}{\partial x_{i} \partial x_{j}}+\sum_{i=1}^{d} b_{i}(x) \frac{\partial}{\partial x_{i}}+c(x) \tag{1}
\end{equation*}
$$

in an open set $M$ in $\boldsymbol{R}^{d}$ with real $C^{\infty}$-smooth coefficients is hypoelliptic if $\operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right) \equiv d$ (for the notation, see $\S 1$ ), where

$$
X_{0}=\sum_{i=1}^{d}\left(b_{i}-\sum_{j=1}^{d} \frac{\partial a_{i j}}{\partial x_{j}}\right) \frac{\partial}{\partial x_{i}},
$$

$$
\begin{equation*}
X_{i}=\sum_{j=1}^{d} a_{i j} \frac{\partial}{\partial x_{j}}, \quad 1 \leqq i \leqq d, \tag{2}
\end{equation*}
$$

and conversely, when the coefficients are real analytic, $\operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}\right.$, $\left.\cdots, X_{d}\right) \equiv d$ if the operator $L$ is hypoelliptic. Chow and Nagano proved (see [7]) that for a set of $C^{\infty}$-smooth vector fields $\left\{X_{0}, X_{1}, \cdots, X_{d}\right\}$ the system

$$
\begin{equation*}
\dot{x}=\sum_{i=0}^{d} \xi_{t} X_{i}(x), \quad \xi_{i} \in R^{1} \tag{3}
\end{equation*}
$$

is controllable in every subdomain in $M$ if $\operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right) \equiv d$, and proved that the converse proposition holds when the vector fields are real analytic. Thus we are led naturally to the following problems:

[^0]Is the system (3) controllable in every subdomain in $M$ if the operator $L$ is hypoelliptic?; conversely, when the vector fields $X_{0}, X_{1}, \cdots, X_{d}$ are real analytic, is the operator $L$ hypoelliptic if the system (3) is controllable in every subdomain in $M$ ? We give an affirmative answer for the former problem (Theorem 2). In view of Hörmander-Oleǐnik-Radkevič's and Nagano's results the latter problem is trivial, so we modify the question: When the vector fields $X_{0}, X_{1}, \cdots, X_{d}$ are merely $C^{\infty}$-smooth, does the answer for the latter problem remain affirmative? The answer is negative in general, but we can show that there is a closed set $F$ in $M$ such that $\dot{F}=\varnothing, F \subset\left\{x \in M ; \operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)<d\right.$ at $\left.x\right\}$ and such that $L$ is hypoelliptic in $M \backslash F$ if the system (3) is controllable in every subdomain in $M$ (Theorem 1 and Remark 1).

In Section 1, Theorem 1 is proved. In Section 2, Theorem 2 is reduced to a certain proposition which is proved in Section 4 by using some probabilistic lemmas prepared in Section 3. We can see easily that the whole statement of this paper remains true when $M$ is a $C^{\infty}$-manifold.

Notations:
$C^{k}(V)$ is the set of all $C^{k}$ functions defined in $V$.
$C_{0}^{k}(V)$ is the set of all functions in $C^{k}(V)$ with compact support in $V$. $\mathscr{D}^{\prime}(V)$ is the set of all distributions in $V$.

## § 1. Proof of Theorem 1.

Let $M$ be an open set in $R^{d}$ and let $L$ be a differential operator in $M$ of the form (1) with real $C^{\infty}$-smooth coefficients. Throughout this paper we assume that $\left(a_{i j}(x)\right)$ is a nonnegative symmetric $d \times d$ matrix for every $x$ in $M$, that is, $L$ is the degenerate elliptic-parabolic operator in $M$. Furthermore we assume that the second order terms and the first order ones of $L$ never vanish simultaneously, i.e.,

$$
\begin{equation*}
\sum_{i, j=1}^{d}\left|a_{i j}(x)\right|+\sum_{i=1}^{d}\left|b_{i}(x)\right| \neq 0 \tag{4}
\end{equation*}
$$

for all $x$ in $M . \quad X_{0}, X_{1}, \cdots, X_{d}$ will denote the vector fields defined by (2) and $\mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)$ will denote the Lie algebra generated over $\boldsymbol{R}$ by the vector fields $X_{0}, X_{1}, \cdots, X_{d}$.

It is easy to show the following lemma which will be used in the proof of Theorem 1.

Lemma 1. If $\operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right) \leqq r$ in an open set $U$ in $M$, then the set $\left\{x \in U ; \operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)=r\right.$ at $\left.x\right\}$ is open.

Theorem 1. If the system (3) is controllable in every subdomain in $M$, then the set $\left\{x \in M ; \operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)<d\right.$ at $\left.x\right\}$ is closed in $M$ and has no interior.

Proof. The closedness follows immediately from Lemma 1. If there is an open set $U$ in which $\operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)<d$, then

$$
\max _{x \in U} \operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)=r<d
$$

Lemma 1 shows that $\operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right) \equiv r$ in some non-empty domain, say $V$, contained in $U$, so $\mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)$ is an $r$-dimensional involutive distribution in $V$. By the Frobenius' theorem, $V$ is able to be cut into slices of $r$-dimensional integral manifolds of the distribution $\mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)$. Hence the system (3) is not controllable in $V$; this is a contradiction.

Remark 1. By combining Theorem 1 with Hörmander-OleǐnikRadkevič's (see [4]) and Nagano's results (see [7]) we obtain the following: If the system (3) is controllable in every subdomain in $M$, then there is a closed set $F$ in $M$ such that $\dot{F}=\varnothing, F \subset\left\{x \in M\right.$; $\operatorname{dim} \mathscr{L}\left(X_{0}, X_{1}, \cdots, X_{d}\right)<$ $d$ at $x\}$ and such that the operator $L$ is hypoelliptic in $M \backslash F$. In particular, when the coefficients $a_{i j}$ and $b_{i}$ are real analytic, we have $F=\varnothing$. It is generally not possible to show $F=\varnothing$, although Fediǐ [1] actually proved this for a certain kind of infinitely degenerate elliptic-parabolic operators.

## § 2. Proof of Theorem 2 (Part 1).

We summarize Sussmann's results ([7]) which are necessary in proving Theorem 2. Let $D$ be the set of vector fields $\left\{X_{0}, X_{1}, \cdots, X_{d}\right\}$ and let $\Delta_{D}$ be the distribution spanned by $D$. For an open set $U$ in $M, G_{D}(U)$ will denote the group of local $C^{\infty}$-diffeomorphisms on $U$ generated by $\left.D\right|_{U}$ (cf. [7]). Sussmann's distribution $S_{D}(U)$ is the smallest $G_{D}(U)$-invariant distribution on $U$ which contains $\left.\Delta_{D}\right|_{U}$, that is, the space $S_{D}(U)(x)$ is the linear hull of all the vectors $v \in T_{x} U$ such that $v \in \Delta_{D}(x)$ or $v=d \varphi(w)$, where $\varphi \in G_{D}(U)$ and, for some $y \in U, x=\varphi(y)$ and $w \in \Delta_{D}(y)$. The distribution $S_{D}(U)$ has the maximal integral manifolds property in the sense of [7] and further, the system (3) is controllable in $U$ if and only if $\operatorname{dim} S_{D}(U) \equiv d$.

Theorem 2. If the operator $L$ is hypoelliptic in $M$, then the system (3) is controllable in every subdomain in $M$.

Proof (Part 1). Assume that there is a subdomain $U$ in $M$ such
that the system (3) is not controllable in $U$. By the Sussmann's result this means that $\operatorname{dim} S_{D}(U)<d$ at some point $p$ in $U$. Since $S_{D}(U)$ has the maximal integral manifolds property, there passes a maximal integral manifold of $S_{D}(U)$ through the point $p$. So it is easy to show, by the inverse function theorem, that there passes a regular maximal integral manifold, say $N$, of $S_{D}(U)$ through $p$. By (4) and $\operatorname{dim} S_{D}(U)(p)<d$, we have easily $1 \leqq \operatorname{dim} N<d$.

Since $N$ is regular and since hypoellipticity is a local property, we may suppose by performing a suitable change of local coordinates that

$$
\begin{equation*}
N=\left\{x \in R^{d} ; x_{r+1}=\cdots=x_{d}=0\right\} \tag{5}
\end{equation*}
$$

in a neighborhood of $p$, where $r=\operatorname{dim} N$; furthermore, by considering $\psi L$ instead of $L$ (where $0 \leqq \psi \leqq 1$ is a $C^{\infty}$-smooth function in $R^{d}$ such that $\psi \equiv 0$ outside a small neighborhood of $p$ and that $\psi \equiv 1$ in a smaller neighborhood of $p$ ) we may suppose that

$$
\begin{equation*}
a \in C_{b d d}^{\infty}\left(R^{d}, S_{d}\right), \quad b \in C_{b d d}^{\infty}\left(R^{d}, R^{d}\right), \quad c \in C_{b d d}^{\infty}\left(R^{d}, R^{1}\right) \tag{6}
\end{equation*}
$$

Here $a$ is the $d \times d$ matrix ( $a_{t j}$ ), b is the vector ( $b_{1}, \cdots, b_{d}$ ) and $S_{d}$ denotes the class of symmetric nonnegative matrices.

Now it will suffice to show the following: There is an open neighborhood $V$ of the point $p$ and a locally integrable function $u$ defined in $V$ such that $N \cap V \subset \operatorname{sing} \operatorname{supp} u$ and $L u=0$ in $V$.

Theorem 2 will be proved completely at the end of Section 4 after a preliminary study in §3.

## § 3. Probabilistic lemmas.

In this section we assume the conditions (6). $\sigma(x)=\left(\sigma_{i j}(x)\right)$ denotes a symmetric nonnegative $d \times d$ matrix such that $a(x)=(1 / 2) \sigma^{2}(x)$. By (6) each $\sigma_{i j}(x)$ is Lipschitz continuous in $R^{d}$, so for any $x$ in $R^{d}$ there exists a unique solution, say $x^{x}(t)$, with

$$
\begin{equation*}
d x(t)=\sigma(x(t)) d w(t)+b(x(t)) d t, \quad x(0)=x \quad \text { a.s. } \tag{7}
\end{equation*}
$$

in $M_{w}^{2}[0, T], T>0$. Here $w(t), t \geqq 0$, is a $d$-dimensional Brownian motion on a probability space $(\Omega, \mathscr{F}, P)$ and $M_{w}^{2}[0, T]$ denotes a set of all nonanticipative functions $f(t)$ satisfying

$$
E\left[\int_{0}^{T}|f(t)|^{2} d t\right]<\infty
$$

$x(t), t \geqq 0$, will denote the time-homogeneous diffusion process that is the
solution of the stochastic differential equation (7).
Definition ([3]). If $N$ is a subset in $R^{d}$ such that $P_{x}[x(t) \notin N$ for all $t \geqq 0]=1$ whenever $x \notin N$, then we say that $N$ is nonattainable by the process $x(t)$.

Lemma 2. Let $N$ be closed in $R^{d}$ and nonattainable by the process $x(t)$ and let $v \in C^{2}\left(R^{d} \backslash N\right)$. Then

$$
\begin{aligned}
& d\left[v(x(t)) \exp \left\{\int_{0}^{t} c(x(s)) d s\right\}\right] \\
& = \\
& \quad \nabla v(x(t)) \exp \left\{\int_{0}^{t} c(x(s)) d s\right\} \sigma(x(t)) d w(t) \\
& \quad+L v(x(t)) \exp \left\{\int_{0}^{t} c(x(s)) d s\right\} d t
\end{aligned}
$$

$P_{x}$-a.s. for all $x \notin N$.
This is a generalization of Itô's formula. For the proof we have only to approximate $v$ by suitable functions in $C^{2}\left(R^{d}\right)$ and use Ito's formula.

Let $V$ be an open set in $R^{d}$ with $C^{\infty}$-smooth boundary. The exit time $\tau$ of $\bar{V}$ is defined by

$$
\tau=\inf \{t \geqq 0 ; x(t) \notin \bar{V}\}
$$

$\Gamma$ and $\Sigma$ are the subsets on the boundary $\partial V$ of $V$ defined by

$$
\Gamma=\left\{x \in \partial V ; P_{x}(\tau>0)=0\right\}
$$

and

$$
\Sigma=\left\{x \in \partial V ;\langle\nu, a(x) \nu\rangle>0 \quad \text { or }\left\langle\nu, X_{0}(x)\right\rangle\langle 0\},\right.
$$

where $\langle$,$\rangle denotes the inner product in R^{d}$ and $\nu$ is the inward normal vector on $\partial V$. For brevity we set

$$
C=\sup _{x \in V} c(x) \vee 0 .
$$

Then we have the following lemma.
Lemma 3. Assume that

$$
\sup _{x \in \bar{V}} E_{x}\left[(1+\tau) e^{\sigma \tau}\right]<\infty
$$

For given $f \in L^{\infty}(V)$ and $g \in L^{\infty}(\Gamma) \cap C(\Sigma)$, the function

$$
\begin{aligned}
u(x)= & E_{x}\left[g(x(\tau)) \exp \left\{\int_{0}^{\tau} c(x(s)) d s\right\}\right. \\
& \left.-\int_{0}^{\tau} f(x(t)) \exp \left\{\int_{0}^{t} c(x(s)) d s\right\} d t\right]
\end{aligned}
$$

is a unique solution of the Dirichlet problem

$$
\left\{\begin{array}{l}
L u=f \quad \text { in } \quad V  \tag{8}\\
\operatorname{ess} \lim u(x)=g(a), \quad a \in \Sigma \\
x \rightarrow V \\
x \in V
\end{array}\right.
$$

in $L^{\infty}(V)$.
Here (8) means that $u$ can be changed on a set of Lebesgue measure zero so that the following relations may be satisfied;

$$
\left\{\begin{array}{l}
\int u L^{*} \varphi d x=\int f \varphi d x, \quad \varphi \in C_{0}^{\infty}(V), \\
\lim _{\substack{x \rightarrow a \\
x \in V}} u(x)=g(a), \quad a \in \Sigma
\end{array}\right.
$$

Lemma 3 is a generalization of Strook-Varadhan's theorem ([6]). The proof is parallel to that of Strook-Varadhan's theorem but we will have to use Lemma 2.1 in [5] instead of Lemma 4.1 in [6].

Lemma 4. Let $V_{\rho}$ be an open neighborhood of a fixed point $p$ in $R^{d}$, with diameter $V_{\rho}=\rho$, and let $\tau_{\rho}$ be the exit time of $\bar{V}_{\rho}$. If the condition (4) is satisfied at the point $p$, then we obtain

$$
\varlimsup_{\rho \neq 0} \sup _{x \in \bar{V}_{\rho}} E_{x}\left[e^{c \tau_{\tau}}\right]<\infty
$$

and

$$
\lim _{\rho \downharpoonright 0} \sup _{x \in \bar{V}_{\rho}} E_{x}\left[\left(\tau_{\rho} e^{C \tau}\right)^{k}\right]=0
$$

for any constant $C \geqq 0$ and any $k=1,2, \cdots$.
Proof. According to Freǐdlin [2] there are small positive constants $\varepsilon$ and $\delta$ independent of all sufficiently small $\rho>0$ such that

$$
P_{x}\left[\tau_{\rho}<\frac{\rho}{\delta}\right]>\varepsilon^{\rho / \delta}
$$

for all $x \in \bar{V}_{\rho}$. The Markov property gives

$$
P_{x}\left[\tau_{\rho} \geqq n \frac{\rho}{\delta}\right] \leqq\left(1-\varepsilon^{\rho / \delta}\right)^{n}
$$

for all $x \in \bar{V}_{\rho}$ and all $n=0,1,2, \cdots$; so we have

$$
\begin{aligned}
E_{x} & {\left[e^{C \tau_{\rho}}\right] } \\
& =\sum_{n=0}^{\infty} E_{x}\left[\chi_{n \rho / \delta \leq \tau_{\rho}<(n+1) \rho / \delta} e^{C \tau_{\rho}}\right] \\
& \leqq \sum_{n=0}^{\infty} e^{C(n+1) \rho / \delta} P_{x}\left[\tau_{\rho} \geqq n \frac{\rho}{\delta}\right] \\
& \leqq e^{C \rho / \delta} \sum_{n=0}^{\infty}\left\{e^{C \rho / \delta}\left(1-\varepsilon^{\rho / \delta}\right)\right\}^{n}
\end{aligned}
$$

Similarly we have

$$
\begin{aligned}
& E_{x}\left[\left(\tau_{\rho} e^{C_{\tau} \rho}\right)^{k}\right] \\
& \quad \leqq\left(\frac{\rho}{\delta} e^{\sigma_{\rho / \delta}}\right)^{k} \sum_{n=0}^{\infty}(n+1)^{k}\left\{e^{k C \rho / \delta}\left(1-\varepsilon^{\rho / \delta}\right)\right\}^{n}
\end{aligned}
$$

This completes the proof.
Remark 2. It follows immediately from Lemmas 3 and 4 that if the condition (4) is satisfied at a point $p$ in $M$, then the equation $L u=f$ is locally solvable at $p$, i.e., there is an open neighborhood $V$ of $p$ such that

$$
L \mathscr{D}^{\prime}(V) \supset C_{0}^{\infty}(V) .
$$

It is to be noted that the leading symbol of $L$ does not always admit an expression in the form of a sum of squares of symbols of principal type.

Lemma 5. Let $\tau$ be a stopping time such that $0 \leqq \tau<\infty$ a.s. and let $f(t)$ belong to $M_{w}^{2}[0, T]$ for each $T>0$. Then

$$
E\left[\sup _{0 \leq t \leq \tau}\left|\int_{0}^{t} f(s) d w(s)\right|^{2}\right] \leqq 4 E\left[\int_{0}^{\tau}|f(s)|^{2} d s\right]
$$

Proof. For any $T>0$ we easily have

$$
\begin{aligned}
& E\left[\sup _{0 \leqq t \leq \tau \wedge T}\left|\int_{0}^{t} f(s) d w(s)\right|^{2}\right] \\
& \quad=E\left[\sup _{0 \leq t \leq \tau \wedge T}\left|\int_{0}^{\tau \wedge t} f(s) d w(s)\right|^{2}\right] \\
& \quad \leqq E\left[\sup _{0 \leq t \leq T}\left|\int_{0}^{\tau \wedge t} f(s) d w(s)\right|^{2}\right] \\
& \quad \leqq 4 E\left[\int_{0}^{\tau \wedge T}|f(s)|^{2} d s\right]
\end{aligned}
$$

by the martingale inequality. Letting $T \rightarrow \infty$, the desired inequality follows.

## § 4. Proof of Theorem 2 (Part 2).

Proof of Theorem 2 (Part 2). Let us take a sufficiently small open neighborhood $V$ of the point $p$, with $C^{\infty}$-smooth boundary. Then one of the desired functions in Part 1 will be given by

$$
u(x)=E_{x}\left[g(x(\tau)) \exp \left\{\int_{0}^{\tau} c(x(s)) d s\right\}\right]
$$

where $g(x)=\log \left(\sum_{i=r+1}^{d} x_{i}^{2}\right)^{-1 / 2}$. Here $x(t)$ is the time-homogeneous diffusion process constructed from the solutions of the stochastic differential equation (7) and $\tau$ is the exit time of $\bar{V}$.

Since the vector fields $X_{i}, 0 \leqq i \leqq d$, are all tangential to $N$ at each point on $N, N$ is nonattainable by the process $x(t)$ (see [3, Section 9.4]). $(L-c) g(x)$ and $\nabla g(x) \sigma(x)$ are bounded functions in $V \backslash N$. In fact, by (5) and (6), $a_{i j}(x)=O\left(\sum_{i=r+1}^{d} x_{i}^{2}\right), b_{i}(x)=O\left(\sum_{i=r+1}^{d} x_{i}^{2}\right)^{1 / 2}$ and $\sigma_{i j}(x)=O\left(\sum_{i=r+1}^{d} x_{i}^{2}\right)^{1 / 2}$ as $d(x, N) \rightarrow 0$ for $r+1 \leqq i, j \leqq d$.

We first show that $u \in L_{1}^{\text {loc }}(V)$ and $u(x) \rightarrow \infty$ as $d(x, N) \rightarrow 0$, i.e., $N \cap$ $V \subset$ sing supp $u$. Since

$$
\nabla g\left(x^{x}(t)\right) \exp \left\{\int_{0}^{t} c\left(x^{x}(x)\right) d s\right\} \sigma\left(x^{x}(t)\right) \in M_{w}^{2}[0, T]
$$

for any $T>0$ if $x \in V \backslash N$, Lemma 2 gives

$$
\begin{aligned}
u(x)= & E_{x}\left[g(x(\tau)) \exp \left\{\int_{0}^{\tau} c(x(s)) d s\right\}\right] \\
= & g(x)+E_{x}\left[\int_{0}^{\tau}(L-c) g(x(t)) \exp \left\{\int_{0}^{t} c(x(s)) d s\right\} d t\right] \\
& +E_{x}\left[\int_{0}^{\tau} c(x(t)) g(x(t)) \exp \left\{\int_{0}^{t} c(x(s)) d s\right\} d t\right] \\
= & g(x)+I_{1}+I_{2}
\end{aligned}
$$

for all $x \in V \backslash N$ and clearly

$$
\begin{aligned}
& \left|I_{1}\right| \leqq C_{1} E_{x}\left[\tau e^{C \tau}\right] \\
& \left|I_{2}\right| \leqq C E_{x}\left[e^{c_{\tau}} \int_{0}^{\tau} g(x(t)) d t\right],
\end{aligned}
$$

where $C=\sup _{x \in V} c(x) \vee 0$ and $C_{1}=\sup _{x \in V \backslash N}|(L-c) g(x)|$. Furthermore, by Lemma 2,

$$
\begin{aligned}
C E_{x} & {\left[e^{C \tau} \int_{0}^{\tau} g(x(t)) d t\right] } \\
= & C E_{x}\left[e ^ { C _ { \tau } } \int _ { 0 } ^ { \tau } \left\{g(x)+\int_{0}^{t} \nabla g(x(s)) \sigma(x(s)) d w(s)\right.\right. \\
& \left.\left.\quad+\int_{0}^{t}(L-c) g(x(s)) d s\right\} d t\right] \\
\leqq & C E_{x}\left[\tau e^{\sigma_{\tau}}\right] g(x)+C\left(E_{x}\left[\left(\tau e^{\sigma \tau}\right)^{2}\right]\right)^{1 / 2} \\
& \times\left(E_{x}\left[\sup _{0 \leq t \leq \tau}\left|\int_{0}^{t} \nabla g(x(s)) \sigma(x(s)) d w(s)\right|^{2}\right]\right)^{1 / 2} \\
& +C C_{1} E_{x}\left[\frac{1}{2} \tau^{2} e^{C \tau}\right]
\end{aligned}
$$

and, by Lemma 5,

$$
\leqq C E_{x}\left[\tau e^{\sigma_{\tau}}\right] g(x)+4 C C_{2}\left(E_{x}\left[\left(\tau e^{C_{\tau}}\right)^{2}\right]\right)^{1 / 2}\left(E_{x}[\tau]\right)^{1 / 2}+C C_{1} E_{x}\left[\frac{1}{2} \tau^{2} e^{\sigma_{\tau}}\right]
$$

where $C_{2}=\sup _{x \in V \backslash N}|\nabla g(x) \sigma(x)|$, and so

$$
\leqq C E_{x}\left[\tau e^{C_{\tau}}\right] g(x)+2 C\left(C_{1}+C_{2}\right)\left\{E_{x}\left[\tau e^{C \tau}\right]+E_{x}\left[\left(\tau e^{C \tau}\right)^{2}\right]\right\} .
$$

Therefore, it will suffice to take diameter $V$ so small that $C \sup _{x \in \bar{V}} \boldsymbol{E}_{x}\left[\tau e^{\sigma \tau}\right]<1$ and $\sup _{x \in \bar{V}} E_{x}\left[\left(\tau e^{C \tau}\right)^{2}\right]<\infty$. Here we have used Lemma 4.

We next show that $L u=0$ in $V$. Set

$$
u_{n}(x)=E_{x}\left[g_{n}(x(\tau)) \exp \left\{\int_{0}^{\tau} c(x(s)) d s\right\}\right],
$$

where $g_{n}(x)=g(x) \wedge n$ and $n=1,2, \cdots$. Let us take diameter $V$ sufficiently small so that $u \in L_{1}^{10 c}(V)$ and $\sup _{x \in \bar{V}} E_{x}\left[(1+\tau) e^{C_{\tau}}\right]<\infty$. Then Lemma 3 shows

$$
\int u_{n} L^{*} \varphi d x=0, \quad \varphi \in C_{0}^{\infty}(V) ;
$$

this easily gives, by letting $n \rightarrow \infty$,

$$
\int u L^{*} \varphi d x=0, \quad \varphi \in C_{0}^{\infty}(V)
$$

The proof of Theorem 2 is now complete.
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Present Address:
Department of Mathematics
Tokyo Metropolitan University
Setagaya-ku, Tokyo 158


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