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## To cite this version:

Michel Benaïm, Stéphane Le Borgne, Florent Malrieu, Pierre-André Zitt. Qualitative properties of certain piecewise deterministic Markov processes. 2012. hal-00688920v1

HAL Id: hal-00688920

## https://hal.science/hal-00688920v1

Preprint submitted on 18 Apr 2012 (v1), last revised 7 Apr 2014 (v4)

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# Qualitative properties of certain piecewise deterministic Markov processes 

Michel Benaïm, Stephane Leborgne, Florent Malrieu, Pierre-André Zitt

April 18, 2012


#### Abstract

We study a class of Piecewise Deterministic Markov Processes with state space $\mathbb{R}^{m} \times E$ where $E$ is a finite set. The continous component evolves according to a smooth vector field that it switched at the jump times of the discrete coordinate. The jump rates may depend on the whole position of the process. Working under the general assumption that the process stays in a compact set, we detail a possible construction of the process and characterize its support, in terms of the solutions set of a differential inclusion. We establish results on the long time behaviour of the process, in relation to a certain set of accessible points, which is shown to be strongly linked to the support of invariant measures. Under Hörmander-type bracket conditions, we prove that there exists a unique invariant measure and that the processes converges to equilibrium in total variation. Finally we give examples where the bracket condition does not hold, and where there may be one or many invariant measures, depending on the jump rates between the flows.


## 1 Introduction and main results

Piecewise deterministic Markov processes (PDMPs in short) are intensively used in many applied areas (molecular biology [22], storage modelling [7], Internet traffic [13, 15, 16], neuronal activity [8, 20],...). Roughly speaking, a Markov process is a PDMP if its randomness is only given by the jump mechanism: in particular, it admits no diffusive dynamics. This huge class of processes has been introduced by Davis (see [9, 10]) in a general framework.

In the present paper, we deal with an interesting subclass of the PDMPs that plays a role in molecular biology (see [22, 8]) and elsewhere [23]. We consider a PDMP evolving on $\mathbb{R}^{m} \times E$, where $E$ is a finite set, as follows: the first coordinate moves continuously on $\mathbb{R}^{m}$ according to a smooth vector field that depends on the second coordinate whereas the second coordinate jumps with a rate depending on the first one. We are interested in the long time qualitative behaviour of these processes. A recent paper by Bakhtin and Hurth [2] consider the particular situation where the jump rates are constants and prove the beautiful result that, under an Hörmander type condition, if there exists an invariant measure for the process, then it is unique and absolutely continuous with respect to the "Lebesgue" measure on $\mathbb{R}^{m} \times E$. Here we consider a more general situation and focus on different questions. However certain of our results heavily rely on Bakhtin and Hurth's paper. This class of Markov processes is reminiscent of the so-called iterated random functions in the discrete time setting (see [11] for a good review of this topic).

Let us define our process more precisely. Let $E$ be a finite set, and for any $i \in E$, $F^{i}: \mathbb{R}^{m} \mapsto \mathbb{R}^{m}$ be a smooth vector field. We assume throughout that each $F^{i}$ is bounded
(i.e. $\left.\sup _{x}\left\|F_{i}(x)\right\| \leqslant L<\infty\right)$ and we let $\Phi^{i}=\left\{\Phi_{t}^{i}\right\}$ denote the flow induced by $F^{i}$. Recall that

$$
t \mapsto \Phi_{t}^{i}(x)=\Phi^{i}(t, x)
$$

is the solution to the Cauchy problem $\dot{x}=F^{i}(x)$ with initial condition $x(0)=x$. We furthermore assume that there exists a compact set $M \subset \mathbb{R}^{m}$ positively invariant under each $\Phi^{i}$. That is

$$
\begin{equation*}
\forall t \geqslant 0, \quad \Phi_{t}^{i}(M) \subset M \tag{1}
\end{equation*}
$$

We consider here a continuous time Markov process $\left(Z_{t}=\left(X_{t}, Y_{t}\right)\right)$ living on $M \times E$ whose infinitesimal generator acts on functions

$$
\begin{aligned}
g: M \times E & \rightarrow \mathbb{R} \\
(x, i) & \mapsto g(x, i)=g^{i}(x)
\end{aligned}
$$

smooth 1 in $x$, according to the formula

$$
\begin{equation*}
L g(x, i)=\left\langle F^{i}(x), \nabla g^{i}(x)\right\rangle+\sum_{j \in E} \lambda(x, i, j)\left(g^{j}(x)-g^{i}(x)\right) \tag{2}
\end{equation*}
$$

where
(i) $x \mapsto \lambda(x, i, j)$ is continuous;
(ii) $\lambda(x, i, j) \geqslant 0$ for $i \neq j$ and $\lambda(x, i, i)=0$;
(iii) For each $x \in M$, the matrix $(\lambda(x, i, j))_{i j}$ is irreducible.

The process is explicitly constructed in Section 2 and some of its basic properties are established. In Section 3 we describe (Theorem 3.2) the support of the law of the process in term of the solutions set of a differential inclusion induced by the collection $\left\{F^{i}: i \in E\right\}$. Section 4 introduces the accessible set which is a natural candidate to support invariant probabilities. We show (Proposition 4.8) that this set is compact, connected, strongly positively invariant and invariant under the differential inclusion induced by $\left\{F^{i}: i \in E\right\}$. In Section 5 basic properties of invariant and empirical occupation probabilities of the processes are established and their support is characterized in terms of the accessible set. In Section 6] we prove the regularity results of [2] in a slightly stronger form that is needed for the convergence estimates. We use these results in Section 7 to show that, under some Hörmander bracket conditions, the law of jump chain and the process converge exponentially in total variation toward the unique invariant probability of the process. Section 8 gives elementary examples, and we show in Section 9 that when the Hörmander condition is violated, the uniqueness of the invariant measure may depend on the jump mechanism between flows, and not only on the flows themselves.

Let us note here that we address related questions in two companion papers: [6] gives under more stringent assumptions an explicit rate of convergence in Wasserstein distance, via a coupling argument; [5] studies in detail an example where the invariance assumption (11) does not hold, and where the behaviour of the process depends strongly on the jump rates.

[^0]
## 2 Construction and Basic properties

Since $M$ is compact and the maps $\lambda(\cdot, i, j)$ are continuous, there exists $\lambda \in \mathbb{R}_{+}$such that

$$
\max _{x \in M, i \in E} \sum_{j \in E, j \neq i} \lambda(x, i, j)<\lambda .
$$

Let us fix such a $\lambda$, and let

$$
Q_{i j}(x)=\frac{\lambda(x, i, j)}{\lambda} \quad \text { for } i \neq j \text { and } \quad Q_{i i}(x)=1-\sum_{j \neq i} Q_{i j}(x) .
$$

Note that $Q(x)$ is an irreducible aperiodic Markov transition matrix and that (2) can be rewritten as

$$
L g=A g+\lambda(Q g-g)
$$

where

$$
A g(x, i)=\left\langle F^{i}(x), \nabla g^{i}(x)\right\rangle
$$

and

$$
Q g(x, i)=\sum_{j \in E} Q_{i j}(x) g^{j}(x)
$$

The process $\left(Z_{t}\right)$ can be constructed as follows. Let $U_{1}, U_{2}, \ldots$ be a sequence of independent identically distributed random variables having an exponential distribution with parameter $\lambda$ : for any $t>0$,

$$
\mathbb{P}\left(U_{i}<t\right)=1-e^{-\lambda t} ;
$$

and let $Z_{0} \in M \times E$ be a random variable independent of $\left(U_{i}\right)_{i \geqslant 1}$. Define a discrete time Markov chain ( $\left.\tilde{Z}_{n}=\left(\tilde{X}_{n}, \tilde{Y}_{n}\right)\right)$ on $M \times E$ recursively by $\tilde{Z}_{0}=Z_{0}$,

$$
\tilde{X}_{n+1}=\Phi^{\tilde{Y}_{n}}\left(U_{n+1}, \tilde{X}_{n}\right),
$$

and

$$
\mathbb{P}\left(\tilde{Y}_{n+1}=j \mid \tilde{X}_{n+1}, \tilde{Y}_{n}=i\right)=Q_{i j}\left(\tilde{X}_{n+1}\right) .
$$

Now, set $T_{0}=0$ and $T_{n+1}=T_{n}+U_{n+1}$ for any $n \geqslant 0$ and define $\left(Z_{t}\right)_{t \geqslant 0}$ by

$$
\begin{equation*}
Z_{T_{n}+s}=\left(\Phi^{\tilde{Y}_{n}}\left(s, \tilde{X}_{n}\right), \tilde{Y}_{n}\right) \tag{3}
\end{equation*}
$$

for all $0 \leqslant s<U_{n+1}$. The memoryless property of exponential random variables make $\left(Z_{t}\right)_{t \geqslant 0}$ a continuous time càd-làg Markov process.

Let us prove that the generator of this process is indeed given by (2).
We let $P=\left(P_{t}\right)_{t \geqslant 0}$ denote the semigroup induced by $\left(Z_{t}\right)_{t \geqslant 0}$. For each $t \geqslant 0, P_{t}$ acts on bounded measurable maps $g: M \times E \rightarrow \mathbb{R}$ according to the formula

$$
P_{t} g(x, i)=\mathbb{E}\left(g\left(Z_{t}\right) \mid Z_{0}=(x, i)\right) .
$$

Let $K_{t}$ and $Q_{t}$ be the operators defined by

$$
K_{t} g(x, i)=g\left(\Phi_{t}^{i}(x), i\right)
$$

and

$$
Q_{t}=K_{t} Q .
$$

Set $Q^{0}=I d$ and let $Q^{n}$ be the random operator $Q_{U_{1}} \circ \ldots \circ Q_{U_{n}}$. It follows from (3) that

$$
\begin{equation*}
P_{t} g=\sum_{n \geqslant 0} \mathbb{E}\left(\mathbb{1}_{\left\{N_{t}=n\right\}} Q^{n} K_{t-T_{n}} g\right) \tag{4}
\end{equation*}
$$

where $N_{t}=\sum_{n \geqslant 1} \mathbb{1}_{\left\{T_{n} \leqslant t\right\}}$ is the Poisson process induced by $\left(T_{n}\right)$.

Remark 2.1. Note that $P_{t}$ can be rewritten as

$$
\begin{equation*}
P_{t}=\sum_{n \geqslant 0} \lambda^{n} e^{-\lambda t} \int_{\left\{u \in \mathbb{R}_{+}^{n}: \sum_{i=1}^{n} u_{i}<t\right\}}\left(Q_{u_{1}} \circ \ldots \circ Q_{u_{n}} K_{t-\sum u_{i}}\right) d u_{1} \ldots d u_{n} . \tag{5}
\end{equation*}
$$

Let $C_{0}=C_{0}(M \times E)$ denote the set of real valued functions $f: M \times E \rightarrow \mathbb{R}$ continuous in the first variable equipped with the uniform norm $\|\cdot\|$. We also let $C_{1} \subset C_{0}$ denote the set of $C^{0}$ functions which are $C^{1}$ in the $x$ variable. We say that $P$ is Feller provided that, for any $t \geqslant 0, P_{t}$ maps $C_{0}$ into itself and $\lim _{t \rightarrow 0}\left\|P_{t} f-f\right\|=0$.

Proposition 2.2. The semigroup $P=\left(P_{t}\right)$ is Feller. Furthermore, for all $g \in C_{1}$,

$$
\lim _{t \rightarrow 0}\left\|\frac{P_{t} g-g}{t}-L g\right\|=0
$$

Proof. By Lebesgue continuity theorem and (4), $P_{t} g \in C_{0}$ whenever $g \in C_{0}$. Moreover, setting apart the first two terms in (4) leads to

$$
\begin{equation*}
P_{t} g=e^{-\lambda t} K_{t} g+\lambda e^{-\lambda t} \int_{0}^{t} Q_{u} K_{t-u} g d u+R(g, t) \tag{6}
\end{equation*}
$$

where $|R(g, t)| \leqslant|g| \mathbb{P}\left(N_{t}>1\right)=|g|\left(1-e^{-\lambda t}(1+\lambda t)\right)$. Therefore $\lim _{t \rightarrow 0}\left\|P_{t} g-g\right\|=0$.
Let us call $A$ the infinitesimal generator of $\left(K_{t}\right)$. If $g$ is $\mathcal{C}^{1}$ in $x$, it is in the domain of $A$ and

$$
\lim _{t \rightarrow 0} \frac{K_{t} g-g}{t}=A g:(x, i) \mapsto\left\langle\nabla g^{i}(x), F^{i}(x)\right\rangle
$$

Therefore, by (6), $\frac{P_{t} g-g}{t} \underset{t \rightarrow 0}{\longrightarrow} A g-\lambda g+\lambda Q_{0} g$, and the result follows.
Similarly, we let $\tilde{P}$ denote the discrete time semigroup of $\tilde{Z}$. It is defined by the formula

$$
\begin{equation*}
\tilde{P} g(x, i)=\mathbb{E}\left(g\left(\tilde{Z}_{1}\right) \mid Z_{0}=(x, i)\right)=\int_{0}^{\infty} Q_{t} g(x, i) \lambda e^{-\lambda t} d t \tag{7}
\end{equation*}
$$

Note that

$$
\tilde{P}=\tilde{K} Q
$$

where

$$
\tilde{K} f=\int_{0}^{\infty} \lambda e^{-\lambda t} K_{t} f d t
$$

It follows from (17) that $\tilde{P}$ is Feller, meaning that it maps $C^{0}$ into itself.
Notation. Throughout the paper we may write $\mathbb{P}_{x, i}(\cdot)$ for $\mathbb{P}\left(\cdot \mid Z_{0}=(x, i)\right)$ and $\mathbb{E}_{x, i}(\cdot)$ for $\mathbb{E}\left(\cdot \mid Z_{0}=(x, i)\right)$.

## 3 Support of the law of paths

In this section we shall describe the support of the law of $\left\{X_{t}: t \geqslant 0\right\}$ in terms of a certain differential inclusion induced by $\left\{F^{i}\right\}$.

For each $x \in \mathbb{R}^{m}$, let $\operatorname{co}(F)(x) \subset \mathbb{R}^{m}$ be the compact convex set defined as

$$
\operatorname{co}(F)(x)=\left\{\sum_{i \in E} \alpha_{i} F^{i}(x): \alpha_{i} \geqslant 0, \sum_{i \in E} \alpha_{i}=1\right\}
$$

Let $C\left(\mathbb{R}, \mathbb{R}^{m}\right)$ denote the set of continuous paths $\eta: \mathbb{R} \mapsto \mathbb{R}^{m}$ equipped with the topology of uniform convergence on compact intervals. A solution to the differential inclusion

$$
\begin{equation*}
\dot{\eta} \in c o(F)(\eta) \tag{8}
\end{equation*}
$$

is an absolutely continuous function $\eta \in C\left(\mathbb{R}, \mathbb{R}^{m}\right)$ such that $\dot{\eta}(t) \in c o(F)(\eta(t))$ for almost all $t \in \mathbb{R}$. We let $S^{x} \subset C\left(\mathbb{R}, \mathbb{R}^{m}\right)$ denote the set of solutions to (8) with initial condition $x$.

Lemma 3.1. The set $S^{x}$ is a nonempty compact connected set.
Proof. Follows from standard results on differential inclusion, since the set-valued map $c o(F)$ is upper-semi continuous, bounded with nonempty compact convex images; see [1] for details.

For $I=[0, T]$ with $T>0$, or $I=\mathbb{R}^{+}$we let $\pi_{I}: C\left(\mathbb{R}, \mathbb{R}^{m}\right) \mapsto C\left(I, \mathbb{R}^{m}\right)$ denote the canonical projection defined by $\pi_{I}(\eta)(s)=\eta(s)$; and we set $S_{I}^{x}=\pi_{I}\left(S^{x}\right)$.

Theorem 3.2. Assume $X_{0}=x$. Then the support of the law of $\left\{X_{t}: t \geqslant 0\right\}$ equals $S_{\mathbb{R}^{+}}^{x}$.
Obviously, any path of $X$ is a solution of the differential inclusion (8). The proof of Theorem 3.2 is made of several steps that consist in different approximations of a solution of (8).

Let $R^{x} \subset S_{\mathbb{R}^{+}}^{x}$ denote the set of paths $\eta$ such that

$$
\dot{\eta}(t) \in \bigcup_{i \in E}\left\{F^{i}(\eta(t))\right\}
$$

for almost all $t \in \mathbb{R}^{+}$.
Lemma 3.3. The set $R^{x}$ is dense in $S_{\mathbb{R}^{+}}^{x}$.
Proof. Clearly $\overline{R^{x}} \subset S_{\mathbb{R}^{+}}^{x}$. Conversely let $\eta \in S_{\mathbb{R}^{+}}^{x}$, and $\epsilon>0$. Set

$$
G_{t}(x)=\left\{v \in \bigcup_{i \in E} F^{i}(x):\langle v-\dot{\eta}(t), x-\eta(t)\rangle<\epsilon\right\} .
$$

Since $\dot{\eta}(t) \in c o(F)(\eta(t))$ almost surely, $G_{t}(x)$ is non empty. Furthermore, $(t, x) \mapsto G_{t}(x)$ is uniformly bounded, lower semicontinuous in $x$, and measurable in $t$. Hence, using a result by Papageorgiou [21], there exists $\xi: \mathbb{R} \mapsto \mathbb{R}^{m}$ absolutely continuous such that $\xi(0)=x$ and $\dot{\xi}(t) \in G_{t}(\xi(t))$ almost surely. In particular,

$$
\frac{d}{d t}\|\xi(t)-\eta(t)\|^{2}=2\langle\dot{\xi}(t)-\dot{\eta}(t), \xi(t)-\eta(t)\rangle<2 \epsilon
$$

so that

$$
\sup _{0 \leqslant t \leqslant T}\|\xi(t)-\eta(t)\|^{2} \leqslant 2 \epsilon T
$$

Let $Q^{x} \subset S_{\mathbb{R}^{+}}^{x}$ denote the set of paths that are obtained by successive integrations of the $F^{i}$ over positive intervals. More precisely, $\eta \in Q^{x}$ if there exist points $i_{0}, i_{1}, \ldots \in E$ and numbers $0=\sigma_{0} \leqslant \sigma_{1} \leqslant \ldots \leqslant \sigma_{n} \uparrow \infty$ such that

$$
\left\{\begin{array}{l}
\eta(0)=x  \tag{9}\\
\eta\left(\sigma_{k}+s\right)=\Phi_{s}^{i_{k}}\left(\eta\left(\sigma_{k}\right)\right) \quad \text { for all } s \in\left[0, \sigma_{k+1}-\sigma_{k}\right]
\end{array}\right.
$$

Lemma 3.4. The set $Q^{x}$ is dense in $S_{\mathbb{R}^{+}}^{x}$.
Proof. By Lemma 3.3, it suffices to show that for all $\xi \in R^{x}, T>0$ and $\varepsilon>0$ there exists $\eta \in Q^{x}$ such that $\sup _{0 \leqslant t \leqslant T}\|\eta(t)-\xi(t)\| \leqslant \varepsilon$. Set $E=\{1, \ldots, d\}$, and define by induction

$$
\begin{aligned}
\Omega_{1} & =\left\{t \in[0, T]: \dot{\xi}(t)=F^{1}(\xi(t))\right\} \\
\forall i=1, \ldots, d-1, \quad \Omega_{i+1} & =\left\{t \in[0, T] \backslash\left(\Omega_{1} \cup \ldots \cup \Omega_{i}\right): \dot{\xi}(t)=F^{i+1}(\xi(t))\right\}
\end{aligned}
$$

Let $\mathcal{C}$ be the algebra consisting of finite union of intervals in $[0, T]$. Since the Borel $\sigma$-field over $[0, T]$ is generated by $\mathcal{C}$ there exists, for all $i=1, \ldots, d, I_{i} \in \mathcal{C}$ such that

$$
\operatorname{Leb}\left(\Omega_{i} \Delta I_{i}\right) \leqslant \epsilon
$$

where Leb stands for the normalized Lebesgue measure over $[0, T]$ and $A \Delta B$ is the symmetric difference of $A$ and $B: A \Delta B=(A \backslash B) \cup(B \backslash A)$.

Now, set

$$
\begin{gathered}
J_{1}=I_{1} \\
J_{i+1}=I_{i+1} \backslash\left(J_{1} \cup \ldots \cup J_{i}\right)
\end{gathered}
$$

for $i=1, \ldots, d-1$ and

$$
J_{d+1}=[0, T] \backslash\left(J_{1} \cup \ldots \cup J_{d}\right)
$$

The family $\left\{J_{i}: i=1, \ldots d+1\right\}$ forms a partition of $[0, T]$. Hence, there exists numbers $0=\sigma_{0}<\sigma_{1}<\ldots<\sigma_{N+1}=T$ and a map $i:\{0, \ldots, N\} \mapsto\{1, \ldots, d+1\}, k \mapsto i_{k}$ such that $] \sigma_{k}, \sigma_{k+1}\left[\subset J_{i_{k}}\right.$. Set $F^{d+1}=F^{1}$ and define $\eta \in Q^{x}$ by Formula (9) for $k=0, \ldots, N$. For all $\sigma_{k} \leqslant t \leqslant \sigma_{k+1}$,

$$
\eta(t)-\xi(t)=\eta\left(\sigma_{k}\right)-\xi\left(\sigma_{k}\right)+\int_{\sigma_{k}}^{t}\left(F^{i_{k}}(\eta(s))-F^{i_{k}}(\xi(s))\right) d s+\int_{\sigma_{k}}^{t}\left(F^{i_{k}}(\xi(s))-\dot{\xi}(s)\right) d s
$$

Hence, by Gronwall's lemma, we get that

$$
\|\eta(t)-\xi(t)\| \leqslant e^{K\left(\sigma_{k+1}-\sigma_{k}\right)}\left(v_{k}+m_{k}\right)
$$

where $K$ is a Lipchitz constant for all the vector fields $\left(F^{i}\right), v_{k}=\left\|\eta\left(\sigma_{k}\right)-\xi\left(\sigma_{k}\right)\right\|$ and

$$
m_{k}=L \operatorname{Leb}\left(\left[\sigma_{k}, \sigma_{k+1}\right] \backslash \Omega_{i_{k}}\right)
$$

with $\left\|F^{i}\right\| \leqslant L$ and the convention $\Omega_{d+1}=\emptyset$. It then follows that, for all $k=0, \ldots, N$ and $\sigma_{k} \leqslant t \leqslant \sigma_{k+1}$,

$$
\begin{equation*}
\|\eta(t)-\xi(t)\| \leqslant \sum_{l=0}^{k} e^{K\left(\sigma_{k+1}-\sigma_{l}\right)} m_{l} \leqslant e^{K T} \sum_{l=0}^{N} m_{l} \leqslant e^{K T} \sum_{i=1}^{d+1} \operatorname{Leb}\left(J_{i} \backslash \Omega_{i}\right) \tag{10}
\end{equation*}
$$

The lemma will follow once we prove the following claim:
Claim. $\operatorname{Leb}\left(\Omega_{i} \Delta J_{i}\right) \leqslant C \varepsilon$ for some constant $C$ depending only on $d$.
For $i \leqslant d, J_{i} \backslash \Omega_{i} \subset I_{i} \subset \Omega_{i}$ so that $\operatorname{Leb}\left(J_{i} \backslash \Omega_{i}\right) \leqslant \varepsilon$. Now,

$$
\Omega_{i} \backslash J_{i}=\Omega_{i} \cap\left[I_{i}^{c} \cup J_{1} \cup \ldots \cup J_{i-1}\right]=\Omega_{i} \backslash\left(I_{i} \cup\left(\bigcup_{l=1}^{i-1} \Omega_{i} \cap J_{l}\right)\right)
$$

For $l \neq i, \Omega_{i} \cap J_{l} \subset \Omega_{i} \cap I_{l}=\Omega_{i} \cap\left(I_{l} \backslash \Omega_{l}\right) \subset I_{l} \backslash \Omega_{l}$. Hence

$$
\operatorname{Leb}\left(\Omega_{i} \backslash J_{i}\right) \leqslant i \varepsilon \leqslant d \varepsilon
$$

We then have shown that $\operatorname{Leb}\left(\Omega_{i} \Delta J_{i}\right) \leqslant(d+1) \varepsilon$ for $i=1, \ldots, d$. It then follows that $\operatorname{Leb}\left(J_{d+1}\right) \leqslant d(d+1) \varepsilon$. This proves the claim and the lemma.

Let us complete the
Proof of Theorem 3.2. In view of Lemma 3.4 it suffices to show that for all $\varepsilon>0, T>0$ and $\eta \in Q^{x}$ the event $\sup _{0 \leqslant t \leqslant T}\left\|X_{t}-\eta(t)\right\| \leqslant \varepsilon$ has positive probability. The map $\eta$ is determined by sequences $\left(i_{n}\right),\left(\sigma_{n}\right)$ and the formula (9). Set $t_{k}=\sigma_{k}-\sigma_{k-1}$. Since we allow $t_{k}$ to be zero we may assume without loss of generality that $Q\left(\eta\left(\sigma_{k+1}\right)\right)\left(i_{k}, i_{k+1}\right)>0$ for all $k$.

Let $n$ be such that $\sigma_{n}<T \leqslant \sigma_{n+1}$ and

$$
\Sigma(n, T)=\left\{\left(s_{1}, \ldots, s_{n}\right) \in \mathbb{R}_{+}^{n}: \sum_{i} s_{i} \leqslant T\right\}
$$

For $\mathbf{s} \in \Sigma(n, T)$ set $\sigma_{k}(\mathbf{s})=s_{1}+\ldots+s_{k}$ for $k=1, \ldots, n, \sigma_{0}(\mathbf{s})=0$ and $\sigma_{n+1}(\mathbf{s})=T$. Now let $\eta_{\mathbf{s}} \in C\left([0, T], \mathbb{R}^{m}\right)$ be the function defined by (9) for $k=0, \ldots, n$ and $\sigma_{k}=\sigma_{k}(\mathbf{s})$. The mapping $\mathbf{s} \mapsto \eta_{\mathbf{s}}$ is easily seen to be continuous. Hence, there exist $\delta_{1}, \delta_{2}>0$ such that

$$
\begin{equation*}
\left\|\eta_{\mathbf{s}}-\eta\right\|_{[0, T]} \leqslant \varepsilon \tag{11}
\end{equation*}
$$

and

$$
\begin{equation*}
Q\left(\eta\left(\sigma_{k+1}(\mathbf{s})\right)\left(i_{k}, i_{k+1}\right) \geqslant \delta_{2}\right. \tag{12}
\end{equation*}
$$

whenever

$$
\max _{i=1, \ldots, n}\left|s_{i}-t_{i}\right| \leqslant \delta_{1}
$$

Let $E_{k}$ be the event defined for $k=1 \ldots n$ by

$$
E_{k}=\left\{\max _{i=1, \ldots, k}\left\|U_{i}-t_{i}\right\| \leqslant \delta_{1}, \tilde{Y}_{1}=i_{1}, \ldots, \tilde{Y}_{n}=i_{n}\right\}
$$

It follows from (11) that $\left\|X_{t}-\eta\right\|_{[0, T]} \leqslant \varepsilon$ on the event $E_{n} \cap\left\{\tau_{n+1}>T\right\}$. Now, there exists $r>0$ such that

$$
\mathbb{P}\left(U_{n+1}>T\right) \geqslant r \text { and } \mathbb{P}\left(\left|U_{k}-t_{k}\right| \leqslant \delta_{1}\right) \geqslant r .
$$

Therefore,

$$
\begin{aligned}
\mathbb{P}\left(E_{n} \cap\left\{\tau_{n+1}>T\right\}\right) & \geqslant r \mathbb{P}\left(E_{n}\right) \\
& =r \mathbb{E}\left(E_{n-1} ; \mathbb{E}\left(Q\left(\tilde{X}_{n}\right)\left(i_{n-1}, i_{n}\right) \mid \mathcal{F}_{n-1}\right) \mathbb{P}\left(\left|U_{n}-t_{n}\right| \leqslant \delta_{1}\right)\right. \\
& \geqslant \delta_{2} r^{2} \mathbb{P}\left(E_{n-1}\right) .
\end{aligned}
$$

By induction we get

$$
\mathbb{P}\left(E_{n}\right) \geqslant r\left(r \delta_{2}\right)^{n}>0
$$

This concludes the proof.

## 4 The accessible set

In order to describe the long term behavior of $\left(X_{t}\right)$ it is natural to define the accessible set of the process $\left(X_{t}\right)$ as the set of points that can be "reached from everywhere" by $\left(X_{t}\right)$. The purpose of this section is to make such a definition precise and to investigate some of its basic properties.

For all $n \in \mathbb{N}_{*}$ let $\mathbb{T}_{n}=E^{n+1} \times \mathbb{R}_{+}^{n}$. Given

$$
(\mathbf{i}, \mathbf{t})=\left(\left(i_{0}, \ldots, i_{n}\right) ;\left(t_{1}, \ldots, t_{n}\right)\right) \in \mathbb{T}_{n}
$$

and $x \in M$ we let

$$
\begin{equation*}
\boldsymbol{\Phi}_{\mathbf{t}}^{\mathbf{i}}(x)=\Phi_{t_{n}}^{i_{n-1}} \circ \ldots \circ \Phi_{t_{1}}^{i_{0}}(x) . \tag{13}
\end{equation*}
$$

The positive trajectory of $x$ is the set

$$
\gamma^{+}(x)=\left\{\boldsymbol{\Phi}_{\mathbf{t}}^{\mathbf{i}}(x):(\mathbf{i}, \mathbf{t}) \in \bigcup_{n \in \mathbb{N}_{*}} \mathbb{T}_{n}\right\} .
$$

The accessible set of $\left(X_{t}\right)$ is the (possibly empty) compact set $\Gamma \subset M$ defined as

$$
\Gamma=\bigcap_{x \in M} \overline{\gamma^{+}(x)} .
$$

Remark 4.1. The accessible set $\Gamma$ is called the set of $D$-approachable points and is denoted by $L$ in [2].

### 4.1 The accessible set and recurrence properties

The following lemmas relate the accessible set to the behavior of the processes $\tilde{Z}$ and $Z$.
Lemma 4.2. Assume that $\Gamma \neq \emptyset$. Let $p \in \Gamma$ and $U$ be a neighborhood of $p$. There exists $m \in \mathbb{N}$ and $\delta>0$ such that for all $i, j \in E$ and $x \in M$

$$
\mathbb{P}_{x, i}\left(\tilde{Z}_{m} \in U \times\{j\}\right) \geqslant \delta .
$$

In particular,

$$
\mathbb{P}_{x, i}\left(\exists m \in \mathbb{N}, \tilde{Z}_{m} \in U \times\{j\}\right)=1
$$

Similarly we have
Lemma 4.3. Assume that $\Gamma \neq \emptyset$. Let $p \in \Gamma$ and $U$ be a neighborhood of $p$. There exists $m \in \mathbb{N}, t_{1}, \ldots, t_{m}>0$ and $\delta>0$ such that for all $i, j \in E$ and $x \in M$

$$
\sum_{i=1}^{m} \mathbb{P}_{x, i}\left(Z_{t_{i}} \in U \times\{j\}\right) \geqslant \delta .
$$

Since $\Gamma$ is positively invariant under each flow $\Phi^{i}$ we deduce from Lemma 4.3 the following result.

Corollary 4.4. Assume $\Gamma$ has nonempty interior. Then

$$
\mathbb{P}_{x, i}\left(\exists t_{0} \geqslant 0, \forall t \geqslant t_{0}, Z_{t} \in \Gamma \times E\right)=1 .
$$

In order to provide the proof of Lemmas 4.2 and 4.3 we firstly establish two technical lemmas. For all $x \in M,(\mathbf{i}, \mathbf{t}) \in \mathbb{T}_{n}$ let

$$
p(x, \mathbf{i}, \mathbf{t})=Q\left(x_{1}\right)\left(i_{0}, i_{1}\right) \times Q\left(x_{2}\right)\left(i_{1}, i_{2}\right) \times \ldots \times Q\left(x_{n}\right)\left(i_{n-1}, i_{n}\right)
$$

where $x_{0}=x$ and $x_{k}=\Phi_{t_{k}}^{i_{k-1}}\left(x_{k-1}\right)$ for $k=1, \ldots, n$.
Definition 4.5. The sequence $(\boldsymbol{i}, \boldsymbol{t}) \in \mathbb{T}_{n}$ is said adapted to $x \in M$ if $p(x, \boldsymbol{i}, \boldsymbol{t})>0$.
For $i, j \in E$, let $\mathbb{T}_{n}^{i j}$ denote the subset of $\mathbb{T}_{n}$ consisting of sequences $(\mathbf{i}, \mathbf{t})$ for which $i_{0}=i$ and $i_{n}=j$. The next lemma show that the positive trajectory of $x$ can be obtained with adapted sequences.

Lemma 4.6. There exists $r \in \mathbb{N}_{*}$ such that for all $x \in M, i, j \in E$

$$
\gamma^{+}(x)=\left\{\boldsymbol{\Phi}_{t}^{i}(x):(\boldsymbol{i}, \boldsymbol{t}) \in \bigcup_{n \in \mathbb{N}_{*}} \mathbb{T}_{n r}^{i j} \text { with }(\boldsymbol{i}, \boldsymbol{t}) \text { adapted to } x\right\}
$$

Proof. By ergodicity of $Q(x)$ there exists $r \in \mathbb{N}$ such that $P^{r}(x)$ has positive entries. By continuity of $P$ and compactness we can choose such an $r$ to be independent of $x$. Now, given $(\mathbf{i}, \mathbf{t}) \in \mathbb{T}_{n}$ we can construct a new sequence in $\mathbb{T}_{r(n+1)}^{i j}$ as follows. Insert before $i_{0}$ a sequence of the form $\left(i, j^{1}, \ldots, j^{r-1}\right)$ with $Q(x)\left(i, j^{1}\right) \times \ldots \times Q(x)\left(j^{r-2}, j^{r}\right) \times$ $Q(x)\left(j^{r-1}, i_{0}\right)>0$ and insert before $t_{1}$ a sequence $(0, \ldots, 0)$ of length $r$. Replace $i_{n}$ by $j$ and for all $0 \leqslant k \leqslant n-1$ insert after $i_{k}$ a sequence $\left(j_{k}^{1}, \ldots, j_{k}^{r-1}\right)$ with $Q\left(x_{k}\right)\left(i_{k}, j_{k}^{1}\right) \times$ $\ldots \times Q\left(x_{k}\right)\left(j_{k}^{r-2}, j_{k}^{r-1}\right) \times Q\left(x_{k}\right)\left(j^{r-1}, i_{k+1}\right)>0$ and insert after $t_{k}$ a sequence $(0, \ldots, 0)$ of length $r$. Such a procedure doesn't affect the value of $\boldsymbol{\Phi}_{\mathbf{t}}^{\mathbf{i}}(x)$. Hence the result.

Lemma 4.7. Assume $\Gamma \neq \emptyset$. Let $p \in \Gamma, U$ be a neighborhood of $p$ and $i, j \in E$. There exist $m \in \mathbb{N}^{*}, \epsilon, \eta>0$, finite sequences $\left(\boldsymbol{i}^{1}, \boldsymbol{t}^{1}\right) \ldots,\left(\boldsymbol{i}^{N}, \boldsymbol{t}^{N}\right) \in \mathbb{T}_{m}^{i j}$ and a open covering $O^{1}, \ldots, O^{N}$ of $M$ (i.e. $M=O^{1} \cup \ldots \cup O^{N}$ ) such that for all $x \in M$ and $\tau \in \mathbb{R}_{+}^{m}$ :

$$
x \in O^{k} \text { and }\left\|\tau-t^{k}\right\| \leqslant \epsilon \Rightarrow \boldsymbol{\Phi}_{\tau}^{i^{k}}(x) \in U \text { and } p\left(x, \boldsymbol{i}^{k}, \tau\right) \geqslant \eta
$$

Furthermore, $m, \epsilon$ and $\eta$ are independent of $i, j \in E$.
Proof. Let $V$ be a neighborhood of $p$ with closure $\bar{V} \subset U$. Define open sets

$$
\begin{aligned}
O(\mathbf{i}, \mathbf{t}, \eta) & =\left\{x \in M: \mathbf{\Phi}_{\mathbf{t}}^{\mathbf{i}}(x) \in V \text { and } p(x, \mathbf{i}, \mathbf{t})>\eta\right\} \\
O(\mathbf{i}, \mathbf{t}) & =\bigcup_{\eta>0} O(\mathbf{i}, \mathbf{t}, \eta)
\end{aligned}
$$

and

$$
O_{i j}^{n}=\bigcup O(\mathbf{i}, \mathbf{t})
$$

where the union is taken over all $(\mathbf{i}, \mathbf{t}) \in \mathbb{T}_{n r}^{i j}$ with $r$ like in Lemma 4.6, Now $O_{i j}^{n+1} \supset O_{i j}^{n}$ (the proof is similar to the proof of Lemma 4.6) and, by Lemma 4.6, $M=\bigcup_{n} O_{i j}^{n}$. It then follows, by compactness of $M$ that $M=O_{i j}^{n(i, j)}$ for some integer $n(i, j)$ and also

$$
M=O_{i j}^{n}
$$

for $n=\max n(i, j)$. By compactness again $M$ can be rewritten as

$$
M=O_{i j}^{n}=\bigcup_{l=1}^{N} O^{l}
$$

where $O^{k}=O\left(\mathbf{i}^{k}, \mathbf{t}^{k}, \eta^{k}\right)$ for some $\eta^{k}>0$ and $\left(\mathbf{i}^{k}, \mathbf{t}^{k}\right) \in \mathbb{T}_{n r}^{i j}$. The results follows with $m=n r$ and $0<\eta<\min _{k=1, \ldots, N} \eta^{k}$ and $\epsilon$ sufficiently small.

Proof of Lemma 4.2. We use the notation of Lemma 4.7. Suppose $x \in O^{k}$ and to shorten notation write $\mathbf{t}$ for $\mathbf{t}^{k}$ and $\mathbf{i}$ for $\mathbf{i}^{k}$. By Lemma 4.7, if

$$
\left\|\left(U_{1}, \ldots, U_{m}\right)-\mathbf{t}\right\| \leqslant \epsilon \text { and }\left(\tilde{Y}_{0}, \tilde{Y}_{1}, \ldots, \tilde{Y}_{m}\right)=\mathbf{i}
$$

then $\tilde{X}_{m} \in U$ and

$$
\begin{equation*}
Q\left(\tilde{X}_{1}\right)\left(i, i_{1}\right) \times \ldots \times Q\left(\tilde{X}_{m}\right)\left(i_{m-1}, j\right) \geqslant \eta \tag{14}
\end{equation*}
$$

Therefore

$$
\begin{aligned}
\mathbb{P}\left(\tilde{X}_{m} \in U, \tilde{Y}_{m}=j \mid \tilde{X}_{0}=x, \tilde{Y}_{0}=i\right) & \geqslant \mathbb{P}\left(\left\|\left(U_{1}, \ldots, U_{m}\right)-\mathbf{t}\right\| \leqslant \epsilon,\left(\tilde{Y}_{0}, \tilde{Y}_{1}, \ldots, \tilde{Y}_{m}\right)=\mathbf{i}\right) \\
& =\mathbb{P}\left(\bigcap_{j=1}^{m}\left\{\left|U_{j}-t_{j}\right| \leqslant \epsilon, \tilde{Y}_{j}=i_{j}\right\}\right)
\end{aligned}
$$

Now

$$
\mathbb{P}\left(\left|U_{m}-t_{m}^{k}\right| \leqslant \epsilon, \tilde{Y}_{m}=j \mid \mathcal{F}_{m-1}\right)=Q\left(\tilde{X}_{m}\right)\left(\tilde{Y}_{m-1}, j\right) \mathbb{P}\left(\left|U_{m}-t_{m}\right| \leqslant \varepsilon\right)
$$

on the event

$$
\bigcap_{j=1}^{m-1}\left\{\left|U_{j}-t_{j}\right| \leqslant \epsilon \text { and } \tilde{Y}_{j}=i_{j}\right\} .
$$

Hence, using (14),

$$
\begin{aligned}
\mathbb{P}\left(\bigcap_{j=1}^{m}\left\{\left|U_{j}-t_{j}\right| \leqslant \epsilon \text { and } \tilde{Y}_{j}=i_{j}\right\}\right) & \geqslant \eta \mathbb{P}\left(\bigcap_{j=1}^{m}\left\{\left|U_{j}-t_{j}\right| \leqslant \epsilon\right)\right. \\
& =\eta e^{-\sum_{j=1}^{m} t_{j}}\left(e^{\varepsilon}-e^{-\varepsilon}\right)
\end{aligned}
$$

The proof of Lemma 4.3 is similar. Details are left to the reader.

### 4.2 Topological properties of the accessible set

The differential inclusion (8) induces a set-valued dynamical system $\Psi=\left\{\Psi_{t}\right\}$ defined by

$$
\Psi_{t}(x)=\Psi(t, x)=\left\{\eta(t): \eta \in S^{x}\right\}
$$

enjoying the following properties
(i) $\Psi_{0}(x)=\{x\}$,
(ii) $\Psi_{t+s}(x)=\Psi_{t}\left(\Psi_{s}(x)\right)$ for all $t, s \geqslant 0$,
(iii) $y \in \Psi_{t}(x) \Rightarrow x \in \Psi_{-t}(y)$.

For subsets $I \subset \mathbb{R}$ and $A \subset \mathbb{R}^{m}$ we set

$$
\Psi(I, A)=\bigcup_{(t, x) \in I \times A} \Psi_{t}(x)
$$

A set $A \subset \mathbb{R}^{m}$ is called strongly positively invariant under $\Psi$ if $\Psi_{t}(A) \subset A$ for $t \geqslant 0$. It is called invariant if for all $x \in A$ there exists $\eta \in S^{x}$ such that $\eta(\mathbb{R}) \subset A$. Given $x \in \mathbb{R}^{m}$, the limit set of $x$ under $\Psi$ is defined as

$$
\omega_{\Psi}(x)=\bigcap_{t \geqslant 0} \overline{\Psi_{[t, \infty[ }(x)}
$$

Lemma 4.8. The set $\omega_{\Psi}(x)$ is compact connected invariant and strongly positively invariant.

Proof. It is not hard to deduce the first three properties from Lemma 3.1. For the last one, let $p \in \omega_{\Psi}(x), s>0$ and $q \in \Psi_{s}(p)$. By Lemma 3.4, for all $\varepsilon>0$ there exists $n \in \mathbb{N}$ and $(\mathbf{i}, \mathbf{t}) \in \mathbb{T}_{n}$ such that $d\left(\boldsymbol{\Phi}_{\mathbf{t}}^{\mathbf{i}}(p), q\right)<\varepsilon$. Continuity of $\Phi_{\mathbf{t}}^{\mathbf{i}}$ makes the set

$$
W=\left\{z \in M: d\left(\boldsymbol{\Phi}_{\mathbf{t}}^{\mathbf{i}}(z), q\right)<\varepsilon\right\}
$$

an open neighborhood of $p$. Hence $W \cap \Psi_{[t, \infty[ }(x) \neq \emptyset$ for all $t>0$. This proves that $d\left(q, \Psi_{[t, \infty[ }(x)\right)<\varepsilon$ and since $\varepsilon$ is arbitrary $q \in \overline{\Psi_{[t, \infty[ }(x)}$.

Remark 4.9. For a general differential inclusion with an upper semi-continuous bounded right hand side with compact convex values, the omega limit set of a point is not (in general) strongly positively invariant, see e.g. [4].

Proposition 4.10. The set $\Gamma$ enjoys the following properties
(i) $\Gamma=\bigcap_{x \in M} \omega_{\Psi}(x)$,
(ii) $\Gamma=\omega_{\Psi}(p)$ for all $p \in \Gamma$,
(iii) $\Gamma$ is compact, connected, strongly positively invariant and invariant under $\Psi$.
(iv) Either $\Gamma$ has empty interior or its interior is dense in $\Gamma$.

Proof. (i)]: Let $x \in M$ and $y \in \Psi_{t}(x)$. Then $\gamma^{+}(y) \subset \Psi_{[t, \infty]}(x)$. Hence $\Gamma \subset \overline{\Psi_{[t, \infty[ }(x)}$ for all $x$. This proves that $\Gamma \subset \bigcap_{x \in M} \omega_{\Psi}(x)$. Conversely let $p \in \bigcap_{x \in M} \omega_{\Psi}(x)$. Then, for all $t>0$ and $x \in M, p \in \overline{\Psi_{[t, \infty[ }(x)} \subset \overline{\gamma^{+}(x)}$ where the latter inclusion follows from Lemma 3.4. This proves the converse inclusion.
(ii): By Lemma [3.4, $\omega_{\Psi}(p) \subset \Gamma$. The converse inequality follows from (i).
(iii) : This follows from (ii) and Lemma 4.8 ,
(iv) : Suppose $\operatorname{int}(\Gamma) \neq \emptyset$. Then there exists an open set $U \subset \Gamma$ and $\bigcup_{\mathbf{t}, \mathbf{i}} \boldsymbol{\Phi}_{\mathbf{t}}^{\mathbf{i}}(U)$ is an open subset of $\Gamma$ dense in $\Gamma$.

An equilibrium $p$ for the flow $\Phi^{1}$ is called an attracting equilibrium if there exists a neighborhood $U$ of $p$ such that

$$
\lim _{t \rightarrow \infty}\left\|\Phi_{t}^{1}(x)-p\right\|=0
$$

uniformly in $x \in U$. In this case, the basin of attraction of $p$ is the open set

$$
B(p)=\left\{x \in \mathbb{R}^{m}: \lim _{t \rightarrow \infty}\left\|\Phi_{t}^{1}(x)-p\right\|=0\right\}
$$

Proposition 4.11. Suppose the flow $\Phi^{1}$ has an attracting equilibrium $p$ with basin of attraction $B(p)$; and that for all $x \in M \backslash B(p), \gamma^{+}(x) \cap B(p) \neq \emptyset$. Then
(i) $\Gamma=\overline{\gamma^{+}(p)}$,
(ii) If furthermore $\Gamma \subset B(p)$. Then $\Gamma$ is contractile. In particular, it is simply connected.

Proof. The proof of (i) is left to the reader. To prove (ii), let $h:[0,1] \times \Gamma \rightarrow \Gamma$ be defined by

$$
h(t, x)= \begin{cases}\Phi^{1}(-\log (1-t), x) & \text { if } t<1 \\ p & \text { if } t=1\end{cases}
$$

It is easily seen that $h$ is continuous. Hence the result.

### 4.3 The two-dimensional case

We assume here that $M$ is a subset of $\mathbb{R}^{2}$ and that one of the vector fields, say $F^{1}$, has a globally attracting equilibrium $p$. Hence, by Proposition 4.11, $\Gamma \neq \emptyset$.

Proposition 4.12. Assume that the eigenvalues of $D F^{1}(p)$ are non real with negative real parts and that $F^{2}(p) \neq 0$. Then $p$ lies in the interior of $\Gamma$.

Proof. The following drawing shows that this is the case: from the equilibrium $p$, one can follow the flow of $F^{2}$ and reach $x$, then follow the spiral, and swith back to $F^{2}$ to reach any point in the shaded region.


Proposition 4.13. If the PDMP is driven by only two planar contractant real analytic vector fields without common trajectories, then the boundary of $\Gamma$ is a finite union of pieces of trajectories.

Proof. Let $V=F^{0}$ and $W=F^{1}$ denote the two vector fields.
First let us remark that in this analytic case the trajectories are real analytic curves. So is the curve defined by $\operatorname{det}(V, W)=0$. In particular all these curves locally intersect one each other finitely many times.

Let $x$ be a point belonging to the boundary $\partial \Gamma$ of $\Gamma$. Suppose first that $x$ does not belong to $\operatorname{det}(V, W)=0$. Then the boundary of $\Gamma$ in the neighbourhood of $x$ has one of the three following forms.


To get this we just have to remark that at $x$ the vector fields of our PDMP cannot point to the exterior of $\Gamma$ and that it must be possible to reach $x$ from the interior of $\Gamma \ldots$

In particular for every $\varepsilon>0$ the set $\partial \Gamma \cap|\operatorname{det}(V, W)|>\varepsilon$ is composed by a finite number of pieces of trajectories of both vector fields.

Let us now consider the case of a point $x$ belonging to $\operatorname{det}(V, W)=0$ but at which the two vector fields are transverse to the curve $\operatorname{det}(V, W)=0$. By conjugation it is possible to suppose that $V$ is an horizontal vector field. At such a point the different possibilities are the following. We distinguish two cases : at $x, V$ and $W$ are colinear, they may have the same sense or opposite sense.


In these cases also the boundary of $\Gamma$ is simple. In the analytic case the preceding descriptions are valid at points $x$ where $V$ and $W$ do not vanish.

Then we just have to say what happens if $x$ is a stationary point for one of the two flows (since there are no common trajectories, a point cannot be stationary for both flows). For example let us suppose that $V(x)=0$. Remark that the derivative of $V$ at $x$ does not have non real eigenvalues (because a stationary point with non real eigenvalues is in the interior of $\Gamma$ ). We can assume that at $x, W$ is an horizontal flow. Take a little circle $\mathcal{C}$ around $x$ and a point $y$ in $\mathcal{C} \cap \partial \Gamma$. Suppose that $V$ is not horizontal at $y$. Now take $z$ another point. If $z$ belongs to $\Gamma$ the $V$-trajectory starting at $z$ stays in $\Gamma$. But if $z$ is in $\Gamma$ and sufficiently near $x$, the only way to reach $z$ by one of the two flows is to follow a $V$-trajectory (otherwise $x$ would not be on the boundary $\partial \Gamma$ ). It means that if $z$ is not on the $V$-trajectory of $y$ then $z$ is not on the boundary $\partial \Gamma$.


This implies that the $V$ trajectory of $y$ is a part of $\partial \Gamma$. As a conclusion in a neighbourhood of $x$ the boundary $\partial \Gamma$ is either a piece of a $W$-trajectory, the union of a piece of a $W$ trajectory and a piece of a $V$-trajectory, or the union of two pieces of $V$-trajectories.

## 5 Invariant Probabilities

Let $\mathcal{M}(M \times E)$ (respectively $\mathcal{M}^{+}(M \times E)$ and $\mathcal{P}(M \times E)$ ) denote the set of signed (respectively positive, and probability) measures on $M \times E$. For $\mu \in \mathcal{M}(M \times E)$ and $f \in L^{1}(\mu)$ we write $\mu f$ for $\int f d \mu$. Recall that we let $C_{0}$ denote $C_{0}(M \times E)$. Given a bounded operator $K: C_{0} \mapsto C_{0}$ and $\mu \in \mathcal{M}(M \times E)$ we let $\mu K \in \mathcal{M}(M \times E)$ denote the measure defined by duality :

$$
\forall g \in C_{0}, \quad(\mu K) g=\mu(K g) .
$$

The mappings $\mu \mapsto \mu P_{t}, \mu \tilde{P}$ preserve the set $\mathcal{M}^{+}(M \times E)$ and $\mathcal{P}(M \times E)$.
For $\mu \in \mathcal{M}(M \times E)$ we let $\mu_{i}$ denote the marginal measure on $M$ given by $\mu_{i}(A)=$ $\mu(A \times\{i\})$.

Recall that a probability measure $\mu$ is called invariant for $\left(P_{t}\right)$ provided

$$
\forall t \geqslant 0, \quad \mu P_{t}=\mu .
$$

It is called invariant for $\tilde{P}$ if $\mu \tilde{P}=\mu$.
We let $\mathcal{P}_{\text {inv }}$ denote the set of invariant probabilities for $\left(P_{t}\right)$ and $\tilde{\mathcal{P}}_{\text {inv }}$ the set of invariant probabilities for $\tilde{P}$.

For $n \in \mathbb{N}$ and $t>0$ we let $\tilde{\Pi}_{n}$ and $\Pi_{t}$ the (random) occupation measures defined by

$$
\tilde{\Pi}_{n}=\frac{1}{n} \sum_{k=1}^{n} \delta_{\tilde{Z}_{k}} \quad \text { and } \quad \Pi_{t}=\frac{1}{t} \int_{0}^{t} \delta_{Z_{t}} .
$$

By standard results for Feller chains on a compact space (see e.g. [12]), one has
Proposition 5.1. The set $\tilde{\mathcal{P}}_{\text {inv }}$ is nonempty, compact (for the weak-ᄎ topology) and convex. Furthermore, with probability one every limit point of $\left(\Pi_{n}\right)$ lies in $\tilde{\mathcal{P}}_{\text {inv }}$.

Proposition 5.2. The mapping $\mu \mapsto \mu \tilde{K}$ maps $\tilde{\mathcal{P}}_{\text {inv }}$ homeomorphically onto $\mathcal{P}_{\text {inv }}$ and extremal points of $\tilde{\mathcal{P}}_{\text {inv }}$ (i.e. ergodic probabilities for $\tilde{P}$ ) onto extremal points of $\mathcal{P}_{\text {inv }}$ (ergodic probabilities for $P$ ).

The inverse homeomorphism is the map $\mu \mapsto \mu Q$ restricted to $\mathcal{P}_{\text {inv }}$.
Proof. For all $f \in C_{1}$, integrating by part $\int_{0}^{\infty} \frac{d K_{t} f}{d t} e^{-\lambda t} d t$ and using the identities $\frac{d K_{t f}}{d t}=$ $A K_{t} f=K_{t} A f$ leads to

$$
\begin{equation*}
\tilde{K}(\lambda I-A) f=\lambda f=(\lambda I-A) \tilde{K} f \tag{15}
\end{equation*}
$$

Let $\mu \in \mathcal{P}(M \times E)$. Then, using (15) and the form of $L$ gives

$$
\begin{align*}
\mu \tilde{K} L f & =\mu \tilde{K}(A-\lambda I) f+\lambda \mu \tilde{K} Q f=\lambda(-\mu f+\mu \tilde{P} f)  \tag{16}\\
\mu L(\tilde{K} f) & =\mu(A-\lambda I) \tilde{K} f+\lambda \mu Q \tilde{K} f=\lambda(-\mu f+\mu Q \tilde{K} f) . \tag{17}
\end{align*}
$$

If $\mu \in \tilde{\mathcal{P}}_{\text {inv }}$, (16) implies $(\mu \tilde{K}) L f=0$ for all $f \in C_{1}$ and since $C_{1}$ is dense in $C_{0}$ this proves that $\mu \tilde{K} \in \tilde{\mathcal{P}}_{\text {inv }}$. Similarly, if $\mu \in \mathcal{P}_{\text {inv }}$, (17) implies $\mu=\mu Q \tilde{K}$. Hence $(\mu Q)=(\mu Q) \tilde{K} Q=$ $(\mu Q) \tilde{P}$ proving that $\mu Q \in \tilde{\mathcal{P}}_{\text {inv }}$. Furthermore the identity $\mu=\mu Q \tilde{K}$ for all $\mu \in \mathcal{P}_{\text {inv }}$ shows that the maps $\mu \mapsto \mu \tilde{K}$ and $\mu \mapsto \mu Q$ are inverse homeomorphisms.

Lemma 5.3. Let $\mu \in \tilde{\mathcal{P}}_{\text {inv }}$. Then $\mu$ and $\mu \tilde{K}$ have the same support.

Proof. Let $(x, i) \in \operatorname{supp}(\mu)$ and let $U$ be a neighborhood of $x$. Then for $t_{0}>0$ small enough and $0 \leqslant t \leqslant t_{0} \Phi_{-t}^{i}(U)$ is also neighborhood of $x$. Thus

$$
(\mu \tilde{K})_{i}(U)=\int \lambda e^{-\lambda t} \mu_{i}\left(\Phi_{-t}^{i}(U)\right) d t \geqslant \lambda \int_{0}^{t_{0}} e^{-\lambda t} \mu_{i}\left(\Phi_{-t}^{i}(U)\right) d t>0
$$

This proves that $\operatorname{supp}(\mu) \subset \operatorname{supp}(\mu \tilde{K})$. Conversely, let $\nu=\mu \tilde{K}(x, i) \in \operatorname{supp}(\nu)$ and let $U$ be a neighborhood of $x$. Then

$$
\mu_{i}(U)=(\nu Q)_{i}(U)=\sum_{j} \int_{U} Q_{j i}(x) \nu_{j}(d x) \geqslant \int_{U} Q_{i i}(x) \nu_{i}(d x)>0
$$

As a consequence, $\operatorname{supp}(\mu) \supset \operatorname{supp}(\mu \tilde{K})$.
The following lemma is useful to compare empirical measures of $\tilde{Z}$ and $Z$.
Lemma 5.4. Let $f: M \times E \mapsto \mathbb{R}$ be a bounded measurable function. Then

$$
\lim _{t \rightarrow \infty} \Pi_{t} f-\tilde{\Pi}_{N_{t}} \tilde{K} f=0 \quad \text { and } \quad \lim _{n \rightarrow \infty} \Pi_{T_{n}} f-\tilde{\Pi}_{n} \tilde{K} f=0
$$

with probability one.
Proof. Write

$$
\Pi_{t} f=\frac{N_{t}}{t}\left(\frac{1}{N_{t}} \sum_{i=0}^{N_{t}-1} \int_{T_{i}}^{T_{i+1}} f\left(Z_{s}\right) d s+r_{t}\right)
$$

where $\left\|r_{t}\right\| \leqslant\|f\| \frac{U_{N_{t}+1}}{N_{t}}$.
Since $\lim _{t \rightarrow \infty} \frac{N_{t}}{t}=1$ almost surely and $\mathbb{P}\left(U_{n} / n \geqslant \varepsilon\right)=e^{-\lambda n \varepsilon}, r_{t} \xrightarrow[t \rightarrow \infty]{a . s .} 0$, so

$$
\Pi_{t} f-\frac{1}{N_{t}} \sum_{i=0}^{N_{t}-1} \int_{T_{i}}^{T_{i+1}} f\left(Z_{s}\right) d s \xrightarrow[t \rightarrow \infty]{\text { a.s. }} 0
$$

Now, note that

$$
\int_{T_{i}}^{T_{i+1}} f\left(Z_{s}\right) d s=\int_{0}^{U_{i+1}} f\left(\phi_{s}^{\tilde{Y}^{i}}\left(\tilde{X}_{i}\right), \tilde{Y}_{i}\right) d s
$$

Therefore

$$
M_{n}=\sum_{i=0}^{n-1}\left(\int_{T_{i}}^{T_{i+1}} f\left(Z_{s}\right) d s-\tilde{K} f\left(\tilde{X}_{i}, \tilde{Y}_{i}\right)\right)
$$

is a martingale with increments bounded in $L^{2}: \mathbb{E}\left(M_{n+1}-M_{n}\right)^{2} \leqslant 2\|f\|^{2} / \lambda^{2}$. Therefore, by the strong law of large numbers for Martingales,

$$
\lim _{n \rightarrow \infty} \frac{M_{n}}{n}=0
$$

Corollary 5.5. The set $\mathcal{P}_{i n v}$ is nonempty compact (for the weak-ᄎ topology) and convex. Furthermore, with probability one, every limit point of $\Pi_{t}$ lies in $\mathcal{P}_{\text {inv }}$.

Proof. This follows from Proposition 5.2 and Lemma 5.4.
The chain $\tilde{Z}$ (respectively $(Z)$ ) is called stable if it has a unique invariant probability.
Proposition 5.6. The chain $\left(Z_{t}\right)$ is stable if and only if $(\tilde{Z})$ is stable.
Proof. This follows from Proposition 5.2,

### 5.1 Support of invariant probabilities

Given $\mu \in \mathcal{P}(M \times E)$ we let $\operatorname{supp}(\mu)$ denote its support. The following proposition relates $\Gamma$ to $\operatorname{supp}(\mu)$ when $\mu$ is invariant for $\tilde{Z}$ or $Z$. We state and prove the result for $\tilde{Z}$ and rely on Lemma 5.3 for $Z$.

Proposition 5.7. (i) If $\Gamma \neq \emptyset$ then $\Gamma \times E \subset \operatorname{supp}(\mu)$ for all $\mu \in \tilde{\mathcal{P}}_{\text {inv }}$ (respectively $\mathcal{P}_{\text {inv }}$ ) and there exists $\mu \in \tilde{\mathcal{P}}_{\text {inv }}$ (resp. $\mathcal{P}_{\text {inv }}$ ) such that $\operatorname{supp}(\mu)=\Gamma \times E$.
(ii) If $\Gamma$ has nonempty interior, then $\Gamma \times E=\operatorname{supp}(\mu)$ for all $\mu \in \tilde{\mathcal{P}}_{\text {inv }}$ (resp. $\left.\tilde{\mathcal{P}}_{\text {inv }}\right)$.
(iii) Suppose ( $\tilde{Z})$ ) (resp. Z) is stable with invariant probability $\pi$. Then supp $(\pi)=\Gamma \times E$.

Proof. (i) follows from Lemma 4.2. Also, since $\Gamma$ is strongly positively invariant, there are invariant measures supported by $\Gamma \times E$. (ii) follows from (i) and Corollary 4.4. To prove (iii), let $(p, i) \in \operatorname{supp}(\pi)$. Let $U, V$ be open neighborhoods of $p$ with $\bar{U} \subset V$ compact. Let $0 \leqslant f \leqslant 1$ be a continuous function which is 1 on $U$ and 0 outside $V$ and let $\tilde{f}(x, j)=f(x) \delta_{j, i}$. Suppose $Z_{0}=(x, j)$. Then with probability one

$$
\begin{aligned}
& \liminf _{n \rightarrow \infty} \frac{1}{n} \sharp\left\{1 \leqslant k \leqslant n: \tilde{Z}_{k} \in V \times\{i\}\right\} \\
\geqslant & \lim _{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^{n} f\left(\tilde{Z}_{k}\right)=\int f \pi \\
\geqslant & \pi(U \times\{i\})>0 .
\end{aligned}
$$

Hence $\left(\tilde{Z}_{n}\right)$ visits infinitely often $U \times\{i\}$. In particular, $p \in \gamma^{+}(x)$. This proves that $\operatorname{supp}(\pi) \subset \Gamma \times E$. The converse statement follows from (i).

Remark 5.8. The example given in Section 9 shows that the inclusion $\Gamma \times E \subset \operatorname{supp}(\mu)$ may be strict when $\Gamma$ has empty interior. On the other hand, the condition that $\Gamma$ has non empty interior is not sufficient to ensure uniqueness of the invariant probability since there exist smooth minimal flows that are not uniquely ergodic. An example of such a flow can be constructed on a 3 -manifold by taking the suspension of an analytic minimal non uniquely ergodic diffeomorphism of the torus constructed by Furstenberg in [14] (see also [19]). As shown in [2] (see also Section [6) a sufficient condition to ensure uniqueness of the invariant probability is that the vector fields verify a Hörmander bracket property at some point in $\Gamma$.

### 5.2 Law of pure types

Assume that $M$ is an embedded submanifold of $\mathbb{R}^{m}$ possibly with boundary. We let $\lambda_{M}$ denote the normalized Riemannian measure on $M$ and $\lambda_{M \times E}=\lambda_{M} \otimes \frac{1}{|E|} \sum_{i \in E} \delta_{i}$.

Proposition 5.9. Let $\mu \in \tilde{\mathcal{P}}_{\text {inv }}$ (respectively $\mathcal{P}_{\text {inv }}$ ) and let $\mu=\mu_{a c}+\mu_{s}$ be the Lebesgue decomposition of $\mu$ with $\mu_{a c}$ the absolutely continuous (with respect to $\lambda_{M \times E}$ ) measure and $\mu_{s}$ the singular (with respect to $\lambda_{M \times E}$ ) measure. Then both $\mu_{a c}$ and $\mu_{s}$ are in $\tilde{\mathcal{P}}_{\text {inv }}$ (respectively $\mathcal{P}_{\text {inv }}$. In particular, if $\mu$ is ergodic, then $\mu$ is either absolutely continuous or singular.

Proof. The key point is that $\tilde{K}$ and $Q$, hence $\tilde{P}=\tilde{K} Q$, map absolutely continuous measures into absolutely continuous measures. For $\mu \in \tilde{\mathcal{P}}_{\text {inv }}$ the result now follows from the following simple Lemma 5.10 applied to $\tilde{P}$.

Lemma 5.10. Let $(\Omega, \mathcal{A}, \lambda)$ be a measure space with $\lambda$ a probability. Let $\mathcal{M}$ (respectively $\mathcal{M}^{+}, \mathcal{P}, \mathcal{M}_{a c}$ ) denote the set of signed (positive, probability, absolutely continuous) measures on $\Omega$. Let $K: \mathcal{M} \mapsto \mathcal{M}, \mu \mapsto \mu K$ be a linear map that maps each of the preceding sets into itself. Then if $\mu \in \mathcal{P}$ is a fixed point for $K$ with Lebesgue decomposition $\mu=\mu_{a c}+\mu_{s}$ both $\mu_{a c}$ and $\mu_{s}$ are fixed point for $K$.

Proof. Write $\mu K=\mu_{a c} K+\mu_{s} K=\mu_{a c} K+\nu_{a c}+\nu_{s}$ with $\mu_{s} K=\nu_{a c}+\nu_{s}$ the Lebesgue decomposition of $\mu_{s} K$. Then, by uniqueness of the decomposition, $\mu_{a c}=\mu_{a c} K+\nu_{a c}$. Thus, $\mu_{a c} \geqslant \mu_{a c} K$. Now either $\mu_{a c}=0$ and there is nothing to prove or, we can normalize by $\mu_{a c}(\Omega)$ and we get that $\mu_{a c}=\mu_{a c} K$.

## 6 Regularity of the law of the process

For simplicity we shall assume here that $M$ is an $m$-dimensional submanifold of $\mathbb{R}^{m}$ (typically a closed ball). However the results here carry out to the more general situation where $M$ is any embedded submanifold of $\mathbb{R}^{m}$.

### 6.1 The results

Let us fix some additional notation. If $u$ is a time, $F$ a vector field, $\Phi$ the flow of $F$, and $G$ another vector field, we can pull back $G$ by $\Phi_{u}$ by defining:

$$
\left.\left(\Phi_{u}^{\star} G\right)_{x}=\left(D\left(\Phi_{u}\right)\right)_{x}\right)^{-1} \cdot G_{\Phi_{u}(x)} .
$$

Similarly we define the push-forward: $\left(\Phi_{u, \star} G\right)_{x}=\left(D\left(\Phi_{u}\right)\right)_{\Phi_{-u}(x)} \cdot G_{\Phi_{-u}(x)}$. If $F$ is the vector field $F^{i}$, we write $\boldsymbol{\Phi}_{u}^{i, \star}$ and $\boldsymbol{\Phi}_{u, \star}^{i}$ the pull-back and push-forward.

Recall that, if $\mathbf{i}$ is a sequence of indices $\mathbf{i}=\left(i_{1}, \ldots i_{K}\right)$ and $\mathbf{u}$ is a sequence of times $\mathbf{u}=\left(u_{1}, \ldots u_{K}\right), \boldsymbol{\Phi}_{\mathbf{u}}^{\mathbf{i}}: M \rightarrow M$ is the composite map defined by (13). We introduce a bit of additional notation.

- If $\mathbf{i}$ and $\mathbf{j}$ are two sequences of indices, we write $\mathbf{i} \cdot \mathbf{j}$ the concatenation of $\mathbf{i}$ and $\mathbf{j}$ (and use the same notation for the $\mathbf{u}$ ).
- For any sequence $\mathbf{i}$, we let $\mathbf{i}_{k}$ be the first $k$ elements of $\mathbf{i}$, and define $\mathbf{u}_{k}$ in the same way.

We can push vector fields forward and pull them back by the composite flows $\boldsymbol{\Phi}_{\mathbf{u}}^{\mathbf{i}}$ : we denote by $\boldsymbol{\Phi}_{\mathbf{u}, \star}^{\mathbf{i},}$ and $\boldsymbol{\Phi}_{\mathbf{u}}^{\mathbf{i}, \star}$ these operations.

The first results are "global" condition for regularity. Similar results may be found in the proofs of [2]; we state them here separately, with an additional uniformity on the starting point.

Theorem 6.1 (Regularity at jump times - global form). Let $x_{0}$ be a point in $M, K$ be an integer, $\mathbf{u}=\left(u_{1}, \ldots u_{K}\right)$ be a sequence of times, and $\mathbf{i}=\left(i_{1}, \ldots i_{K}\right)$ a sequence of indices. Suppose that $\mathbf{i}, \mathbf{u}$ is adapted to $x_{0}$ in the sense of Definition 4.5. For $k=1, \ldots K$, let $x_{k}=\Phi_{\mathbf{u}_{k}}^{\mathbf{i}_{k}}\left(x_{0}\right)$ be the successive points in the trajectory that follows the flow of $F^{i_{l}}$ for a time $u_{l}$.

Consider the family of tangent vectors at $x_{0}$ obtained by collecting tangent vectors at each jump, and pulling them all back to $x_{0}$ along the flows:

$$
\begin{equation*}
\left\{F^{i_{1}}\left(x_{0}\right),\left(\boldsymbol{\Phi}_{\mathbf{u}, \mathbf{1}}^{\mathbf{i}_{1}, \star} F^{i_{2}}\right)\left(x_{0}\right), \ldots\left(\boldsymbol{\Phi}_{\mathbf{u}_{K-1}}^{\mathbf{i}_{K-1, \star},} F^{i_{K}}\right)\left(x_{0}\right)\right\} . \tag{18}
\end{equation*}
$$



The trajectory begins by following $F^{1}$. At the first jump, it starts following $F^{2}$; we pull this tangent vector back to $x_{0}$. The next tangent vector $F^{3}$ (at $x_{2}$ ) has to be pulled back by the two flows. If the three tangent vectors we obtain at $x_{0}$ span $T_{x_{0}} M$, the law of $X_{T_{3}}$, the process at its third jump, is partly regular.

Figure 1: The global condition

If this collection spans the entire tangent plane $T_{x_{0}} M$, then the law of the process at its $K$-th jump is partly regular. More precisely, there exist $\mathcal{U}_{0}$ a neighborhood of $x_{0}, \mathcal{V}_{0}$ a neighborhood of $x_{K}$, and a constant $c>0$ such that:

$$
\begin{equation*}
\forall x \in \mathcal{U}_{0}, \quad \mathbb{P}_{x, i_{1}}\left[X_{T_{K}} \in \cdot\right] \geqslant c \lambda_{\mathbb{R}^{m}}\left(\cdot \cap \mathcal{V}_{0}\right) \tag{19}
\end{equation*}
$$

The condition on the starting index $i_{1}$ may be relaxed: there exist $\mathcal{U}_{0}^{\prime}$ and $\mathcal{V}_{0}^{\prime}$, a integer $K^{\prime}$ and $a$ constant $c^{\prime}$ such that:

$$
\begin{equation*}
\forall x \in \mathcal{U}_{0}^{\prime}, \forall i, \quad \mathbb{P}_{x, i}\left[X_{T_{K^{\prime}}} \in \cdot\right] \geqslant c \lambda_{\mathbb{R}^{m}}\left(\cdot \cap \mathcal{V}_{0}\right) \tag{20}
\end{equation*}
$$

Theorem 6.2 (Regularity at fixed time - global form). Let $\left(x_{k}\right)$, u, i be as above. Consider the following family:

$$
\begin{align*}
\left\{F^{i_{1}}\left(x_{0}\right)-\left(\boldsymbol{\Phi}_{\mathbf{u}_{K}}^{\mathbf{i}_{K}, \star} F^{i_{K+1}}\right)\left(x_{0}\right),\right. \\
\left(\boldsymbol{\Phi}_{\mathbf{u}_{1}}^{\mathbf{i}_{1}, \star} F^{i_{2}}\right)\left(x_{0}\right)-\left(\boldsymbol{\Phi}_{\mathbf{u}_{K}}^{\mathbf{i}_{K}, \star} F^{i_{K+1}}\right)\left(x_{0}\right), \\
\vdots  \tag{21}\\
\left.\left(\boldsymbol{\Phi}_{\mathbf{u}_{K-1}}^{\mathbf{i}_{K-1}, \star} F^{i_{K}}\right)\left(x_{0}\right)-\left(\boldsymbol{\Phi}_{\mathbf{u}_{K}}^{\mathbf{i}_{K}, \star} F^{i_{K+1}}\right)\left(x_{0}\right)\right\} .
\end{align*}
$$

It this family spans $T_{x_{0}} M$, then for any $t_{0}>u_{1}+\cdots+u_{K}$, there exist two constants $c$ and $\epsilon$, and two non-empty open sets $\mathcal{U}_{0}$ and $\mathcal{V}_{0}$ such that:

$$
\begin{equation*}
\forall x \in \mathcal{U}_{0}, \forall i, \forall t \in\left[t_{0}, t_{0}+\epsilon\right], \quad \mathbb{P}_{x, i}\left[X_{t} \in \cdot\right] \geqslant c \lambda_{\mathbb{R}^{m}}\left(\cdot \cap \mathcal{V}_{0}\right) \tag{22}
\end{equation*}
$$

Remark 6.3. Since the derivative $\left(D \mathbf{\Phi}_{\mathbf{i}_{K}, \mathbf{u}_{K}}\right)_{x_{0}}$ is an isomorphism from $T_{x_{0}} M$ to $T_{x_{K}} M$, we can write down this condition at $T_{x_{K}} M$ by pushing all tangent vectors forward to the end of the trajectory rather than pulling them back to the starting point.
Remark 6.4. The condition for regularity at the $K$-th jump involves $K$ vectors. Therefore, we need at least $d$ jumps to get regularity in dimension $d$.

This condition for having a regular part is necessary and sufficient (if it does not hold for any $n, \mathbf{u}, \mathbf{i}$ then the process is not regular). Unfortunately, it is not very easy to check, since one needs to "solve" the flows. However, it translates to a very nice local condition. To write down this condition, we need a bit of additional notation. Let $\mathcal{F}_{0}$ the collection of vector fields $\left(F^{i}: i \in E\right)$. Let $\mathcal{F}_{k}=\mathcal{F}_{k-1} \cup\left\{\left[F^{i}, V\right], V \in \mathcal{F}_{k-1}\right\}$, and $\mathcal{F}_{k}(x)$ the vector space (included in $T_{x} M$ ) spanned by $\left\{V(x), V \in \mathcal{F}_{k}\right\}$.

Similarly, starting from $\mathcal{G}_{0}=\left\{F^{i}-F^{j}, i \neq j\right\}$, we define $\mathcal{G}_{k}$ by taking Lie brackets with the $F^{i}$, and $\mathcal{G}_{k}(x)$ the corresponding subspace of $T_{x} M$.

Definition 6.5. We say that the weak bracket condition is satisfied at $x$ if there exists $k$ such that $\mathcal{F}_{k}(x)=T_{x} M$. If for some $k, \mathcal{G}_{k}(x)=T_{x} M$, we say that the strong bracket condition holds.

Since $\mathcal{G}_{k}(x)$ is a subspace of $\mathcal{F}_{k}(x)$, the strong condition implies the weak one. The converse is false, a counter-example is given below in Section 8.1.

These two conditions are called A (for the stronger) and B (for the weaker) in [2]. The following result is a version of Theorem 2 from [2], with an additional uniformity with respect to the initial point and the time $t$.

Theorem 6.6 (Regularity - local form, bracket condition). If the weak bracket condition holds at $x_{0}$, then the process is partly regular at jump times: there exist an integer $K^{\prime}$, a constant $c>0$ and non-empty open sets $\mathcal{U}_{0}, \mathcal{V}_{0}$ such that (20) holds.

If the strong bracket condition holds, the process is partly regular: there is a $t_{0}$, two constants $c>0$ and $\epsilon>0$, and two non-empty open sets $\mathcal{U}_{0}, \mathcal{V}_{0}$ such that (22) holds.

### 6.2 Three useful lemmas

The first result is a "uniform" local inversion lemma, for functions of $\mathbf{t}$ that depend on a parameter $x$.

Remark 6.7. Even if $x$ lives in some $\mathbb{R}^{k}$, we do not write it in boldface, for the sake of coherence with the rest of the paper.

Lemma 6.8. Let $m, n, k$ be integers, with $m \leqslant n$. Let $f$ be a $\mathcal{C}^{1}$ map from $\mathbb{R}^{n} \times \mathbb{R}^{k}$ to $\mathbb{R}^{n}$,

$$
f:(\mathbf{t}, x) \mapsto f(\mathbf{t}, x)=f_{x}(\mathbf{t})
$$

For any fixed $x, f_{x}$ maps $\mathbb{R}^{n}$ to itself; we denote its derivative at $\mathbf{t}$ by $\left(D f_{x}\right)_{\mathbf{t}}$. Suppose that, for some points $x_{0}$ and $\mathbf{t}_{0},\left(D f_{x_{0}}\right)_{\mathbf{t}_{0}}$ is invertible. Then we can find a neighborhood $J \subset \mathbb{R}^{k}$ of $x_{0}$, an open set $I \subset \mathbb{R}^{n}$ and, for all $x \in J$, an open set $W_{x} \subset \mathbb{R}^{n}$, such that:

$$
\tilde{f}_{x}: \begin{cases}W_{x} & \rightarrow I, \\ \mathbf{t} & \mapsto f_{x}(\mathbf{t})\end{cases}
$$

is a diffeomorphism. Moreover, for any $m \leqslant n$, and any neighborhood $W$ of $\mathbf{t}_{0}$, we can choose $I, J$ and the $W_{x}$ so that:
i) $I$ is a cartesian product $I_{1} \times I_{2}$ where $I_{1} \subset \mathbb{R}^{m}, I_{2} \subset \mathbb{R}^{n-m}$;
ii) $\forall x \in J, \quad W_{x} \subset W$.

Proof. We "complete" the map $f$ by defining:

$$
H: \begin{cases}\mathbb{R}^{n} \times \mathbb{R}^{k} & \rightarrow \mathbb{R}^{n} \times \mathbb{R}^{k} \\ (\mathbf{t}, x) & \mapsto\left(f_{x}(\mathbf{t}), x\right)\end{cases}
$$

$H$ is $\mathcal{C}^{1}$, and its derivative can be written in block form:

$$
D H_{(\mathbf{t}, x)}=\left(\begin{array}{cc}
\left(D f_{x}\right)_{\mathbf{t}} & \star \\
0 & I_{k}
\end{array}\right) .
$$

Since $\left(D f_{x_{0}}\right)_{\mathbf{t}_{0}}$ is invertible, $(D H)_{\mathbf{t}_{0}, x_{0}}$ is invertible. We apply the local inversion theorem to $H$ : there exist open sets $\mathcal{U}_{0}, \mathcal{V}_{0}$ such that $H$ maps $\mathcal{U}_{0}$ to $\mathcal{V}_{0}$ diffeomorphically. In order to
satisfy the properties i) andii), we restrict $H$ two times. First we define $\mathcal{U}_{1}=\mathcal{U}_{0} \cap\left(W \times \mathbb{R}^{k}\right)$, and $\mathcal{V}_{1}=H\left(\mathcal{U}_{1}\right)$. Since $\mathcal{V}_{1}$ is open it contains a product set $\mathcal{V}=I_{1} \times I_{2} \times J$, and we let $\mathcal{U}=H^{-1}(\mathcal{V})$. For any $(y, x) \in I \times J$, define $g_{x}(y)$ the first component of $H^{-1}(y, x)$ : composing by $H$, we see that $f_{x}\left(g_{x}(y)\right)=y$.

Now, define $W_{x}=\left\{\mathbf{t} \in \mathbb{R}^{n} ;(\mathbf{t}, x) \in \mathcal{U}\right\}$. $W_{x}$ is open, and included in $W$. Since $f_{x}$ maps $W_{x}$ to $I, g_{x}$ is its inverse and both are continuous, so $\tilde{f}_{x}$ is a diffeomorphism.

Lemma 6.9. Let $T$ be a continuous random variable in $\mathbb{R}^{n}$, with density $f_{T}$. Let $m \leqslant n$, and let $\phi$ be a $\mathcal{C}^{1}$ map from $\mathbb{R}^{n} \times \mathbb{R}^{m}$ to $\mathbb{R}^{m}$ :

$$
\phi:(\mathbf{t}, x) \mapsto \phi_{x}(\mathbf{t})
$$

Suppose that, for some $x_{0}, \mathbf{t}_{0},\left(D \phi_{x_{0}}\right)_{\mathbf{t}_{0}}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ has full rank $m$. Suppose additionally that $f_{T}$ is bounded below by $c>0$ on a neighborhood of $\mathbf{t}_{0}$.

Then there exist a constant $c^{\prime}>0$, a neighborhood $J$ of $x_{0}$ and a neighborhood $I_{1}$ of $\phi_{x_{0}}\left(\mathbf{t}_{0}\right)$ such that:

$$
\begin{equation*}
\forall x \in J, \quad \mathbb{P}[\phi(T, x) \in \cdot] \geqslant c^{\prime} \lambda_{\mathbb{R}^{m}}\left(\cdot \cap I_{1}\right) \tag{23}
\end{equation*}
$$

In other words, $\phi(T, x)$ has an absolutely continuous part w.r.t. the Lebesgue measure.
Proof. We know that $\left(D \phi_{x_{0}}\right)_{\mathbf{t}_{0}}$ has rank $m$. Without loss of generality, we suppose that the first $m$ columns are independent. In other words, writing $\mathbf{t}=(\mathbf{u}, \mathbf{v}) \in \mathbb{R}^{m} \times \mathbb{R}^{n-m}$, we suppose that the derivative of $\psi_{x, \mathbf{v}}: \mathbf{u} \mapsto \phi_{x_{0}}(\mathbf{u}, \mathbf{v})$ is invertible in $\mathbf{u}_{0}$ for $\mathbf{v}=\mathbf{v}_{0}$.

Once more, we "complete" $\phi$ and define:

$$
f_{x}: \begin{cases}\mathbb{R}^{m} \times \mathbb{R}^{n-m} & \rightarrow \mathbb{R}^{m} \times \mathbb{R}^{n-m} \\ (\mathbf{u}, \mathbf{v}) & \mapsto\left(\phi_{x}(\mathbf{u}, \mathbf{v}), \mathbf{v}\right)\end{cases}
$$

By Lemma 6.8, we can find $I_{1} \subset \mathbb{R}^{m}, I_{2} \subset \mathbb{R}^{n-m}, J \subset \mathbb{R}^{m}$ and $\left(W_{x}\right)_{x \in J} \subset \mathbb{R}^{n}$ such that $f_{x}$ maps diffeomorphically $W_{x}$ to $I_{1} \times I_{2}$. Call $\tilde{f}_{x}$ this diffeomorphism. By property ii) of the lemma, we can ensure that $W_{x}$ is included in a given neighborhood of $\mathbf{t}_{0}$. Since $D f_{x}=\left(\begin{array}{cc}D \psi_{x, \mathbf{v}} & \star \\ 0 & I\end{array}\right)$, we can choose this neighborhood so that:

$$
\begin{equation*}
\forall x \in J, \forall \mathbf{t} \in W_{x}, \quad f_{T}(\mathbf{t})\left|\operatorname{det}\left(\left(D f_{x}\right)_{\mathbf{t}}\right)\right|^{-1} \geqslant c^{\prime \prime}>0 \tag{24}
\end{equation*}
$$

for some strictly positive constant $c^{\prime \prime}$.
Write the random variable $T$ as a couple $(U, V)$, and let $A$ be a Borel set included in $I_{1}$.

$$
\begin{aligned}
\mathbb{P}[\phi(T, x) \in A] & \geqslant \mathbb{P}\left[\phi(T, x) \in A, V \in I_{2}\right] \\
& =\mathbb{P}\left[f_{x}(U, V) \in A \times I_{2}\right] \\
& \geqslant \mathbb{P}\left[(U, V) \in \tilde{f}_{x}^{-1}\left(A \times I_{2}\right)\right] \\
& =\int_{\tilde{f}_{x}^{-1}\left(A \times I_{2}\right)} f_{T}(\mathbf{u}, \mathbf{v}) d \mathbf{u} d \mathbf{v} \\
& =\int_{\tilde{f}_{x}^{-1}\left(A \times I_{2}\right)} f_{T}(\mathbf{u}, \mathbf{v})\left|\operatorname{det}\left(\left(D \tilde{f}_{x}\right)_{\mathbf{u}, \mathbf{v}}\right)\right|^{-1} \cdot\left|\operatorname{det}\left(\left(D \tilde{f}_{x}\right)\right)_{\mathbf{u}, \mathbf{v}}\right| d \mathbf{u} d \mathbf{v}
\end{aligned}
$$

Since $\tilde{f}_{x}^{-1}\left(A \times I_{2}\right) \subset W_{x}$, we may use the bound (24). Then we can change variables by defining $(\mathbf{s}, \mathbf{v})=\tilde{f}_{x}(\mathbf{u}, \mathbf{v})$. We obtain:

$$
\begin{aligned}
\mathbb{P}[\phi(T, x) \in A] & \geqslant c^{\prime \prime} \int_{\tilde{f}_{x}^{-1}\left(A \times I_{2}\right)}\left|\operatorname{det}\left(\left(D \tilde{f}_{x}\right)\right)_{\mathbf{u}, \mathbf{v}}\right| d \mathbf{u} d \mathbf{v} \\
& =c^{\prime \prime} \int_{A \times I_{2}} d \mathbf{s} d \mathbf{v} \\
& \geqslant c^{\prime \prime} \lambda_{\mathbb{R}^{m}}(A) \lambda_{\mathbb{R}^{n-m}}\left(I_{2}\right)
\end{aligned}
$$

Therefore (23) holds with $c=c^{\prime \prime} \lambda_{\mathbb{R}^{n-m}}\left(I_{2}\right)$.
Finally we need the following result.
Lemma 6.10. Let $U$ be a non-empty open set. There exist $0<\epsilon_{1}<\epsilon_{2}, k \in \mathbb{N}$ and an open set $U^{\prime} \subset U$ such that:

$$
\begin{array}{rr}
\forall x \in U^{\prime}, \forall i, j, & \mathbb{P}_{x, i}\left[\forall t \in\left[\epsilon_{1}, \epsilon_{2}\right], X_{t} \in U, Y_{t}=j\right] \geqslant c, \\
\forall x \in U^{\prime}, \forall i, j & \mathbb{P}_{x, i}\left[X_{T_{k}} \in U, Y_{T_{k}}=j\right] \geqslant c .
\end{array}
$$

Proof. The vector fields $F^{i}$ are continuous and the space is compact, so the speed of the process is bounded: given the open set $U$, we can find $U^{\prime} \subset U$ and $\epsilon_{2}$ such that, starting from $U^{\prime}$, the process cannot exit $U$ before time $\epsilon_{2}$. Let $\epsilon_{1}=\epsilon_{2} / 2$ : there is a positive probability that between $t=0$ and $t=\epsilon_{1}$, the index jumps exactly one time, from $i$ to $j$, and does not jump again before time $t=\epsilon_{2}$. The second result is similar; if all jump rates are bounded below, we can even choose $k=1$.

### 6.3 Proofs for the global criteria

Since the flows are, by essence, deterministic, the randomness can only come from one source: the jump times. The proofs of (19) and (22) are very similar. We prove the former in a simple case, to illustrate the ideas, and only give a detailed proof of the latter.

Let us begin by the regularity of the law after the second jump, when the dimension is $d=2$, and there are only two flows governed by $F^{1}$ and $F^{2}$. Let the process start from $x_{0}$ along $F^{1}$, and let $X_{T_{2}}$ be its position at the moment of its second jump.

Let $E_{1}, E_{2}$ be independent exponential random variables. Using the flows, we can write:

$$
X_{T_{2}}=\Phi^{2}\left(E_{2}, \Phi^{1}\left(E_{1}, x_{0}\right)\right)=\phi\left(E_{1}, E_{2}\right)
$$

where $\phi$ is defined by:

$$
\begin{aligned}
\phi: \mathbb{R}^{2} & \rightarrow M \\
\left(u_{1}, u_{2}\right) & \mapsto \boldsymbol{\Phi}_{\left(u_{1}, u_{2}\right)}^{(1,2)}\left(x_{0}\right) .
\end{aligned}
$$

The derivative of $\phi$ is given by:

$$
\begin{aligned}
\frac{\partial \phi}{\partial u_{1}} & =D \phi \cdot \partial_{1}=\left(D \Phi^{2}\left(u_{2}, \cdot\right)\right)_{x_{1}} \cdot F^{1}\left(x_{1}\right)=\left(\mathbf{\Phi}_{u_{2}, \star}^{2} F^{1}\right)_{x_{2}} \\
\frac{\partial \phi}{\partial u_{2}} & =D \phi \cdot \partial_{2}=F^{2}\left(x_{2}\right)
\end{aligned}
$$

This family of two vectors in $T_{x_{2}} M$ has the same rank as its image by $\left(D \boldsymbol{\Phi}_{\left(u_{1}, u_{2}\right)}^{(1,2)}\right)^{-1}$ :

$$
\left(\boldsymbol{\Phi}_{u_{1}}^{1, \star} F^{1}, \mathbf{\Phi}_{\left(u_{1}, u_{2}\right)}^{(1,2), \star} F^{2}\right)_{x}=\left(F^{1}, \boldsymbol{\Phi}_{u_{1}}^{1, \star} F^{2}\right)_{x} .
$$

The hypothesis precisely states that this family spans $T_{x_{0}} M$, so $\phi$ has full rank at $\left.\left(u_{1}, u_{2}\right)\right)$. The product law $\mathcal{E}(1) \otimes \mathcal{E}(1)$ has a continuous density, which is strictly positive near $\left(u_{1}, u_{2}\right)$, so Lemma 6.9 applies and (19) holds, with $\mathcal{U}_{0}$ and $\mathcal{V}_{0}$ given by $J$ and $I_{1}$ of Lemma 6.9. The more general (20) follows by an application of the second item in Lemma 6.10,

Let us now turn to the proof of (22), i.e. the regularity of the law at a fixed time under the global hypothesis. The idea is very similar. To avoid additional complications we suppose that the jump rates $\lambda(i, j, x)$ are strictly positive, so that all sequences are adapted. Let $\mathbf{i}=\left(i_{1}, \ldots i_{K}\right), \mathbf{u}=\left(u_{1}, \ldots u_{K}\right)$ be sequences of indices and times, and suppose $t_{0}$ satisfies $t_{0}>\sum_{l} u_{l}$. Let $\tilde{\mathbf{i}}=\mathbf{i} \cdot\left(i_{K+1}\right)$, $\tilde{\mathbf{u}}=\mathbf{u} \cdot\left(t_{0}-\sum u_{k}\right)$. The final position of the process is given by $X_{t_{0}}=\boldsymbol{\Phi}_{\tilde{\mathbf{u}}}^{\tilde{\mathbf{u}}}\left(x_{0}\right)$. Fixing $\tilde{\mathbf{i}}$, this position can be seen as $\psi(\mathbf{u})$, where the function $\psi$ (which depends on $\tilde{\mathbf{i}}$ and $t_{0}$ ) maps $\mathbb{R}^{K}$ to $M$. Once more, we differentiate with respect to $u_{k}$. Since $u_{k}$ appears twice in $\tilde{\mathbf{u}}$, we get:

$$
\frac{\partial \psi}{\partial u_{k}}(\mathbf{u})=(D \psi) \cdot \frac{\partial}{\partial u_{k}}=-F^{i_{K+1}}(\psi(\mathbf{u}))+\left(\boldsymbol{\Phi}_{u_{k+1}, \ldots u_{K}, t_{0}-\sum u_{j}, \star}^{i_{k+1}, \ldots i_{K+1}} F^{i_{k}}\right)(\psi(\mathbf{u}))
$$

which is a tangent vector at $\psi(\mathbf{u})$. Mapping it back to $T_{x_{0}} M$, together with the vectors obtained for the other choices of $k$, we obtain the family of vectors described in the hypothesis. Therefore $\psi$ has full rank as soon as the vectors defined in (21) span $T_{x_{0}} M$.

Define two events
$A=$ "the process makes $k$ jumps before time $t "$
$B=$ "the first $k$ jumps are real and the indices followed are given by $\mathbf{i} "$.
and let $\mathcal{F}$ be the sigma field generated by the waiting times $E_{k}$. The construction of the process via independent exponential random variables (Section (2) shows that:

$$
\mathbb{P}[B \mid \mathcal{F}]=\prod_{l} \frac{\lambda\left(i_{l-1}, i_{l}, \Phi_{\left(E_{1}, \ldots E_{l}\right)}^{\mathbf{i}_{l}}\left(x_{0}\right)\right)}{\lambda} \geqslant c
$$

Therefore

$$
\begin{aligned}
\mathbb{P}\left[X_{t} \in \cdot\right] & \geqslant \mathbb{P}\left[A \cap B \cap\left(X_{t} \in \cdot\right)\right] \\
& =\mathbb{P}\left[A \cap B \cap\left(\psi\left(E_{1}, \ldots E_{k}\right) \in \cdot\right)\right] \\
& =\mathbb{E}\left[\mathbb{P}[B \mid \mathcal{F}] \mathbb{1}_{A} \mathbb{1}_{\psi\left(E_{1}, \ldots E_{k}\right) \in \cdot}\right] \\
& \geqslant c \mathbb{P}\left[A \cap \psi\left(E_{1}, \ldots E_{k}\right) \in \cdot\right] \\
& \geqslant c e^{-\lambda t_{0}} \mathbb{P}\left[\psi\left(E_{1}, \ldots E_{k}\right) \in \cdot\right]
\end{aligned}
$$

since $A \supset\left\{E_{k+1} \geqslant t_{0}\right\}$. Thanks to Lemma 6.9, this implies (22), but only with $i=i_{1}$ and $\epsilon=0$.

To prove the general form of (22) with the additional freedom in the choice of $i$ and $t$, we need to restrict the starting point. By Lemma 6.10, we can find a neighborhood $\mathcal{U}_{0}^{\prime}$ of $x_{0}$, and three constants $0<\epsilon_{1}<\epsilon_{2}$ and $c>0$ such that:

$$
\forall x \in \mathcal{U}_{0}^{\prime}, \quad \mathbb{P}_{x, i}\left[\forall t \in\left[\epsilon_{1}, \epsilon_{2}\right], X_{t} \in \mathcal{U}_{0}, I_{t}=i_{1}\right] \geqslant c
$$

Let $t_{0}^{\prime}=t_{0}-\epsilon_{1}$ and $\epsilon=\epsilon_{2}-\epsilon_{1}$, so that $\left[t_{0}^{\prime}, t_{0}^{\prime}+\epsilon\right]=\left[t_{0}+\epsilon_{1}, t_{0}+\epsilon_{2}\right]$. Then, for any $x \in \mathcal{U}_{0}^{\prime}$, and any $t \in\left[t_{0}^{\prime}, t_{0}^{\prime}+\epsilon\right]$,

$$
\begin{aligned}
\mathbb{P}_{x, i}\left[X_{t} \in \cdot\right] & \geqslant \mathbb{E}_{x, i}\left[\mathbb{1}_{X_{t-t_{0}} \in \mathcal{U}_{0}, I_{t-t_{0}}=i_{1}} \mathbb{P}_{X_{t-t_{0}}, I_{t-t_{0}}}\left[X_{t_{0}} \in \cdot\right]\right] \\
& \geqslant c^{\prime} \lambda_{\mathbb{R}^{m}}\left(\cdot \cap \mathcal{V}_{0}\right) .
\end{aligned}
$$

This concludes the proof of Theorem 6.2,

### 6.4 Proof of the local criteria

Suppose that the weak bracket condition is satisfied at some point $x$. Then, by Theorem 5 of [2], the global condition (18) holds for some integer $K$. Similarly, under the strong bracket condition, Theorem 4 of [2] shows that (21) holds. This concludes the proof of Theorem 6.6.

## 7 Ergodicity

### 7.1 The embedded chain

Theorem 7.1. Suppose there exists $p \in \Gamma$ at which the weak bracket condition holds. Then the chain $\tilde{Z}$ admits a unique invariant probability $\tilde{\pi}$, absolutely continuous with respect to $\lambda$ and

$$
\left\|\mathbb{P}\left(\tilde{Z}_{n} \in .\right)-\tilde{\pi}\right\| \leqslant C \rho^{n}
$$

for some constants $C, 0 \leqslant \rho<1$ where $\|\cdot\|$ stands for the total variation norm.
Proof. By Theorem 6.6, (20) holds for some $K^{\prime}$. It is easy to see that, up to a restriction of $\mathcal{U}_{0}^{\prime}$ and a change of $c$, (20) also holds for $K^{\prime \prime}=K^{\prime}+1$. By Proposition 4.2, the chain comes back to $\mathcal{U}_{0}^{\prime}$. Therefore, by usual regeneration/coupling arguments for Harris chains, detailed e.g. in [18], Sections III.8-10, $\tilde{\pi}$ is unique; the fact that (20) holds for $K^{\prime \prime}$ guarantees that the embedded renewal process defined by the times of return to $\mathcal{U}_{0}^{\prime}$ is aperiodic. Once more by usual arguments, two copies of $\tilde{Z}$ may be coupled in a time $T$ that has geometric tails; this implies the exponential convergence (see e.g. [18], Section I. 3 for details).

To see that $\tilde{\pi} \ll \lambda$ observe that by Proposition 4.2 and Theorem $6.6 \pi \geqslant \delta c \psi$, where $\psi$, the Lebesgue measure on $\mathcal{V}_{0} \times E$, is absolutely continuous with respect to $\lambda$. Therefore $\tilde{\pi}_{a c}$ (the absolutely continuous part of $\tilde{\pi}$ with respect to $\lambda$ ) is non zero and the result follows from Proposition 5.10,

Corollary 7.2. Under the assumptions of Theorem [7.1, with probability one

$$
\lim _{n \rightarrow \infty} \tilde{\Pi}_{n}=\tilde{\pi}
$$

and

$$
\lim _{t \rightarrow \infty} \Pi_{t}=\tilde{\pi} \tilde{K}
$$

Proof. Follows from Theorem 7.1 and Propositions 5.2, 5.1) and 5.5.

### 7.2 The continuous time process

Lemma 7.3. If the strong bracket condition holds at $x_{0} \in \Gamma$, then for any neighborhood $U$ of $x_{0}$, there exist $t_{0}, t_{1}$, two non-empty open sets $\mathcal{U}_{0} \subset U$ and $\mathcal{V}_{1}$, and three constants $\alpha_{0}, \alpha_{1}$ and $\epsilon$ such that:

$$
\begin{gather*}
\forall x \in \mathcal{U}_{0}, \forall t \in\left[t_{0}, t_{0}+\epsilon\right], \forall i \quad \mathbb{P}_{x, i}\left[X_{t} \in \cdot \cap \mathcal{V}_{1}\right] \geqslant \alpha_{0} \lambda\left(\cdot \cap \mathcal{V}_{1}\right),  \tag{25}\\
\forall y \in \mathcal{V}_{1}, \forall i, \quad \mathbb{P}_{y, i}\left[X_{t_{1}} \in \mathcal{U}_{0}\right] \geqslant \alpha_{1} . \tag{26}
\end{gather*}
$$

Proof. The regularity result of Theorem 6.6 gives us $\mathcal{U}_{0}, \mathcal{V}_{0}, t_{0}$ and $\epsilon$ such that:

$$
\begin{equation*}
\forall x \in \mathcal{U}_{0}, \forall t \in\left[t_{0}, t_{0}+\epsilon\right], \forall i, \quad \mathbb{P}_{x, i}\left[X_{t} \in \cdot\right] \geqslant c \lambda\left(\cdot \cap \mathcal{V}_{0}\right) . \tag{27}
\end{equation*}
$$

At this stage we can replace $\mathcal{U}_{0}$ by $\mathcal{U}_{0} \cap U$ to ensure $\mathcal{U}_{0} \subset U$. By Lemma 4.3 we cover $M$ by a finite number of open sets $O_{1}, \ldots O_{k}$, such that, for some $\left(u_{1}, \ldots u_{k}\right)$,

$$
\begin{equation*}
\forall y \in O_{k}, \forall i, \quad \mathbb{P}_{y, i}\left[X_{u_{k}} \in \mathcal{U}_{0}\right] \geqslant \alpha_{2} \tag{28}
\end{equation*}
$$

One of those open sets (say $O_{l}$ ) has a non-empty intersection with $\mathcal{V}_{0}$ : call $\mathcal{V}_{1}$ this intersection. Since $\mathcal{V}_{1} \subset \mathcal{V}_{0}$, (27) implies (25). Since $\mathcal{V}_{1} \subset O_{l}$, (28) implies (26), with $t_{1}=u_{l}$.

Lemma 7.4. If the strong bracket condition holds at $x_{0} \in \Gamma$, then for any neighborhood $U$ of $x_{0}$, there exists a $t$ and an $\alpha>0$ such that for all $x$ and $i$,

$$
\begin{equation*}
\mathbb{P}_{x, i}\left[X_{t} \in U\right] \geqslant \alpha \tag{29}
\end{equation*}
$$

Proof. We reuse Lemma 4.3 to define open sets $O_{1}, \ldots O_{k}$, such that, for some $\left(u_{1}, \ldots u_{k}\right)$, Equation (28) holds. Now we apply Lemma 7.3 to find $t_{0}, t_{1}$ such that (25) and (26) hold.

Suppose without loss of generality that $\min \left(u_{i}\right)=u_{1} \leqslant u_{k}=\max \left(u_{i}\right)$. Let $s_{k}=t_{0}+t_{1}$. For any $n$, define $t=u_{k}+n s_{k}$. For $j<k$, define $s_{j}$ by:

$$
s_{j}=\frac{t-u_{j}}{n}=s_{k} \frac{t-u_{j}}{t-u_{k}}
$$

We have the following bounds:

$$
s_{k} \leqslant s_{j} \leqslant \frac{t-u_{1}}{t-u_{k}} s_{k}
$$

Choose $n$ large enough to ensure that $\frac{t-u_{1}}{t-u_{k}} \leqslant\left(1+\epsilon / s_{k}\right)$. Then each $s_{j}$ is is the interval $\left[t_{0}+t_{1}, t_{0}+\epsilon+t_{1}\right]$. Therefore we have found $t, n$ and $\left(s_{j}\right)$ such that:

$$
\begin{array}{r}
\forall j, \quad s_{j} \in\left[t_{0}+t_{1}, t_{0}+\epsilon+t_{1}\right], \\
\forall j, \quad t=u_{j}+n s_{j} .
\end{array}
$$

For $x \in O_{j}$, we have:

$$
\begin{aligned}
\mathbb{P}_{x, i}\left[X_{t} \in \mathcal{U}_{0}\right] & =\mathbb{P}_{x, i}\left[X_{u_{j}} \in \mathcal{U}_{0}, X_{u_{j}+s_{j}} \in \mathcal{U}_{0}, \ldots, X_{t} \in \mathcal{U}_{0}\right] \\
& \geqslant \mathbb{E}_{x, i}\left[\prod_{l=1}^{n} \mathbb{1}_{X_{u_{j}+l s_{j}} \in \mathcal{U}_{0}}\right] \\
& =\mathbb{E}_{x, i}\left[\prod_{l=1}^{n-1} \mathbb{1}_{X_{u_{j}+l s_{j}}} \mathcal{U}_{0} \cdot \mathbb{P}_{X_{u_{j}+(n-1) s_{j}}, I_{u_{j}+(n-1) s_{j}}}\left[X_{s_{j}} \in \mathcal{U}_{0}\right]\right]
\end{aligned}
$$

by the Markov property. Now, for any $x \in \mathcal{U}_{0}$ and any $i$,

$$
\begin{aligned}
\mathbb{P}_{x, i}\left[X_{s_{j}} \in \mathcal{U}_{0}\right] & \geqslant \mathbb{P}_{x, i}\left[X_{s_{j}-t_{1}} \in \mathcal{V}_{1}, X_{s_{j}} \in \mathcal{U}_{0}\right] \\
& \geqslant \mathbb{E}_{x, i}\left[\mathbb{1}_{X_{s_{j}-t_{1}} \in \mathcal{V}_{1}} \mathbb{P}_{X_{s_{j}-t_{1}, I_{s_{j}-t_{1}}}}\left[X_{t_{1}} \in \mathcal{U}_{0}\right]\right] \\
& \geqslant \alpha_{1} \mathbb{P}_{x, i}\left[X_{s_{j}-t_{1}} \in \mathcal{V}_{1}\right] \\
& \geqslant \alpha_{0} \alpha_{1} \lambda\left(\mathcal{V}_{1}\right)
\end{aligned}
$$

since $s_{j}-t_{1} \in\left[t_{0}, t_{0}+\epsilon\right]$. Therefore

$$
\mathbb{P}_{x, i}\left[X_{t} \in \mathcal{U}_{0}\right] \geqslant \alpha_{0} \alpha_{1} \lambda\left(\mathcal{V}_{1}\right) \mathbb{E}_{x, i}\left[\prod_{l} \mathbb{1}_{X_{u_{j}+l s_{j}} \in \mathcal{U}_{0}}\right]
$$

By induction and thanks to (28), we get:

$$
\mathbb{P}_{x, i}\left[X_{t} \in U\right] \geqslant \alpha_{2}\left(\alpha_{0} \alpha_{1} \lambda\left(\mathcal{V}_{1}\right)\right)^{n}
$$

Therefore (29) holds with $\alpha=\alpha_{2}\left(\alpha_{0} \alpha_{1} \lambda\left(\mathcal{V}_{1}\right)\right)^{n}$.
Theorem 7.5. Supose that there is a point $x \in \Gamma$ at which the strong bracket condition is satisfied. Let $\tilde{\pi}$ be the unique invariant probability of $\tilde{Z}$ and let $\pi=\pi \tilde{K}$. Then

$$
\begin{equation*}
\exists \alpha>0, C>0, \quad\left\|\mathbb{P}\left(Z_{t} \in \cdot\right)-\pi\right\| \leqslant C \exp (-\alpha t) \tag{30}
\end{equation*}
$$

Remark 7.6. Theorem 1 of [2] states that, if the weak bracket condition is satisfied at $x \in \Gamma$, and if $\left(Z_{t}\right)$ has an invariant measure, then it is unique and absolutely continuous with respect to the Lebesgue measure on $M \times E$. Under the strong bracket assumption, we have shown that the distribution of $Z_{t}$ itself converges, and not only its empirical measure.

Proof of Theorem 7.5. Lemma 7.4 shows that two processes starting from anywhere can be coupled in some time $t$ with positive probability. This implies (30) by the usual coupling argument.

## 8 Elementary examples

We give here a few examples of systems given by (2).

### 8.1 On the torus

Consider the system defined on the torus $\mathbb{T}^{n}=\mathbb{R}^{n} / \mathbb{Z}^{n}$ by the constant vector fields $F^{i}=e^{i}$, where $\left(e_{1}, \ldots e_{n}\right)$ is the standard basis on $\mathbb{R}^{n}$. Then, as argued in [2], the weak bracket condition holds everywhere, and the strong condition does not hold. Therefore the chain $\tilde{Z}$ is ergodic and converges exponentially fast, the empirical means of $\tilde{Z}_{n}$ and $Z_{t}$ converge, but the law of $Z_{t}$ is singular with respect to the invariant measure.

### 8.2 Two planar linear flows

Let $A$ be a $2 \times 2$ real matrix whose eigenvalues $\eta_{1}, \eta_{2}$ have negative real parts. Set $E=\{0,1\}$ and consider the process defined on $\mathbb{R}^{2} \times E$ by

$$
F^{0}(x)=A x \quad \text { and } \quad F^{1}(x)=A(x-a)
$$

for some $a \in \mathbb{R}^{2}$. The associated flows are $\Phi_{t}^{0}(x)=e^{t A} x$ and $\Phi_{t}^{1}(x)=e^{t A}(x-a)+a$.
First note that, by using the Jordan decomposition of $A$, it is possible to find a scalar product $\langle\cdot\rangle$ on $\mathbb{R}^{2}$ (depending on $A$ ) and some number $0<\alpha \leqslant \min \left(-\operatorname{Re}\left(\eta_{1}\right),-\operatorname{Re}\left(\eta_{2}\right)\right)$ such that $\langle A x, x\rangle \leqslant-\alpha\langle x, x\rangle$. Therefore

$$
\langle A(x-a), x\rangle \leqslant-\alpha\langle x, x\rangle-\langle A a, x\rangle \leqslant\|x\|(-\alpha\|x\|+\|A a\|) .
$$

This shows that, for $R>\|A a\| / \alpha$, the ball $M=\left\{x \in \mathbb{R}^{2},\|x\| \leqslant R\right\}$ is positively invariant by $\Phi^{0}$ and $\Phi^{1}$. Moreover every solution to the differential inclusion induced by $\left\{F^{0}, F^{1}\right\}$ eventually enters $M$. In particular $M \times E$ is an absorbing set for the process $\left(Z_{t}\right)$.

Another remark that will prove useful in our analysis is that

$$
\operatorname{det}\left(F^{0}(x), F^{1}(x)\right)=\operatorname{det}(A) \operatorname{det}(a, x)
$$

so that

$$
\begin{equation*}
\operatorname{det}\left(F^{0}(x), F^{1}(x)\right)>0 \quad(\text { resp. }=0) \Leftrightarrow \operatorname{det}(a, x)>0 \quad(\text { resp. }=0) \tag{31}
\end{equation*}
$$

## Case 1: $a$ is an eigenvector

If $a$ is an eigenvector of $A$ the line $\mathbb{R} a$ is invariant by both flows, so that

$$
\Gamma=\overline{\gamma^{+}(0)}=[0, a]
$$

and there is one unique invariant probability $\pi$ (whose support has to be $\Gamma$ by Proposition [5.7) Indeed, it is easily seen that $\Gamma$ is an attractor for the set-valued dynamics induced by $F^{0}$ and $F^{1}$. Therefore the support of every invariant measure equals $\Gamma$. If we consider the process restricted to $\Gamma$, it becomes one-dimensional and the strong bracket condition holds, proving uniqueness.

Remark 8.1. If $X(0) \notin \mathbb{R} a, X$ will never reach $\Gamma$. As a consequence, the law of $X_{t}$ and $\pi$ are singular. In particular, their total variation distance is constant, equal to 1 . Note also that the strong bracket condition being satisfied everywhere except on $\mathbb{R} a$, the law of $X_{t}$ at finite times has a regular part.

Remark 8.2. Consider the following example: $A=-I d, a=(1,0)$ and $\mathbb{R} a$ is identified to $\mathbb{R}$. If the jump rates are constant and equal to $\lambda$, it is easy to check (see [17, 22]) that the invariant measure $\mu$ on $[0,1] \times\{0,1\}$ is given by:

$$
\mu=\frac{1}{2}\left(\mu_{0} \times \delta_{0}+\mu_{1} \times \delta_{1}\right)
$$

where $\mu_{0}$ and $\mu_{1}$ are Beta laws on $[0,1]$,

$$
\begin{aligned}
& \mu_{0}(d x)=C_{\lambda} x^{\lambda-1}(1-x)^{\lambda} \\
& \mu_{1}(d x)=C_{\lambda} x^{\lambda}(1-x)^{\lambda-1}
\end{aligned}
$$

In particular, this example shows that the density of the invariant measure (with respect to the Lebesgue measure) may be unbounded: when the jump rate $\lambda$ is smaller than 1 , the densities blow up at 0 and 1.

## Case 2: Eigenvalues are reals and $a$ is not an eigenvector

Suppose $\eta_{1}, \eta_{2}<0$ and that $a$ is not an eigenvector.
Let $\gamma_{0}=\left\{\Phi_{t}^{0}(a), t \geqslant 0\right\}, \gamma_{1}=\left\{\Phi_{t}^{1}(0), t \geqslant 0\right\}$. Note that $\gamma_{1}$ and $\gamma_{0}$ are image of each other by the transformation $T(x)=a-x$. The curve $\gamma_{0}$ (respectively $\gamma_{1}$ ) crosses the line $\mathbb{R} a$ only at point $a$ (respectively 0 ). For, otherwise, the trajectory $t \mapsto \Phi_{t}^{0}(a)$ would have to cross the line $\operatorname{Ker}\left(A-\lambda_{1} I\right)$ which is invariant. This makes the curve $\gamma=\gamma_{0} \cup \gamma_{1}$ a simple closed curve in $\mathbb{R}^{2}$ crossing $\mathbb{R} a$ at 0 and $a$. By Jordan curve Theorem, $\mathbb{R}^{2} \backslash \gamma=B \cup U$ where $B$ is a bounded component and $U$ an unbounded one. We claim that

$$
\Gamma=\bar{B}
$$

To prove this claim, observe that thanks to (31), $F^{0}$ and $F^{1}$ both point inward $B$ at every point of $\gamma$. This makes $\bar{B}$ positively invariant by $\Phi^{0}$ and $\Phi^{1}$. Thus $\Gamma \subset \bar{B}$. Conversely, $\gamma \subset \Gamma$ (because 0 and $a$ are accessibles from everywhere). If $x \in B$ there exists $s>0$ such that $\Phi_{-s}^{0}(x) \in \gamma\left(\right.$ because $\left.\lim _{t \rightarrow-\infty} \Phi_{t}^{0}(a)=-\infty\right)$ and necessarily $\Phi_{-s}^{0}(x) \in \gamma_{1}$. This proves that $x \in \gamma^{+}(0)$. Finally note that the strong bracket condition is verified in $\Gamma \backslash \mathbb{R} a$ proving uniqueness and absolute continuity of the invariant probability.


Figure 2: Double rotation.

## Case 3: Eigenvalues are complex conjugates

Suppose now that the eigenvalues have a nonzero imaginary part. By Jordan decomposition, it is easily seen that trajectories of $\Phi^{i}$ converge in spiralling, so that the mappings $\tau^{i}(x)=\inf \left\{t>0: \Phi_{t}^{i}(x) \in \mathbb{R} a\right\}$ and $h^{i}(x)=\Phi_{\tau^{i}(x)}^{i}$ are well defined and continuous. Let $H: \mathbb{R} a \mapsto \mathbb{R} a$ be the map $h^{0} \circ h^{1}$ restricted to $\mathbb{R} a$. Since two different trajectories of the same flow have empty intersection, the sequence $x_{n}=H^{n}(0)$ is decreasing (for the ordering on $\mathbb{R} a$ inherited from $\mathbb{R}$.) Being bounded (recall that $M$ is compact and positively invariant), it converges to $x^{*} \in \mathbb{R} a$ such that $x^{*}=H\left(x^{*}\right)$. Let now $\gamma^{0}=\left\{\Phi_{t}^{1}\left(x^{*}\right), 0 \leqslant t \leqslant \tau^{1}\left(x^{*}\right)\right\}, \gamma^{1}=\left\{\Phi_{t}^{0}\left(h^{1}\left(x^{*}\right)\right), 0 \leqslant t \leqslant \tau^{0}\left(h^{1}\left(x^{*}\right)\right)\right\}$ and $\gamma=\gamma^{0} \cup \gamma^{1}$. Reasoning as previously shows that $\Gamma$ is the bounded component of $\mathbb{R}^{2} \backslash \gamma$ and that there is a unique invariant and absolutely continuous invariant probability.

We illustrate this situation in Figure 2, with

$$
A=\left(\begin{array}{cc}
-1 & -1 \\
1 & -1
\end{array}\right) \quad \text { and } \quad a=\binom{1}{0} .
$$

Remark 8.3. Note that if the jump rates are small, the situation is similar to the one described in Remark [8.2, the process spends most of its time near the attractive points, and the density is unbounded at these points. Since they are in the interior of $\Gamma$, the density is not even continuous in the interior of $\Gamma$.

## 9 Knowing the flows is not enough

In this section we study in detail a PDMP on $\mathbb{R}^{2}$, where the strong bracket condition holds everywhere except on $\Gamma$, and where there may be one or more invariant measures, depending on the dynamics of the discrete part of the process.

This model has been suggested by O. Radulescu. The continuous part of the process takes its values on $\mathbb{R}^{2}$ whereas its discrete part belongs to $\{0,1\}$. For sake of simplicity we will denote (in a different way than in the beginning of the paper) by $\left(X_{t}, Y_{t}\right) \in \mathbb{R}^{2}$
the continuous component. The discrete component $\left(I_{t}\right)_{t \geqslant 0}$ is a continuous time Markov chain on $E=\{0,1\}$ with jump rates $\left(\lambda_{i}\right)_{i \in E}$. Let $\alpha>0$. The two vector fields $F^{0}$ and $F^{1}$ are given by

$$
F^{0}(x, y)=\binom{-x+\alpha}{-y+\alpha} \quad \text { and } \quad F^{1}(x, y)=\binom{-x+\frac{\alpha}{1+y^{2}}}{-y+\frac{\alpha}{1+x^{2}}}
$$

with $(x, y) \in \mathbb{R}^{2}$. Notice that the quarter plane $(0,+\infty)^{2}$ is invariant under the action of the vector fields $F^{0}$ and $F^{1}$. If the support of the initial law of $(X, Y)$ is included in the quarter plane (which is assumed from now on), then it is still the case for the law of ( $X_{t}, Y_{t}$ ) at any time.

### 9.1 General properties of the two vector fields

Obviously, the vector fields $F^{0}$ has a unique stable point ( $\alpha, \alpha$ ), whereas $F^{1}$ may admits one or three critical points, according to the value of $\alpha$.

Lemma 9.1. Let us define

$$
a=\frac{\alpha+\sqrt{\left|\alpha^{2}-4\right|}}{2} \quad \text { and } \quad b=\left(\frac{\sqrt{4 / 27+\alpha^{2}}+\alpha}{2}\right)^{1 / 3}-\left(\frac{\sqrt{4 / 27+\alpha^{2}}-\alpha}{2}\right)^{1 / 3} .
$$

Notice that $b$ is positive and is the unique real solution of $b^{3}+b=\alpha$. Then

- if $\alpha \leqslant 2$, then $F^{1}$ admits a unique critical point $(b, b)$ and it is stable,
- if $\alpha>2$, then $F^{1}$ admits three critical points: $(b, b)$ is unstable whereas $\left(a, a^{-1}\right)$ and $\left(a^{-1}, a\right)$ are stable.

Proof. If $(x, y)$ is a critical point of $F^{1}$ then $(x, y)$ is solution of

$$
\left\{\begin{array}{l}
x\left(1+y^{2}\right)=\alpha \\
y\left(1+x^{2}\right)=\alpha .
\end{array}\right.
$$

As a consequence, $x$ is solution of

$$
0=x^{5}-\alpha x^{4}+2 x^{3}-2 \alpha x^{2}+\left(1+\alpha^{2}\right) x-\alpha=\left(x^{2}-\alpha x+1\right)\left(x^{3}+x-\alpha\right) .
$$

The equation $x^{3}+x-\alpha$ admits a unique real solution $b$. It belongs to $(0, \alpha)$. Obviously, if $\alpha \leqslant 2,(b, b)$ is the unique critical point of $F^{1}$ whereas, if $\alpha>2$ and $a$ and $a^{-1}$ are the roots of $x^{2}-\alpha x+1=0$, then $F^{1}$ admits the three critical points: $(b, b),\left(a, a^{-1}\right)$ and $\left(a^{-1}, a\right)$. Let us have a look to the stability of $(b, b)$. The Jacobian matrix of $F^{1}$ at $(x, y)$ is given by

$$
\operatorname{Jac}\left(F^{1}\right)(x, y)=\left(\begin{array}{cc}
-1 & -\frac{2 \alpha y}{\left(1+y^{2}\right)^{2}} \\
-\frac{2 \alpha x}{\left(1+x^{2}\right)^{2}} & -1
\end{array}\right) .
$$

Since $1+b^{2}=\alpha / b$ one gets that

$$
\operatorname{Jac}\left(F^{1}\right)(b, b)=\left(\begin{array}{cc}
-1 & -2+\frac{2 b}{\alpha} \\
-2+\frac{2 b}{\alpha} & -1
\end{array}\right)
$$

and its eigenvalues are given by

$$
\eta_{1}=-3+\frac{2 b}{\alpha}=-1-2 \frac{\alpha-b}{\alpha} \quad \text { and } \quad \eta_{2}=1-\frac{2 b}{\alpha}=\frac{b^{3}-b}{\alpha}
$$

and are respectively associated to the eigenvectors $(1,1)$ and $(1,-1)$. Since $b<\alpha, \eta_{1}$ is smaller than -1 . Moreover, $\eta_{2}$ has the same sign than $b-1$ i.e. the same sign than $\alpha-2$. As a conclusion, $(b, b)$ is stable (resp. unstable) if $\alpha<2$ (resp. $\alpha>2$ ).

Assume now that $\alpha>2$. Then

$$
\operatorname{Jac}\left(F^{1}\right)\left(a, a^{-1}\right)=\left(\begin{array}{rr}
-1 & -\frac{2 a}{\alpha} \\
-\frac{2}{\alpha a} & -1
\end{array}\right)
$$

and its two eigenvalues $-1 \pm 2 \alpha^{-1}$ are negative. The critical points $\left(a, a^{-1}\right)$ and $\left(a^{-1}, a\right)$ are stable.

In the sequel, we assume that $\alpha>2$. The sets

$$
\begin{aligned}
D & =\{(x, x): x>0\} \\
L & =\{(x, y): x>0 \text { and } 0<y<x\}, \\
U & =\{(x, y): y>0 \text { and } 0<x<y\}
\end{aligned}
$$

are invariant under the action of the flows $F^{0}$ and $F^{1}$. Moreover, the set $D$ (and in particular the unique stable point $(\alpha, \alpha)$ of $\left.F^{0}\right)$ is included in the stable manifold of the unstable equilibrium $(b, b)$ of $F^{1}$.

What happens if $(X, Y)$ starts at a point $(x, y) \in L$ ? The answer may depend on the parameters $\lambda_{0}, \lambda_{1}, \alpha$.

### 9.2 Transience

Lemma 9.2. Assume that $\left(X_{0}, Y_{0}\right) \in L$. Then, for any $t>0$,

$$
0 \leqslant X_{t}-Y_{t} \leqslant\left(X_{0}-Y_{0}\right) \exp \left(-\int_{0}^{t} \alpha\left(I_{s}\right) d s\right)
$$

with $\alpha(0)=1$ and $\alpha(1)=1-c \alpha<0$ with $c=(3 / 8) \sqrt{3}$.
Proof. If $I_{t}=0$ then

$$
\frac{d}{d t}\left(X_{t}-Y_{t}\right)=-\left(X_{t}-Y_{t}\right)
$$

On the other hand, if $I_{t}=1$ then

$$
\begin{aligned}
\frac{d}{d t}\left(X_{t}-Y_{t}\right) & =-\left(X_{t}-Y_{t}\right)+\alpha \frac{X_{t}^{2}-Y_{t}^{2}}{\left(1+X_{t}^{2}\right)\left(1+Y_{t}^{2}\right)} \\
& =-\left(1-\alpha h\left(X_{t}, Y_{t}\right)\right)\left(X_{t}-Y_{t}\right)
\end{aligned}
$$

where the function $h$ is defined on $[0, \infty)^{2}$ by

$$
h(x, y)=\frac{x+y}{\left(1+x^{2}\right)\left(1+y^{2}\right)}
$$



Figure 3: Trajectory of $(X, Y)$ (red line) in the plane with $\lambda_{0}=1, \lambda_{1}=0.6$ and $\alpha=5$.

The unique critical point of $h$ on $[0, \infty)^{2}$ is $(1 / \sqrt{3}, 1 / \sqrt{3})$ and $h$ reaches its maximum at this point:

$$
c:=\sup _{x, y>0} h(x, y)=\frac{3 \sqrt{3}}{8} .
$$

As a consequence, for any $t \geqslant 0$,

$$
\frac{d}{d t}\left(X_{t}-Y_{t}\right) \leqslant-\alpha\left(I_{t}\right)\left(X_{t}-Y_{t}\right) \quad \text { where } \quad\left\{\begin{array}{l}
\alpha(0)=1 \\
\alpha(1)=1-c \alpha
\end{array}\right.
$$

Integrating this relation concludes the proof.
Corollary 9.3. Assume that $\left(X_{0}, Y_{0}\right) \in L$. If $\lambda_{1}>\lambda_{0}(c \alpha-1)$ then $\left(X_{t}, Y_{t}\right)$ converges exponentially fast to $D$ almost surely. More precisely,

$$
\begin{equation*}
\limsup _{t \rightarrow \infty} \frac{1}{t} \log \left(X_{t}-Y_{t}\right) \leqslant-\frac{\lambda_{1}-(c \alpha-1) \lambda_{0}}{\lambda_{0}+\lambda_{1}}<0 \quad \text { a.s. } \tag{32}
\end{equation*}
$$

In particular, the process $(X, Y, I)$ admits a unique invariant measure $\mu$ which support is the set

$$
S=\{(x, x): x \in[b, \alpha]\} .
$$

Proof. The ergodic theorem for the Markov process $\left(I_{t}\right)_{t \geqslant 0}$ ensures that

$$
\frac{1}{t} \int_{0}^{t} \alpha\left(I_{s}\right) d s \xrightarrow[t \rightarrow \infty]{\text { a.s. }} \int \alpha(i) d \nu(i)
$$

where the invariant measure $\nu$ of the process $\left(I_{t}\right)_{t \geqslant 0}$ is the Bernoulli measure with parameter $\lambda_{0} /\left(\lambda_{0}+\lambda_{1}\right)$. The upper bound (32) is a straightforward consequence of Lemma 9.2, This ensures that the sets $L$ and $U$ are transient. At last, it is quite obvious that the set of recurrent points in $D$ is exactly $S$.

One can also get an estimate for the $p^{t h}$ moment of $X_{t}-Y_{t}$.


Figure 4: Trajectories of $X$ (blue line) and $Y$ (red line) with $\lambda_{0}=1, \lambda_{1}=0.6$ and $\alpha=5$.

Corollary 9.4. Assume that $\left(X_{0}, Y_{0}\right) \in L$. Let $p>0$ such that

$$
\begin{equation*}
\lambda_{1}>\left(\lambda_{0}+p\right)(c \alpha-1) . \tag{33}
\end{equation*}
$$

Then there exists two positive constants $c_{p}, \mu_{p}$ such that

$$
\mathbb{E}\left(\left|X_{t}-Y_{t}\right|^{p}\right) \leqslant c_{p} \mathbb{E}\left(\left|X_{0}-Y_{0}\right|^{p}\right) e^{-\mu_{p} t}
$$

Proof. Once again, Lemma 9.2 ensures that

$$
0 \leqslant \mathbb{E}\left(\left|X_{t}-Y_{t}\right|^{p}\right) \leqslant \mathbb{E}\left(\left|X_{0}-Y_{0}\right|^{p}\right) \mathbb{E}\left[\exp \left(-\int_{0}^{t} p \alpha\left(I_{s}\right) d s\right)\right]
$$

According to [3, Prop. 4.1], there exists $c_{p} \geqslant 1$ such that, for any $t \geqslant 0$,

$$
\frac{1}{c_{p}} e^{-\mu_{p} t} \leqslant \mathbb{E}\left[\exp \left(-\int_{0}^{t} p \alpha\left(I_{s}\right) d s\right)\right] \leqslant c_{p} e^{-\mu_{p} t}
$$

where $\mu_{p}=-\max \left\{\operatorname{Re} \eta: \eta \in \operatorname{Spec}\left(M_{p}\right)\right\}$ and

$$
M_{p}=\left(\begin{array}{cc}
-\lambda_{0}-p & \lambda_{0} \\
\lambda_{1} & -\lambda_{1}+p(c \alpha-1)
\end{array}\right)
$$

The real parts of the eigenvalues of $M_{p}$ are negative if and only if their sum $S$ is negative and their product $P$ is positive with

$$
\begin{aligned}
-S & =\lambda_{0}+\lambda_{1}+p(2-c \alpha) \\
P & =p\left(\lambda_{1}-(c \alpha-1)\left(\lambda_{0}+p\right)\right)
\end{aligned}
$$

The sum $S$ is always negative and the positivity of $P$ is given by (33).

### 9.3 Recurrence

In this section, we aim to show that $(X, Y, I)$ may admit several invariant measures if the jump rate $\lambda_{0}$ is large enough. Let us define

$$
U_{t}=\frac{X_{t}+Y_{t}}{2} \quad \text { and } \quad V_{t}=\frac{X_{t}-Y_{t}}{2}
$$

Of course $(U, V, I)$ is still a PDMP. If

$$
\frac{d}{d t}\binom{X_{t}}{Y_{t}}=F^{1}\left(X_{t}, Y_{t}\right) \quad \text { then } \quad \frac{d}{d t}\binom{U_{t}}{V_{t}}=G^{1}\left(U_{t}, V_{t}\right)
$$

with

$$
G^{1}(u, v)=\frac{1}{2}\left(\begin{array}{cc}
1 & 1 \\
1 & -1
\end{array}\right) F^{1}(u+v, u-v)=\binom{-u+\frac{\alpha\left(1+u^{2}+v^{2}\right)}{\left(1+(u+v)^{2}\right)\left(1+(u-v)^{2}\right)}}{-v+\frac{2 \alpha u v}{\left(1+(u+v)^{2}\right)\left(1+(u-v)^{2}\right)}}
$$

Corollary 9.3 ensures that, if $\lambda_{1} / \lambda_{0}$ is large enough, then $V_{t}$ goes to 0 exponentially fast.
Let us show that this is no longer true if $\lambda_{1} / \lambda_{0}$ is small enough. Let $\varepsilon>0$. Assume that, with positive probability, $V_{t} \in(0, \varepsilon)$ for any $t \geqslant 0$. Then, for any time $t \geqslant 0$, $\left(U_{t}, V_{t}\right) \in[b, \alpha] \times[0, \varepsilon]$. Indeed, one can show that the set $X_{t}+Y_{t} \geqslant 2 b$ for any $t \geqslant 0$ as soon as it is true at the initial time.

Lemma 9.5. Assume that $(u, v) \in[b, \alpha] \times[0, \varepsilon]$. Then there exists $u_{c} \in(b, \alpha)$ and $K, \delta, \gamma, \tilde{\gamma}>0$ (that do not depend on $\varepsilon$ ) such that $b_{\varepsilon}=b+K \varepsilon^{2}$ and

$$
G_{1}^{1}(u, v) \leqslant H_{1}^{1}(u, v) \quad \text { with } \quad H_{1}^{1}(u, v)=-\delta\left(u-b_{\varepsilon}\right)
$$

and

$$
G_{2}^{1}(u, v) \geqslant H_{2}^{1}(u, v) \quad \text { with } \quad H_{2}^{1}(u, v)=\left((\gamma+\tilde{\gamma}) \mathbb{1}_{\left\{u \leqslant u_{c}\right\}}-\tilde{\gamma}\right) v
$$

Proof. Notice firstly that

$$
\begin{equation*}
\left|\left(1+(u+v)^{2}\right)\left(1+(u-v)^{2}\right)-\left(1+u^{2}\right)^{2}\right| \leqslant K \varepsilon^{2} \tag{34}
\end{equation*}
$$

Thus, using that $u^{3}+u-\alpha=(u-b)\left(u^{2}+b u+\alpha / b\right)$ we get that

$$
\begin{aligned}
G_{1}^{1}(u, v) & \leqslant-u+\frac{\alpha}{1+u^{2}}+K \varepsilon^{2} \\
& \leqslant-(u-b) \frac{u^{2}+b u+\alpha / b}{1+u^{2}}+K \varepsilon^{2} \\
& \leqslant-(u-b) \frac{2 b^{2}+\alpha / b}{1+\alpha^{2}}+K \varepsilon^{2} .
\end{aligned}
$$

We get the desired upper bound for $G_{1}^{1}$ with

$$
\delta=\frac{2 b^{2}+\alpha / b}{1+\alpha^{2}} \quad \text { and } \quad b_{\varepsilon}=b+(K / \delta) \varepsilon^{2}
$$

Similarly, Equation (34) ensures that

$$
G_{2}^{1}(u, v) \geqslant v k(u) \quad \text { with } \quad k(u)=\frac{2 \alpha u}{\left(1+u^{2}\right)^{2}}-1-K \varepsilon^{2}
$$

Obviously, if $\varepsilon$ is small enough, $k(b)>0, k(\alpha)<0$ and $k$ is decreasing. Thus, if $\tilde{u}$ is the unique zero of $k$ on $(b, \alpha)$, then one can choose

$$
u_{c}=\frac{\tilde{u}+b}{2}, \quad \gamma=k\left(u_{c}\right) \quad \text { and } \quad \tilde{\gamma}=k(\alpha)
$$

To get a simpler bound in the sequel we can even set $\tilde{\gamma}=k(\alpha) \vee 1$.
Finally, define $H_{1}^{0}(u, v)=G_{1}^{0}(u, v)=-(u-\alpha)$ and $H_{2}^{0}(u, v)=G_{2}^{0}(u, v)=-v$ and introduce the $\operatorname{PDMP}(\tilde{U}, \tilde{V}, \tilde{I})$ where $\tilde{I}=I$ is the switching process of $(U, V, I)$ and $(\tilde{U}, \tilde{V})$ is driven by $H^{0}$ and $H^{1}$ instead of $G^{0}$ and $G^{1}$. From Lemma 9.5, we get that

$$
U_{t} \leqslant \tilde{U}_{t} \quad \text { and } \quad \tilde{V}_{t} \leqslant V_{t} \quad(t \geqslant 0)
$$

assuming that $\left(\tilde{U}_{0}, \tilde{V}_{0}, \tilde{I}_{0}\right)=\left(U_{0}, V_{0}, I_{0}\right)$. The last step is to study briefly the process $(\tilde{U}, \tilde{V}, \tilde{I})$. Let us firstly notice that if $\lambda_{1} / \lambda_{0}$ is small enough, then $\left(I_{s}, \tilde{U}_{s}\right)$ spends an arbitrary large amount of time near $\left(1, b_{\varepsilon}\right)$ (and $b_{\varepsilon}$ can be assumed smaller than $u_{c}$ if $\varepsilon$ is small enough). Thus

$$
\frac{1}{t} \log \frac{\tilde{V}_{t}}{\tilde{V}_{0}} \geqslant \frac{1}{t} \int_{0}^{t}\left((\gamma+\tilde{\gamma}) \mathbb{1}_{\left\{I_{s}=1, \tilde{U}_{s}<u_{c}\right\}}-\tilde{\gamma}\right) d s
$$

since $\tilde{\gamma} \geqslant 1$. The right hand side converges almost surely to a positive limit as soon as $\lambda_{1} / \lambda_{0}$ is small enough. This implies that $V$ cannot be bounded by $\varepsilon$ forever.

Corollary 9.6. If $\lambda_{1} / \lambda_{0}$ is small enough, the process $(X, Y, I)$ admits three ergodic measures.

Acknowledgements FM and PAZ thank MB for his kind hospitality and his coffee breaks. We acknowledge financial support from the Swiss National Foundation Grant FN 200021-138242/1 and the French ANR projects EVOL and ProbaGeo.

## References

[1] J.P. Aubin and A. Cellina, Differential Inclusions: Set-Valued Maps and Viability Theory, SpringerVerlag New York, Inc. Secaucus, NJ, USA, 1984.
[2] Y. Bakhtin and T. Hurth, Invariant densities for dynamical systems with random switching, preprint available on arXiv, 2012. 1, 1, 4.1, 5.8, 6.1, 6.1, 6.4, 7.6, 8.1
[3] J.-B. Bardet, H. Guérin, and F. Malrieu, Long time behavior of diffusions with Markov switching, ALEA 7 (2010), 151-170. 9.2
[4] M. Benaïm, J. Hofbauer, and S. Sorin, Stochastic approximations and differential inclusions, I. SIAM Journal on Optimization and Control 44 (2005), 328-348. 4.9
[5] M. Benaïm, S. Leborgne, F. Malrieu, and P.-A. Zitt, On the stability of planar randomly switched systems, preprint available on arXiv, 2012. 11
[6] _ Quantitative ergodicity for some switched dynamical systems, preprint available on arXiv, 2012. 1
[7] O. Boxma, H. Kaspi, O. Kella, and D. Perry, On/Off Storage Systems with State-Dependent Inpout, Outpout and Swithching Rates, Probability en the Engineering and Informational Siences 19 (2005), 1-14. 1
[8] E. Buckwar and M. G. Riedler, An exact stochastic hybrid model of excitable membranes including spatio-temporal evolution, Preprint http://www.ma.hw.ac.uk/maths/deptreps/HWM10-6.html, 2010. 1
[9] M. H. A. Davis, Piecewise-deterministic Markov processes: a general class of nondiffusion stochastic models, J. Roy. Statist. Soc. Ser. B 46 (1984), no. 3, 353-388, With discussion. MR MR790622 (87g:60062)
[10] ___, Markov models and optimization, Monographs on Statistics and Applied Probability, vol. 49, Chapman \& Hall, London, 1993.
[11] P. Diaconis and D. Freedman, Iterated random functions, SIAM Rev. 41 (1999), no. 1, 45-76. MR 1669737 (2000c:60102) 1
[12] M. Duflo, Random Iterative Models, Springer Paris, 2000. 5
[13] V. Dumas, F. Guillemin, and Ph. Robert, A Markovian analysis of additive-increase multiplicativedecrease algorithms, Adv. in Appl. Probab. 34 (2002), no. 1, 85-111. 1
[14] H. Furstenberg, Strict ergodicity and transformation of the torus, Amer. J. Math. 83 (1961), 573-601. MR 0133429 (24 \#A3263) 5.8
[15] C. Graham and Ph. Robert, Interacting multi-class transmissions in large stochastic networks, Ann. Appl. Probab. 19 (2009), no. 6, 2334-2361. 1
[16] _ Self-adaptive congestion control for multi-class intermittent connections in a communication network, arXiv, 2010. 1
[17] R. Karmakar and I. Bose, Graded and binary responses in stochastic gene expression, Physical Biology 197 (2004), no. 1, 197-214. 8.2
[18] T. Lindvall, Lectures on the coupling method, Wiley Series in Probability and Mathematical Statistics: Probability and Mathematical Statistics, John Wiley \& Sons Inc., New York, 1992, A WileyInterscience Publication. MR 1180522 (94c:60002) 7.1
[19] R. Mañé, Ergodic theory and differentiable dynamics, Ergebnisse der Mathematik und ihrer Grenzgebiete (3) [Results in Mathematics and Related Areas (3)], vol. 8, Springer-Verlag, Berlin, 1987, Translated from the Portuguese by Silvio Levy. MR 889254 (88c:58040) 5.8
[20] K. Pakdaman, M. Thieullen, and G. Wainrib, Fluid limit theorems for stochastic hybrid systems with application to neuron models, arXiv, 2009.
[21] N. S Papageorgiu, Existence Theorems for Differential Inclusions with Nonconvex Right Hand Side, Internat. J. Math. \& Sci. 9 (1986), no. 3, 459-469.
[22] O. Radulescu, A. Muller, and A. Crudu, Théorèmes limites pour des processus de Markov à sauts. Synthèse des résultats et applications en biologie moléculaire, Technique et Science Informatiques $\mathbf{2 6}$ (2007), no. 3-4, 443-469. 1. 8.2
[23] G. G. Yin and C. Zhu, Hybrid switching diffusions, Stochastic Modelling and Applied Probability, vol. 63, Springer, New York, 2010, Properties and applications. MR 2559912 (2010i:60226) 1

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[^0]:    ${ }^{1}$ meaning that $g^{i}$ is the restriction to $M$ of a smooth function on $\mathbb{R}^{m}$

