# ON THE BERNOULLI PROPERTY OF PLANAR HYPERBOLIC BILLIARDS 

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#### Abstract

We consider billiards in non-polygonal domains of the plane with boundary consisting of curves of three different types: straight segments, strictly convex inward curves and strictly convex outward curves of a special kind. The billiard map for these domains is known to have non-vanishing Lyapunov exponents a.e. provided that the distance between the curved components of the boundary is sufficiently large, and the set of orbits having collisions only with the flat part of the boundary has zero measure. Under a few additional conditions, we prove that there exists a full measure set of the billiard phase space such that each of its points has a neighborhood contained up to a zero measure set in one Bernoulli component of the billiard map. Using this result, we show that there exists a large class of planar hyperbolic billiards that have the Bernoulli property. This class includes the billiards in convex domains bounded by straight segments and strictly convex inward arcs constructed by Donnay.


## 1. Introduction

A planar billiard is the mechanical system consisting of a pointparticle moving freely inside a bounded domain $\Omega \subset \mathbb{R}^{2}$ with piecewise differentiable boundary, and being reflected off $\partial \Omega$ so that the angle of reflection equals the angle of incidence. This paper concerns hyperbolic billiards, i.e., billiards for which the corresponding map has no vanishing Lyapunov exponents. Maps with this property are not necessarily uniformly hyperbolic, but exhibit a weak form of hyperbolicity called non-uniform hyperbolicity [27].

The study of hyperbolic billiards was initiated by Sinai. In his seminal paper [28], he proved that billiards in 2-dimensional toral domains containing finitely many obstacles with strictly convex outward boundary are hyperbolic and K-mixing. In fact, Sinai billiards enjoy the Bernoulli property as well [17].

Later on, Bunimovich proved that also billiards in domains with boundary formed by strictly convex inward arcs and straight segments can be hyperbolic $[1,2]$. The most celebrated example of such a domain

[^0]is the stadium, the region bounded by two semi-circles connected by two parallel segments. The only strictly convex inward arcs allowed in Bunimovich billiards were arcs of circles. This limitation was eventually overcome by several researchers. Using new techniques for establishing the positivity of Lyapunov exponents [23, 32], Wojtkowski, Markarian and Donnay proved independently that besides arcs of circles many other strictly convex inward arcs can be used to construct hyperbolic billiards [16, 23, 25, 32]. Similar results were obtained by Bunimovich for a class of strictly convex arcs related to those of Donnay [5]. All these results showed that billiards in non-polygonal planar domains are hyperbolic if three conditions are fulfilled: B1) the strictly convex boundary components of the domain are of a special type, B2) the distance between these components and the other curved boundary components is sufficiently large, and B3) orbits having only collisions with straight segments of the boundary form a set of zero measure. A precise formulation of these conditions is given in Section 5 .

In this paper, we address the question whether a hyperbolic billiard has the Bernoulli property (for short, 'it is Bernoulli'), i.e., whether it is isomorphic to a Bernoulli shift. The Bernoulli property is the strongest among the ergodic properties: it implies $K$-property, mixing and ergodicity. The Bernoulli property was proved for several billiards, including the Sinai billiards, the Bunimovich billiards, the Wojtkowski billiards and other special hyperbolic billiards [4, 11, 12, 13, 22, 24, 30]. Despite that, there remain many planar hyperbolic billiards for which not even the ergodicity has been proved. Notably, among them, there are the billiards constructed by Donnay [16]. The goal of this paper is to fill this gap: we show that there exists a large class of hyperbolic billiards, including the Donnay billiards, that have the Bernoulli property.

The key ingredient in the proof of this result is a local ergodic theorem for hyperbolic planar billiards. Roughly speaking, our local ergodic theorem states that if a planar billiard satisfies conditions B1-B3 and the extra condition B4 (see Section 5), then there exists a full measure set $H$ in the billiard phase space with the property that each element of $H$ has a neighborhood contained $(\bmod 0)$ in one ergodic component of the billiard.

Condition B4 regards the singular set of the billiard, the set formed by the elements where the billiard map is not defined or is not twicedifferentiable. This set corresponds to the trajectories that hit a corner of the billiard domain or have a tangential collision with its boundary. Condition B4 requires the elements of the singular set whose trajectories have eventually collisions only with straight segments to form a negligible subset (in the measure theoretical sense) of the singular set.

As a matter of fact, the neighborhood in the conclusion of the local ergodic theorem belongs (mod 0$)$ to a single Bernoulli component
of the billiard map (for the definition of a Bernoulli component, see Theorem A. 6 in the appendix). As a consequence, every Bernoulli component of a billiard satisfying B1-B4 is open $(\bmod 0)$. We stress, that although this is a remarkable property, it is not enough to yield the Bernoulli property.

Our local ergodic theorem for billiards is derived from a local ergodic theorem for general hyperbolic symplectomorphisms with singularities [14], which is a generalization of a theorem of Liverani and Wojtkowski [22]. Whereas in [22], it is assumed the existence of a cone field that is continuous everywhere on the billiard phase space, in [14], the existence of such a cone field is required only on an open subset of the billiard phase space. The billiards considered in this paper admit in general only an invariant piecewise-continuous cone field, and so the local theorem in [22] does not apply to all them. The proof of both [14] and [22] rely on the method developed by Sinai to prove the ergodicity of dispersing billiards [28]. Refinements of Sinai's method were also obtained in [9, 21, 29].

We also mention that the trick used in [30] to prove the ergodicity of Wojtkowski billiards and based on rescaling a certain semi-metric does not work with the majority of our billiards. It works with Wojtkowski billiards, because their invariant cones have the property that $G_{y}^{-}=$ $G_{y}^{+}=d(x)$ (for the meaning of these symbols, see Section 4), but this property is not satisfied in general by our billiards.

The paper is organized as follows. In Section 2, we recall the definition of the billiard map and its main properties. In Section 3, we give the definition of the focusing times. We also give the definition of a focusing arc introduced by Donnay, and collect the main properties of these arcs. Section 4 is concerned with the theory of invariant cone fields, and their construction for billiards. In this section, we also recall further results on focusing arcs. In Section 5, we give a detailed description of the hyperbolic billiards considered, and state the main results of this paper. In Section 6, we prove Conditions L1-L3 of the local ergodic theorem. Section 7 is entirely devoted to the proof of Condition L4. In Section 8, we introduce the class of billiards in polygons with pockets and bumps, and prove that they have the Bernoulli property. As a corollary, we obtain that the Donnay billiards have the Bernoulli property. In the Appendix, we state the local ergodic theorem for general hyperbolic symplectomorphisms with singularities from [14] together with relevant concepts and observations.

Sections 2, 3 and the beginning of Section 4 are meant to be an introduction to basic concepts on billiards and geometric optics, Donnay arcs and the theory of invariant cone fields. The reader already familiar with these notions may read quickly or skip these sections, and move directly to the second part of Section 4.

## 2. Generalities on billiards

A billiard system can be described either by a flow or a map. In this paper, we focus on the billiard map. For the relation between the billiard map and billiard flow, we refer the reader to the book [10]. In this section, we define the billiard map for a 2-dimensional domain, and single out its basic properties: regularity, natural invariant probability and singular sets.
2.1. Billiard domain. Let $0 \leq k \leq \infty$. A subset $\Gamma \subset \mathbb{R}^{2}$ is called an arc of class $C^{k}$ if $\Gamma$ is the image of a $C^{k}$ embedding $\gamma:[0,1] \rightarrow \mathbb{R}^{2}$. The boundary of $\Gamma$ is the set $\partial \Gamma=\gamma(0) \cup \gamma(1)$. A subset $\Gamma \subset \mathbb{R}^{2}$ is called a closed curve of class $C^{k}$ if $\Gamma$ is $C^{k}$ diffeomorphic to the unit circle $S^{1}$. For a closed curve, we set $\partial \Gamma:=\emptyset$.

Let $\Omega$ be an open bounded connected subset of $\mathbb{R}^{2}$ with boundary $\partial \Omega$ consisting of finitely many arcs and closed curves $\Gamma_{1}, \ldots, \Gamma_{n}$ of class $C^{3}$. The set $\Omega$ may have a finite number of holes (obstacles), i.e., it does not have to be simply connected. Let $\ell_{i}$ be the length of $\Gamma_{i}$, and consider the parametrization $\gamma_{i}:\left[0, \ell_{i}\right] \rightarrow \mathbb{R}^{2}$ of $\Gamma_{i}$ by arc-length with the property that the interior of $\Omega$ remains on the left of the tangent vector $\gamma_{i}^{\prime}(s)$ for $s \in\left[0, \ell_{i}\right]$. We assume that the curvature of each $\Gamma_{i}$ computed with respect to the parametrization $\gamma_{i}$ is either strictly negative, or strictly positive, or identically zero at every point of $\Gamma_{i}$ (even at its boundary points). If $\Gamma_{i}$ has zero curvature, then it is just a straight segment. We also assume that $\Gamma_{i} \cap \Gamma_{j} \subset \partial \Gamma_{i} \cap \partial \Gamma_{j}$ for $i \neq j$.

The sets $\Gamma_{1}, \ldots, \Gamma_{n}$ are called the components of $\partial \Omega$. The union of all components with positive curvature (focusing components), negative curvature (dispersing components) and zero curvature (flat components) are denoted by $\Gamma^{+}, \Gamma^{-}$and $\Gamma^{0}$, respectively. A point of $\bigcup_{i=1}^{n} \partial \Gamma_{i}$ is called a corner of $\Omega$.
2.2. Billiard phase space. For each $i=1, \ldots, n$, define

$$
M_{i}=\left[0, \ell_{i}\right] \times[0, \pi]
$$

with the elements $(0, \theta)$ and $\left(\ell_{i}, \theta\right)$ identified if $\Gamma_{i}$ is a closed curve. Hence $M_{i}$ is either a rectangle or a cylinder. The billiard phase space $M$ is the disjoint union of $M_{1}, \ldots, M_{n}$. An element $x \in M$ specifies a unit vector of $\mathbb{R}^{2}$ with base point at $s \in \Gamma_{i}$ and forming an angle $\theta$ with $\gamma_{i}^{\prime}(s)$. It corresponds to the position and the velocity of the point-particle immediately after a collision with $\partial \Omega$. For this reason, the elements ${ }^{1}$ of $M$ are called collisions. We denote by $M^{+}, M^{-}, M^{0}$ the subsets of $M$ obtained by taking the disjoint unions of sets $M_{i}$ with $\Gamma_{i}$ belonging to $\Gamma^{+}, \Gamma^{-}$and $\Gamma^{0}$, respectively.

[^1]The set $M$ is a smooth manifold with boundary $\partial M=\bigsqcup_{i=1}^{n} \partial M_{i}$. We equip $M$ with the Riemannian metric $g=\left\{g_{x}\right\}_{x \in M}$ and the symplectic form $\omega=\left\{\omega_{x}\right\}_{x \in M}$ given by $g_{x}=d s^{2}+d \theta^{2}$ and $\omega_{x}=\sin \theta(x) d s \wedge d \theta$. The norm generated by $g$ is denoted by $\|\cdot\|$. The Riemannian metric $g$ induces in the usual way a distance $d$ on each $M_{i}$, which can be extended to the entire set $M$ by setting $d(x, y)=\bar{d}$ whenever $x \in M_{i}$ and $y \in M_{j}$ with $i \neq j$, where $\bar{d}>0$ is a sufficiently large constant depending on $\partial \Omega$. Let $m$ be the volume generated by $g$. Then $\mu=(2 \ell)^{-1} \sin \theta(x) m$ is the probability measure generated by $\omega$ with $\ell:=\sum_{i=1}^{n} \ell_{i}$ being the total length of $\partial \Omega$.

Finally, we observe that there exists a natural involution $\mathcal{I}: M \rightarrow M$ given by $\mathcal{I}(s, \theta)=(s, \pi-\theta)$ whenever $(s, \theta) \in M_{i}$.
2.3. Billiard map and singular sets. Given $x=(s, \theta) \in M_{i}$, let $q(x)=\gamma_{i}(s) \in \Gamma_{i}$, and let $v(x)$ be the unit vector of $\mathbb{R}^{2}$ forming an angle $\theta$ with $\gamma_{i}^{\prime}(s)$. Let $\rho(x)=\{\tau>0:(q(x), q(x)+\tau v(x)) \subset \Omega\}$, where $(q(x), q(x)+\tau v(x))$ is the open segment $\{q(x)+t v(x): 0<t<\tau(x)\}$. Then, define $t(x)=0$ if $\rho(x)$ is empty, and $t(x)=\sup \rho(x)$ otherwise.

Now, define

$$
q_{1}(x)=q(x)+t(x) v(x),
$$

and

$$
M^{\prime}=\left\{x \in M: q_{1}(x) \text { is not a corner of } \Omega\right\} .
$$

If $x \in M^{\prime}$, then there exists a unique $i$ such that $q_{1}(x)$ belongs to the interior of $\Gamma_{i}$, and we set $s_{1}(x)=\gamma_{i}^{-1}\left(q_{1}(x)\right)$. Now, let

$$
v_{1}(x)=-v(x)+2\left\langle\gamma_{i}^{\prime}\left(s_{1}(x)\right), v(x)\right\rangle \gamma_{i}^{\prime}\left(s_{1}(x)\right),
$$

where $\langle\cdot, \cdot\rangle$ denotes the usual scalar product in $\mathbb{R}^{2}$. In other words, $v_{1}(x)$ is the unit vector obtained by reflecting $v(x)$ about the direction $\gamma_{i}^{\prime}\left(s_{1}(x)\right)$. Let $\theta_{1}(x) \in[0, \pi]$ be the oriented angle between $\gamma_{i}^{\prime}\left(s_{1}(x)\right)$ and $v_{1}(x)$. The billiard map for the domain $\Omega$ is the transformation $T: M^{\prime} \rightarrow M$ given by

$$
T x=\left(s_{1}(x), \theta_{1}(x)\right), \quad x \in M^{\prime} .
$$

The regularity (continuity, differentiability, etc.) of $T$ depends on the regularity of the boundary of $\Omega$. To clarify this point, we define:

$$
\begin{aligned}
& A_{1}=\{x \in M: q(x) \text { is a corner of } \Omega\}, \\
& A_{2}=\{x \in M: \theta(x) \in\{0, \pi\}\}, \\
& A_{3}=\left\{x \in M: q_{1}(x) \text { is a corner of } \Omega\right\}, \\
& A_{4}=\left\{x \in M^{\prime}: T x \in A_{2}\right\} .
\end{aligned}
$$

Let $S_{1}^{+}=A_{3} \cup A_{4}$, and let $S_{1}^{-}=\mathcal{I}\left(S_{1}^{+}\right)$. Both $S_{1}^{+}$and $S_{1}^{-}$are compact sets. Next, define

$$
R_{1}^{+}=\partial M \cup S_{1}^{+},
$$

and for all $j \geq 1$, define iteratively

$$
R_{j+1}^{+}=R_{j}^{+} \cup T^{-1} R_{j}^{+} \quad \text { and } \quad R_{j}^{-}=\mathcal{I}\left(R_{j}^{+}\right)
$$

The billiard map $T$ is a local $C^{k-1}$ diffeomorphism at $x \in M \backslash R_{1}^{+}[20$, Theorem 4.1]. Thus, the points where $T$ is not continuous or more generally is not $C^{k-1}$ are contained in $R_{1}^{+}$. Analogously, $R_{j}^{+}$(resp. $R_{j}^{-}$) contains the points where $T^{j}$ (resp. $T^{-j}$ ) is not $C^{k-1}$. We call $R_{j}^{+}$ (resp. $R_{j}^{-}$the singular set of $T^{j}$ (resp. $T^{-j}$ ).

The map $T: M \backslash R_{1}^{+} \rightarrow M \backslash R_{1}^{-}$is a $C^{k-1}$ diffeomorphism preserving the symplectic form $\omega$ and the probability measure $\mu$ provided that $k>1$ (see [20, Corollaries 4.1 and 4.4, Part V]). Under proper conditions on the components $\Gamma_{1}, \ldots, \Gamma_{n}$, which are satisfied by the billiards considered in this paper ${ }^{2}$, the sets $R_{j}^{+}$and $R_{j}^{-}$are unions of finitely many arcs of class $C^{2}$ [15]. Hence, $\mu\left(R_{j}^{+}\right)=\mu\left(R_{j}^{-}\right)=0$ for every $j \geq 1$. Finally, we observe that $T$ is time-reversible, which means that $\mathcal{I} \circ T=T^{-1} \circ \mathcal{I}$ on $M \backslash R_{1}^{+}$.

## 3. Focusing times and focusing components

We now introduce the concept of focusing times. This notion is borrowed from geometrical optics, and permits to obtain an intuitive description of the action of the derivative of the billiard map on the projective line.
3.1. Focusing times. Let $x \in M$ and $u \in T_{x} M \backslash\{0\}$. There exists a differentiable curve $\varphi:[-\delta, \delta] \rightarrow M$ with $\delta>0$ such that $\varphi(0)=x$ and $\varphi^{\prime}(0)=u$. This curve defines a 1-parameter family of lines $\gamma_{u}$ in $\mathbb{R}^{2}$, which is a differentiable variation of the line $l(x)$ through $q(x)$ and parallel to $v(x)$. In linear approximation, all lines of $\gamma_{u}$ focus at the point $q^{+} \in l(x)$. If $\gamma_{u}$ consists in linear approximation of parallel lines, then $q^{+}=\infty$. The point $q^{+}$depends only on $u$, and not on the curve $\varphi$. By reflecting the lines of $\gamma_{u}$ about the boundary of $\Omega$, we obtain another differentiable 1-parameter family of lines, which is generated by the curve $\mathcal{I}(\varphi)$. Denote by $q^{-}$the point where the lines of this family focus in linear approximation.

The forward focusing time $\tau^{+}(x, u)$ of the vector $u$ is equal to length of ( $\left.q^{+}-q(x)\right)$ multiplied by 1 or -1 depending on whether the vectors $\left(q^{+}-q(x)\right)$ and $v(x)$ have the same direction or opposite directions. The backward focusing time $\tau^{-}(x, u)$ of the vector $u$ is defined similarly with $q^{+}$and $v(x)$ replaced by $q^{-}$and $v(\mathcal{I}(x))$, respectively. Note that $\tau^{+}(x, u)$ and $\tau^{-}(x, u)$ may be negative numbers.

Define $\kappa_{i}(x)$ to be the curvature of $\Gamma_{i}$ for every $x=(s, \theta) \in M_{i}$ such that $s \in\left(0, \ell_{i}\right)$. Then extend $\kappa_{i}$ to the entire $M_{i}$ by continuity. Finally, let $\kappa: M \rightarrow \mathbb{R}$ be the function given by $\kappa(x)=\kappa_{i}(x)$ if $x \in M_{i}$.

[^2]Let $x \in M$ and $u \in T_{x} M \backslash\{0\}$. Write $u=\left(u_{s}, u_{\theta}\right) \in \mathbb{R}^{2}$ in coordinates $(s, \theta)$. Let $m(u)=u_{\theta} / u_{s} \in \mathbb{R} \cup\{\infty\}$. A straightforward computation (for example, see [32, Section 2]) shows that

$$
\tau^{ \pm}(x, u)= \begin{cases}\frac{\sin \theta(x)}{\kappa(x) \pm m(u)} & \text { if } m(u) \neq \mp \kappa(x)  \tag{1}\\ \infty & \text { otherwise }\end{cases}
$$

Let $d(x)=\sin \theta(x) / \kappa(x)$. From (1), one can easily derive the well known Mirror Equation of geometrical optics, relating the focusing times $\tau^{+}(x, u)$ and $\tau^{-}(x, u)$ :

$$
\begin{equation*}
\frac{1}{\tau^{+}(x, u)}+\frac{1}{\tau^{-}(x, u)}=\frac{2}{d(x)} . \tag{2}
\end{equation*}
$$

Definition 3.1. Let $x \in M \backslash R_{k}^{+}$for some $k \in \mathbb{N}$. Then define $\tau_{k}^{ \pm}(x, u)=\tau^{ \pm}\left(T^{k} x, D_{x} T^{k} u\right)$ for all $u \in T_{x} M \backslash\{0\}$.
3.2. Focusing components. In this paper, we assume that the focusing components of $\partial \Omega$ are arcs of the type introduced by Donnay [16], which for simplicity will be called Donnay arcs.

The first property that a Donnay arc $\Gamma_{i}$ must satisfy is the following:

$$
\begin{equation*}
\int_{0}^{\ell_{i}} \kappa(s(x)) d s \leq \pi . \tag{3}
\end{equation*}
$$

This condition amounts to saying that the tangents to $\Gamma_{i}$ at its endpoints form an angle that is not larger than $\pi$. In particular, no Donnay arc can be a closed curve. Hence, no billiard orbit is trapped by $\Gamma_{i}$, except possibly the periodic orbit of period 2 whose trajectory coincides with the segment joining the endpoints of $\Gamma_{i}[16$, Lemma 1.1].
Definition 3.2. Define $n: M^{+} \rightarrow\{0,1,2, \ldots\} \cup\{+\infty\}$ by
$n(x)= \begin{cases}\sup \left\{j \geq 0: T^{k} x \in M_{i} \quad \forall 0 \leq k \leq j\right\}, & x \in M_{i} \backslash A_{2} \subset M^{+}, \\ +\infty, & x \in M^{+} \cap A_{2} .\end{cases}$
The number $n(x)$ is the number of consecutive collisions of $x$ with $\Gamma_{i}$ before either leaving it or hitting one of its endpoints.
Definition 3.3. For every $\Gamma_{i} \subset \Gamma^{+}$, define

$$
E_{i}=\left\{x \in M_{i}: n(\mathcal{I}(x))=0\right\} .
$$

Also, define $E^{+}=\bigcup_{i: \Gamma_{i} \subset \Gamma^{+}} E_{i}$.
Equivalently, $x \in E_{i}$ if and only if $\mathcal{I}(x)$ either leaves $\Gamma_{i}$ or hits one of the endpoints of $\Gamma_{i}$. Note that if $x \in E_{i}$, then $\left\{x, T x, \ldots, T^{n(x)} x\right\} \subset M_{i}$ is the maximal sequence of consecutive collisions with $\Gamma_{i}$ that cannot be extended forward or backward.

Definition 3.4. Let $x \in E_{i}$, and denote by $u_{x} \in T_{x} M$ the tangent vector such that $\tau^{-}\left(x, u_{x}\right)=\infty$ or equivalently $m(x)=\kappa(x)$. In other words, the variation associated to $u_{x}$ consists of lines parallel to $v(x)$.

Definition 3.5. We say that $x \in E_{i}$ is focused by $\Gamma_{i}$ if
(1) $0<\tau_{i}^{+}\left(x, u_{x}\right)<t\left(T^{i} x\right)$ for $0 \leq i<n(x)$,
(2) $0<\tau_{n(x)}^{+}\left(x, u_{x}\right)<+\infty$.

The concept of a focused collision is key in the definition of a Donnay arc. In words, $x$ is focused by $\Gamma_{i}$ if the infinitesimal family of parallel trajectories focuses between every two consecutive collisions with $\Gamma_{i}$ and after the last one.

Definition 3.6. A focusing component $\Gamma_{i}$ is a Donnay arc if i) $\Gamma_{i}$ is of class $C^{\infty}$, ii) $\Gamma_{i}$ satisfies (3), and iii) if every $x \in E_{i}$ is focused by $\Gamma_{i}$.

The definition of a Donnay arc is similar to that of an absolutely focusing arc introduced by Bunimovich [3, 5]. The relation between the two definitions is discussed in [5].

Billiards in convex regions bounded by Donnay arcs connected by straight segments sufficiently long have non-zero Lyapunov exponents almost everywhere [16, Theorem 2]. In this paper, we consider billiard domains that are more general than those considered by Donnay: our billiard domains are not necessarily convex, and their boundaries are allowed to have components with negative curvature.

The key property of Donnay arcs used in the construction of hyperbolic billiards is the uniform boundedness of the focusing time $\tau_{n(x)}^{+}\left(x, u_{x}\right)$ for $x \in E_{i}$ [16, Theorem 4.4]. For this property to hold, the arc does not have to be smooth as required by Donnay. Indeed, Wojtkowski and Markarian found $C^{4}$ curves satisfying that property [23, 32].

Examples of Donnay arcs are arcs of circles, arcs of cardioids, arcs of logarithmic spirals and elliptical arcs. An example of a Donnay arc that is not one of the arcs discovered by Wojtkowski and Markarian is the half-ellipse $\left\{(x, y) \in \mathbb{R}^{2}: x^{2} / a^{2}+y^{2} / b^{2}=1\right.$ and $\left.x \geq 0\right\}$ with $a / b<\sqrt{2}$ [16, Theorem 7.1].

We mention two interesting facts concerning the existence and robustness under perturbation of Donnay arcs. The first fact is that given a smooth arc $\Gamma$ with positive curvature, every sufficiently short arc contained in $\Gamma$ is a Donnay arc [16, Theorem 3]. A similar conclusion when $\Gamma$ is only of class $C^{4}$, but satisfies also the condition $d^{2}\left(\kappa^{-1 / 3}\right) / d s^{2}>0$ was obtained by Markarian [23]. The second fact we want to mention is that given a Donnay arc $\Gamma$, every sufficiently small perturbation of $\Gamma$ in the $C^{6}$ topology is still a Donnay arc [16, Theorem 4].

## 4. Cone fields for billiards

In this section, we recall the notions of an invariant cone field and a monotone quadratic form. We restrict ourselves to the 2-dimensional setting, since we are dealing only with planar billiards in this paper. Next, we introduce a specific family of cone fields for planar billiards.

In Section 5, we will show that these cone fields are eventually strictly invariant if the billiards satisfy certain geometric conditions. The main reference for Subsections 4.1 and 4.2 is [22].
4.1. Cone fields. Let $V$ be a 2 -dimensional vector space with a symplectic form $\alpha$. Given two linearly independent vectors $X_{1}$ and $X_{2}$ of $V$, we can write $v=v_{1}+v_{2}$ for every $v \in V$ with $v_{i} \in X_{i}$ for $i=1,2$. We call the cone generated by $X_{1}$ and $X_{2}$ the set

$$
C\left(X_{1}, X_{2}\right):=\left\{v \in V: \alpha\left(v_{1}, v_{2}\right) \geq 0\right\} .
$$

The cone $C^{\prime}\left(X_{1}, X_{2}\right):=C\left(X_{2}, X_{1}\right)$ is called the complementary cone of $C\left(X_{1}, X_{2}\right)$. The set $\operatorname{int} C\left(X_{1}, X_{2}\right):=\left\{v \in V: \alpha\left(v_{1}, v_{2}\right)>0\right\}$ is called the interior of $C\left(X_{1}, X_{2}\right)$. To our knowledge, the description of a cone using symplectic forms first appeared in [31].

Now, consider a billiard in a domain $\Omega \subset \mathbb{R}^{2}$ with phase space $M$, billiard map $T$ and the symplectic form $\omega=\left\{\omega_{x}\right\}_{x \in M}$ as in Section 2. Let $U$ be an open set of $M$, and let $X_{1}$ and $X_{2}$ be two measurable vector fields on $U$ such that $X_{1}(x)$ and $X_{2}(x)$ are linearly independent for all $x \in U$. The cone field $C$ on $U \subset M$ generated by $X_{1}$ and $X_{2}$ is the family of cones $\{C(x)\}_{x \in U}$ with

$$
C(x):=C\left(X_{1}(x), X_{2}(x)\right)=\left\{u \in T_{x} M: \omega_{x}\left(u_{1}, u_{2}\right) \geq 0\right\} \subset T_{x} M
$$

for every $x \in U$, and it is denoted by $(U, C)$.
The cone field $(U, C)$ is called continuous if $X_{1}$ and $X_{2}$ are continuous. The cone field $(U, C)$ is called invariant (resp. strictly invariant) if $x \in U$ and $T^{k} x \in U$ with $k>0$ implies that $D_{x} T^{k} C(x) \subset C\left(T^{k} x\right)$ (resp. $\left.D_{x} T^{k} C(x) \subset \operatorname{int} C\left(T^{k} x\right) \cup\{0\}\right)$. The cone field $(U, C)$ is called eventually strictly invariant if it is invariant, and for a.e. $x \in U$, there exists an integer $k(x)>0$ such that $T^{k(x)} x \in U$ and $D_{x} T^{k(x)} C(x) \subset$ $\operatorname{int} C\left(T^{k(x)}(x)\right) \cup\{0\}$.

A crucial property of the 2-dimensional cone $C(x)$ is that it can be identified with a closed interval of the projective space $\mathbb{P}\left(T_{x} M\right)$. Therefore, since $m, \tau^{-}, \tau^{+}$are all projective coordinates, each of the following provides an alterative definition of the cone field $(U, C)$ : for all $x \in U$,

$$
\begin{aligned}
& C(x)=\left\{u \in T_{x} M \backslash\{0\}: m(u) \in I(x)\right\} \cup\{0\}, \\
& C(x)=\left\{u \in T_{x} M \backslash\{0\}: \tau^{+}(x, u) \in I_{+}(x)\right\} \cup\{0\}, \\
& C(x)=\left\{u \in T_{x} M \backslash\{0\}: \tau^{-}(x, u) \in I_{-}(x)\right\} \cup\{0\}
\end{aligned}
$$

with $I(x), I_{-}(x), I_{+}(x)$ being proper closed intervals of the extended real line $\hat{\mathbb{R}}$.
4.2. Quadratic forms. Consider a cone field $(U, C)$ generated by the vector fields $X_{1}$ and $X_{2}$. The quadratic form $Q_{C}=\left\{Q_{C}(x, \cdot): x \in U\right\}$ associated to $(U, C)$ is defined by

$$
Q_{C}(x, u)=\omega_{x}\left(u_{1}, u_{2}\right) \quad \text { for } x \in U \text { and } u \in T_{x} M .
$$

The form $Q_{C}$ is monotone (resp. strictly monotone) if $Q_{C}\left(T^{k} x, D_{x} T^{k} u\right) \geq$ $Q_{C}(x, u)$ (resp. $\left.Q_{C}\left(T^{k} x, D_{x} T^{k} u\right)>Q_{C}(x, u)\right)$ for all $u \in T_{x} M \backslash\{0\}$ whenever $x \in U$ and $T^{k} x \in U$ with $k>0$. The form $Q_{C}$ is eventually strictly monotone if it is monotone, and for a.e. $x \in U$, there exists an integer $k(x)>0$ such that $Q_{C}\left(T^{k(x)} x, D_{x} T^{k(x)} u\right)>Q_{C}(x, u)$ for all $u \in T_{x} M \backslash\{0\}$.

Following [22], to measure the expansion generated by the action of $D_{x} T^{k}$ with $k>0$ on vectors of $C_{x}$, we define

$$
\sigma_{C}\left(D_{x} T^{k}\right)=\inf _{u \in \operatorname{int} C(x)} \sqrt{\frac{Q_{C}\left(T^{k} x, D_{x} T^{k} u\right)}{Q_{C}(x, u)}}
$$

and

$$
\sigma_{C}^{*}\left(D_{x} T^{k}\right)=\inf _{u \in \operatorname{int} C(x)} \frac{\sqrt{Q_{C}\left(T^{k} x, D_{x} T^{k} u\right)}}{\|u\|}
$$

We denote by $\left(U, C^{\prime}\right)$ the family of cones $\left\{C^{\prime}(x)\right\}_{x \in U}$ with $C^{\prime}(x)$ being the complementary cone of $C(x)$. We observe that $C$ is invariant (resp. strictly invariant) if and only if $C^{\prime}$ is invariant (resp. strictly invariant) with respect to $T^{-1}$. The relation between the expansion of $D_{x} T^{k}$ on $C$ and the expansion of $D_{T^{k} x} T^{-k}$ on $C^{\prime}$ is given by $\sigma_{C^{\prime}}\left(D_{T^{k} x} T^{-k}\right)=$ $\sigma_{C}\left(D_{x} T^{k}\right)$.

The cone field ( $U, C$ ) is invariant (resp. strictly invariant) if and only if the quadratic form $Q_{C}$ is monotone (resp. strictly monotone), and $(U, C)$ is eventually strictly invariant if and only if $Q_{C}$ is eventually strictly monotone. Furthermore, $D_{x} T^{k} C(x) \subset C\left(T^{k} x\right)$ (resp. $\left.D_{x} T^{k} C(x) \subset \operatorname{int} C\left(T^{k} x\right) \cup\{0\}\right)$ for $x \in U$ such that $T^{k} x \in U$ with $k>0$ is equivalent to $\sigma_{C}\left(D_{x} T^{k}\right) \geq 1$ (resp. $\sigma_{C}\left(D_{x} T^{k}\right)>1$ ).
4.3. Cone fields for billiards. Let $E^{-}=\bigcup_{i: \Gamma_{i} \subset \Gamma^{-}} M_{i}$, and let $E=$ $E^{+} \cup E^{-}$(for the definition of $E^{+}$, see Definition 3.3). In this section, we introduce a family of continuous cone fields for billiards. The cones are defined separately for focusing and dispersing components, and to make their geometrical meaning more transparent, several equivalent definitions are provided. Rather than considering a single continuous cone field, as it is common in the literature on hyperbolic billiards, we consider a family of continuous cone fields $\left\{\left(U_{x}, C_{x}\right): x \in E\right\}$. The reason for using $\left\{\left(U_{x}, C_{x}\right): x \in E\right\}$ instead of a single continuous cone field on $E$ is that for most of the billiards considered in this paper, we do not know whether such a cone field exists ${ }^{3}$.

As a consequence of this choice, we will be able to apply the Local Ergodic Theorem (Theorem A.13) only to $U_{x}$ with $x \in E$. This is not a serious limitation, because the hypotheses imposed on our billiards imply that almost every element of $M$ visits $E$, and so knowing that

[^3]each $U_{x}$ belongs $(\bmod 0)$ to an ergodic component of $T$ suffices to conclude that the same holds true for almost every point of $M$.
4.3.1. Dispersing components. Let $x \in M_{i} \subset M^{-}$for some $i$. Then, define $U_{x}=M_{i}$, and for every $y \in U_{x}$,
$$
C_{x}(y)=\left\{u \in T_{y} M_{i} \backslash\{0\}: m(u) \leq \kappa(y)\right\} \cup\{0\} .
$$

The resulting cone field $\left(U_{x}, C_{x}\right)$ is clearly continuous. Each cone $C_{x}(y)$ consists of tangent vectors focusing inside the disk tangent to $\Gamma_{i}$ at $q(y)$ of radius $1 /(4|\kappa(y)|)$. This can be easily checked by using (1): if $u \in C_{x}(y) \backslash\{0\}$, then $d(y) \leq 2 \tau^{+}(y, u) \leq 0$ and $0 \leq \tau^{-}(y, u) \leq+\infty$.
4.3.2. Focusing components. To define $\left(U_{x}, C_{x}\right)$ for every $x \in E^{+}$, we first recall a result of Donnay.

For every $x \in M^{+}$and every real number $g<\kappa(x)$, let

$$
\mathcal{D}(x, g)=\left\{u \in T_{x} M \backslash\{0\}: g \leq m(u) \leq \kappa(x)\right\}
$$

Theorem 4.1. Suppose that $\Gamma_{i}$ is a Donnay arc. Then there exist positive constants $a_{i}, m_{i}, \theta_{i}, t_{i}^{-}, t_{i}^{+}$, and for every $x \in E_{i}$, there exist a neighborhood $U_{x} \subset M_{i} \backslash A_{2}$ of $x$ and a continuous function $g_{x}: U_{x} \rightarrow \mathbb{R}$ with $\left|g_{x}\right|<\left.\kappa\right|_{U_{x}}$ such that if $x \in E_{i}, y \in U_{x}, j \in\{0, \ldots, n(y)\}, y_{j}:=$ $T^{j} y, u \in \mathcal{D}\left(y, g_{x}(y)\right)$ and $u_{j}:=D_{y} T^{j} u$, then
(1) $-\kappa\left(y_{j}\right)+m_{i} \leq m\left(u_{j}\right) \leq \kappa\left(y_{j}\right)$,
(2) $\min \{\theta(y), \pi-\theta(y)\}<\theta_{i} \Longrightarrow\left|m\left(u_{j}\right)\right| \leq a_{i} \cdot \min \{\theta(y), \pi-\theta(y)\}$,
(3) $j<n(y) \Longrightarrow d\left(y_{j}\right) \leq 2 \tau^{+}\left(y_{j}, u_{j}\right)<2 t\left(y_{j}\right)-d\left(T y_{j}\right)$,
and if $x \in E_{i}$ and $y \in U_{x}$, then
(4) $\inf _{u \in \mathcal{D}\left(y, g_{x}(y)\right)} \tau^{-}(y, u) \leq t_{i}^{-}$,
(5) $\sup _{u \in \mathcal{D}\left(y, g_{x}(y)\right)} \tau_{n(y)}^{+}(y, u) \leq t_{i}^{+}$.

Proof. For the proof of Parts (1) and (2), see the proofs of [16, Theorems 4.4 and 5.6], whereas for the proof of Parts (3)-(5), see [16, Proposition 4.1 and Theorem 4.4].

Theorem 4.1 has the following geometrical interpretation. Part (1) states that the iterates along consecutive collisions with $\Gamma_{i}$ of tangent vectors with initial slope between $g_{x}(y)$ and $\kappa(y)$ have slopes uniformly bounded for all $x \in E_{i}$ and all $y \in U_{x}$. This conclusion is strengthened in Part (2) provided that $\theta(y)$ is uniformly close to 0 or $\pi$. In particular, in this case, the iterates of the vectors remain uniformly close to the horizontal direction. Parts (3)-(5) implies the every $y \in U_{x}$ is focused (c.f. 3.5), and that the backward and forward focusing times of the tangent vectors considered are uniformly bounded by constants depending only on $\Gamma_{i}$. In fact, the constants $t_{i}^{-}$and $t_{i}^{+}$are continuous functions of $\Gamma_{i}$ with respect to the $C^{6}$ topology [16, Theorem 4.4].

Suppose that $x \in E^{+}$. Let $U_{x}$ be the neighborhood of $x$ and $g_{x}$ be the continuous function as in Theorem 4.1. For every $y \in U_{x}$, define

$$
C_{x}(y)=\mathcal{D}\left(y, g_{x}(y)\right) \cup\{0\} .
$$

Remark 4.2. The following are alternative descriptions of the cone $C_{x}(y)$ for every $x \in E$ in terms of the projective coordinates $\tau^{+}$and $\tau^{-}$:

$$
\begin{aligned}
C_{x}(y) & =\left\{u \in T_{y} M_{i} \backslash\{0\}: d(y) / 2 \leq \tau^{+}(y, u) \leq G_{x}^{+}(y)\right\} \cup\{0\} \\
& =\left\{u \in T_{y} M_{i} \backslash\{0\}: G_{x}^{-}(y) \leq \tau^{-}(y, u) \leq+\infty\right\} \cup\{0\},
\end{aligned}
$$

where $G_{x}^{ \pm}(y):=\sin \theta(y) /\left(\kappa(y) \pm g_{x}(y)\right)$ for all $y \in U_{x}$ if $x \in E^{+}$, and $G_{x}^{ \pm}(y):=0$ for all $y \in U_{x}$ if $x \in E^{-}$. Note that for $x \in E^{+}$, the numbers $G_{x}^{+}(y)$ and $G_{x}^{-}(y)$ are, respectively, the forward and backward focusing times of the vectors with slope equal to $g_{x}(y)$.

## 5. Results

In this section, we give a detailed description of the billiards that we want to study, and formulate the main results of the paper.

The billiards in question are characterized by four conditions called B1-B4. It is well known that B1-B3 are sufficient to guarantee the hyperbolicity of the billiard map $T$ (see Proposition 5.3). This together with the Spectral Theorem implies that $T$ has at most countably many ergodic components of positive measure with respect to $\mu$, with each ergodic component further decomposed into finitely many Bernoulli components cyclically permuted by $T$ (see Theorem A.6).

The main result of this paper is the following: if in addition to B1B3, a billiard satisfies also Condition B4, then there exists a measurable set $H \subset M$ of full measure such that for every $x \in H$, there is a neighborhood of $x$ in $M$ contained up to a set of zero measure in a single Bernoulli component of $T$ (see Theorem 5.6). This result implies immediately that every Bernoulli component of $T$ is open up to a set of zero measure. Results of this type are often called Local Ergodic Theorems.

Local ergodicity alone is not enough to conclude that $T$ is Bernoulli. This is obtained by imposing on $T$ some extra conditions. It is not an easy task to formulate these conditions for the generality of the billiards considered in this paper. Therefore, rather than trying to formulate the optimal condition for the Bernoulli property of hyperbolic billiards, we limit ourselves here to give a simple condition, called B5, that yields the Bernoulli property of interesting subclasses of hyperbolic billiards (i.e., billiards with domains without straight boundaries, i.e., $\Gamma^{0}=\emptyset$ ). It is unfortunate that B5 does not hold for Donnay billiards. Nevertheless, in Section 8, we prove that these billiards and some generalizations (billiards with pockets and bumps) are Bernoulli, using a proof that does not require B5.
5.1. Important sets. Next, we introduce several sets involved in the formulation of Conditions B1-B5. Recall that $R_{k}^{ \pm}$are the singular sets defined in Section 2. Define

- $R_{\infty}^{ \pm}=\bigcup_{k \geq 1} R_{k}^{ \pm}$,
- $R=R_{\infty}^{-} \cap R_{\infty}^{+}$,
- $N^{ \pm}=\left\{x \in M \backslash R_{\infty}^{ \pm}: \exists k>0\right.$ s.t. $T^{( \pm) n} x \in M_{0}$ for all $\left.n \geq k\right\}$,
- $N=N^{-} \cap N^{+}$,
- $N^{\prime}=\left(R_{\infty}^{-} \cap N^{+}\right) \cup\left(R_{\infty}^{+} \cap N^{-}\right)$,
- $H=M \backslash\left(R \cup N \cup N^{\prime}\right)$.

The geometric meaning of these sets is the following: $R_{\infty}^{+}$(resp. $R_{\infty}^{-}$) is the set of collisions with finite positive (resp. negative) semiorbit; $R$ is the set of collisions with finite orbit; $N^{+}$(resp. $N^{-}$) is the set of collisions with positive (resp. negative) semi-orbit visiting eventually only flat components of $\partial \Omega ; N$ is the set of collisions with both semi-orbits visiting eventually only flat components of $\partial \Omega ; N^{\prime}$ is the set of collisions with one semi-orbit being finite and the other semi-orbit visiting eventually only flat components of $\partial \Omega ; H$ is the set of collisions with one semi-orbit visiting the curved components of $\partial \Omega$ infinitely many times.
5.2. Hyperbolic billiards. We are ready to formulate Conditions B1B5.

B1 (Non-polygonal domain): The domain $\Omega$ is not a polygon, and its boundary components can only be of the following type: straight segments, dispersing arcs of class $C^{3}$ and focusing arcs.

B2 (Distance between boundary components): For each curved component $\Gamma_{i}$, we define $\lambda_{i}^{ \pm}=0$ if $\Gamma_{i}$ is dispersing, and $\lambda_{i}^{ \pm}=t_{i}^{ \pm}$with $t_{i}^{ \pm}$as in Theorem 4.1 if $\Gamma_{i}$ is focusing. Given two curved components $\Gamma_{i}$ and $\Gamma_{j}$, denote by $t_{i j} \geq 0$ the infimum of the Euclidean length of all finite billiard orbits $\left\{x_{0}, \ldots, x_{n}\right\}$ with $n>0$ such that $x_{0} \in M_{i}$ and $x_{n} \in M_{j}$. We assume that
(1) there exists $\lambda>0$ such that if $\Gamma_{i}$ or $\Gamma_{j}$ is focusing, then

$$
t_{i j} \geq \lambda_{i}^{-}+\lambda_{j}^{+}+\lambda
$$

(2) the distance between each focusing component $\Gamma_{i}$ and the set of corners of $\partial \Omega$ formed by two straight segments is greater than $\lambda_{i}^{-}$.

B3 (Neutral orbits): We assume that $\mu\left(N^{-}\right)=0$.
B4 (Singular-Neutral orbits): We assume $m_{-}\left(S_{1}^{-} \cap N^{+}\right)=0$, where $m_{-}$is the measure induced by the Riemann metric $g$ on $S_{1}^{-}$.

B5 (Connectedness): The set $H \cap M_{i}$ is connected for every component $\Gamma_{i}$ of $\partial \Omega$.

Condition B2 has a couple of obvious consequences for the geometry of $\Omega$ : i) the internal angle between a focusing component and an adjacent curved component is greater than $\pi$, and ii) the internal angle between a focusing component and an adjacent flat component is greater than $\pi / 2$. Also, note that Conditions B1-B4 allow $\partial \Omega$ to have cusps formed by two dispersing components or a dispersing and a flat component.
Remark 5.1. Since $N^{+}=\mathcal{I}\left(N^{-}\right)$and $S_{1}^{+} \cap N^{-}=\mathcal{I}\left(S_{1}^{-} \cap N^{+}\right)$, Conditions B3 and B4 imply that $\mu\left(N^{+}\right)=0$ and $m_{+}\left(S_{1}^{+} \cap N^{-}\right)=0$, where $m_{+}$is the measure induced by the Riemann metric $g$ on $S_{1}^{+}$.

From Conditions B1 and B2, it follows that cone field $\left\{\left(U_{x}, C_{x}\right)\right\}_{x \in E}$ is strictly invariant along a piece of an orbit connecting two elements of $E$. The next lemma is proved in [15, Lemma 5.2].

Lemma 5.2. Suppose that the billiard in $\Omega$ satisfies Conditions B1 and B2. Also, suppose that there exist $x_{1}, x_{2} \in E$ and $y \in E \cap U_{x_{1}} \backslash R_{k}^{+}$for some $k>0$ such that $T^{k} y \in E \cap U_{x_{2}}$. Then

$$
D_{y} T^{k} C_{x_{1}}(y) \subset \operatorname{int} C_{x_{2}}\left(T^{k} y\right) \cup\{0\}
$$

From the previous lemma, one obtains the hyperbolicity of the billiard map $T$ provided that Conditions B1-B3 are satisfied. This is well known fact, but we give its proof for completeness.
Proposition 5.3. If a billiard in a domain $\Omega$ satisfies Conditions B1B3, then the Lyapunov exponents of the map $T$ are non-zero a.e. on $M$.

Proof. Let $E^{\prime}$ be the subset of $E \backslash\left(R_{\infty}^{+} \cup N^{+}\right)$defined by

$$
E^{\prime}=\left\{x \in E \backslash\left(R_{\infty}^{+} \cup N^{+}\right): \exists n_{k} \nearrow+\infty \quad \text { s.t. } \quad T^{n_{k}} x \in V_{x} \quad \forall k>0\right\}
$$

where $\left(U_{x}, C_{x}\right)$ is the cone field associated to $x$. Since $\left(U_{x}, C_{x}\right)$ is strictly invariant for every $x \in E^{\prime}$ by Lemma 5.2, results of Wojtkowski [32] imply that the Lyapunov exponents of $T$ are non-vanishing at every point of $E^{\prime}$. By $\mu\left(R_{\infty}^{+}\right)=\mu\left(N^{+}\right)=0$ and the Poincaré Recurrence Theorem, we obtain that $\mu\left(E^{\prime}\right)=\mu(E)>0$. This fact together with $\mu\left(R_{\infty}^{+}\right)=\mu\left(N^{+}\right)=0$ gives that the orbit of a.e. point of $M$ visits $E^{\prime}$. Since the Lyapunov exponents are constant along orbits, we can finally conclude that the Lyapunov exponents of $T$ are non-vanishing at a.e. point of $M$.

Conditions B1-B3 are sufficient for the hyperbolicity of the map $T$. In fact, even part (2) of B2 can be dropped if we are only interested in the hyperbolicity of $T$. The extra Condition B4 is required to prove that $T$ is locally ergodic. This condition is related to the Sinai-Chernov Ansatz (c.f. Condition L3 of Theorem A.13). Note also that B3 is a necessary condition for the hyperbolicity of $T$.

Remark 5.4. We do not know whether or not, for a domain $\Omega$ satisfying B1 and B2, B3 or B4 is automatically satisfied. We also do not know whether B3 and B4 are independent. These questions are strictly related to the problem of understanding the distribution of orbits in polygonal billiards.

Lemma 5.5. We have $\mu(H)=1$ provided that B1 and B3 and are satisfied.

Proof. Since $\mu\left(R_{1}^{+}\right)=0$ for billiards satisfying B1 (see the end of Subsection 2.3), we trivially obtain $\mu(R)=\mu\left(N^{\prime}\right)=0$. From B3, we obtain immediately $\mu(N)=0$.
5.3. Main results. The central result of this paper is the following theorem. Its proof is given in Section 6. For the definition of a Bernoulli component of $T$, see Appendix A.

Theorem 5.6. If a billiard in a domain $\Omega$ satisfies Conditions B1B4, then every point of $H$ has a neighborhood contained (mod 0) in a Bernoulli component of $T$.

We now prove that the map $T$ is Bernoulli if it also satisfies Condition B5. As already explained in the introduction to this section, B5 applies only to a small subclass of billiards satisfying B1-B4 (see Theorem 8.7). We could have weakened considerably B5 for it to include many more hyperbolic billiards, but at the price of a much more technical formulation. Instead of attempting to give the weakest formulation of B5, we opted for a strong condition but with a simple formulation that allows for a relatively simple proof of the Bernoulli property for billiards.

Corollary 5.7. If a billiard in a domain $\Omega$ satisfies Conditions B1-B4, then every Bernoulli component of $T$ is open (mod 0).

Proof. Let $B$ be a Bernoulli component. Since $\mu(B)>0$, we have $\mu(B \cap H)>0$. Let $x \in B \cap H$, and let $U$ be the neighborhood of $x$ as in Theorem 5.6. The set $V:=\bigcup_{n \in \mathbb{Z}} T^{n} U$ is open. Moreover, since $V$ is invariant and contained $(\bmod 0)$ in $B$, it follows that $B=V(\bmod$ $0)$.

Corollary 5.8. If the billiard in a domain $\Omega$ satisfies Conditions B1B5, then the map $T$ is Bernoulli.

Proof. By Theorem 5.6, every point of $H$ has a neighborhood contained up to a set of zero measure in a Bernoulli component of $T$. The same is true for every connected component of $H$, and so for every $M_{i} \cap H$ such that $M_{i} \subset M^{-} \cup M^{+}$by the first part of Condition B5. Since $\mu(H)=1$ (see Remark 5.1), we conclude that every set $M_{i} \subset M^{-} \cup M^{+}$ is contained $(\bmod 0)$ in a single Bernoulli component of $T$.

We now show that if $\Gamma_{i}$ and $\Gamma_{j}$ intersect, then $M_{i}$ and $M_{j}$ are contained in the same Bernoulli component of $T$. First, note that $S_{1}^{-} \backslash\left(R_{\infty}^{+} \cup N^{+}\right)$is contained in $H$. Next, since $S_{1}^{-} \cap R_{k}^{+}$is finite for every $k>0$ (see [15, Propositions 6.17-6.19]), it follows that $S_{1}^{-} \cap R_{\infty}^{+}$is countable. This together with B4 implies that $m_{-}$-a.e. element of $S_{1}^{-}$is contained in $H$. Hence, if $p \in \Gamma_{i} \cap \Gamma_{j}$ is a corner of $\partial \Omega$, then we can find $x \in H \cap S_{1}^{-}$such that the ray emerging from $\mathcal{I}(x)$ is arbitrarily close to $\ell$, the line bisecting $p$. Now, let $U$ be the neighborhood of $x$ contained $(\bmod 0)$ in one Bernoulli component of $T$ as stated by Theorem 5.6. It is clear that $T^{-1} U$ is contained $(\bmod 0)$ in the same Bernoulli component. But, if $\mathcal{I}(x)$ is sufficiently close to $\ell$, then $M_{i} \cap T^{-1} U$ and $M_{j} \cap T^{-1} U$ are non-empty open sets so that $M_{i}$ and $M_{j}$ must belong $(\bmod 0)$ to the same Bernoulli component.

The previous conclusion implies that $q^{-1}(\Sigma)$ is contained $(\bmod 0)$ in a Bernoulli component for every connected component $\Sigma$ of $\partial \Omega$. Since $\Omega$ is connected, it follows that all sets $M_{i}$ belongs to the same Bernoulli component, i.e., $T$ is Bernoulli.

## 6. Local ergodicity

In this section, we prove Theorem 5.6 by applying Theorem A. 13 to the billiard map $T$. Theorem A. 13 is a version of a Local Ergodic Theorem for hyperbolic symplectomorphisms with singularities in arbitrary dimension [14] specialized to planar billiards. Because of its length, Theorem A. 13 is given in the Appendix together with all the definitions required for its formulation. Its main hypotheses are Conditions L1-L4. This section is devoted to the proof of Conditions L1-L3, whereas Section 7 is entirely devoted to the proof of Condition L4.

Proof of Theorem 5.6. The wanted conclusion follows at once by applying Theorem A. 13 to points of $H$. Accordingly, we show that each point of $H$ is sufficient (see Definition A.1) and satisfies Conditions L1-L4. This is achieved in Corollary 6.3, Propositions 6.4-6.6 and Proposition 7.30.
6.1. Sufficient points. We now prove that every point of $H$ is sufficient (see Definition A.1).

Given $x \in E$ and $k \in \mathbb{N}$ such that $T^{k} x \in E$, define

$$
\widetilde{\sigma}_{x}\left(D_{y} T^{k}\right)=\inf _{u \in \operatorname{int} C_{x}(y)} \sqrt{\frac{Q_{C_{T^{k} x}}\left(T^{k} y, D_{y} T^{k} u\right)}{Q_{C_{x}}(y, u)}}
$$

and

$$
\widetilde{\sigma}_{x}^{*}\left(D_{y} T^{k}\right)=\inf _{u \in \operatorname{int} C_{x}(y)} \frac{\sqrt{Q_{C_{T^{k} x}}\left(T^{k} y, D_{y} T^{k} u\right)}}{\|u\|}
$$

for every $y \in U_{x} \cap T^{-k} U_{T^{k} x}$. The quantity $\widetilde{\sigma}_{x}$ (resp. $\widetilde{\sigma}_{x}^{*}$ ) is a generalization of $\sigma_{C}$ (resp. $\sigma_{C}^{*}$ ), because if $U=U_{x}=U_{T_{k} x}$ and $C=C_{x}=C_{T^{k} x}$, then $\widetilde{\sigma}_{x}=\sigma_{C}\left(\right.$ resp. $\left.\widetilde{\sigma}_{x}^{*}=\sigma_{C}^{*}\right)$.

Under the previous hypotheses $x, T^{k} x \in E$, also define

$$
t\left(x, T^{k} x\right)=\sum_{i=0}^{m-1} t\left(T^{i} x\right)
$$

In other words, $t\left(x, T^{k} x\right)$ is the sum of the length of the segments $[q(x), q(T x)], \ldots,\left[q\left(T^{k-1} x\right), q\left(T^{k} x\right)\right]$. Furthermore, let $\lambda>0$ be the constant in Condition B2.

Lemma 6.1. There exists a constant c depending only on the billiard table $\Omega$ such that if $x \in E, T^{m} x \in E$ for some $m \in \mathbb{N}$ and $T^{i} x \notin E$ for every $0<i<m$, then

$$
\tilde{\sigma}_{x}\left(D_{x} T^{m}\right) \geq \sqrt{1+c \delta}+\sqrt{c \delta}
$$

where $\delta=\lambda$ if $x \in M^{+}$or $T^{m} x \in M^{+}$, and $\delta=t\left(x, T^{m} x\right)$ if $x \in M^{-}$ and $T^{m} x \in M^{-}$.

Proof. The hypotheses imply that if $x \in M^{+}$and $m>n(x)+1$, then $T^{i} x \in M^{0}$ for every $n(x)<i<m$. Similarly, if $x \in M^{-}$and $m>1$, then $T^{i} x \in M^{0}$ for every $1<i<m$. Define $\bar{m}=n(x)$ if $x \in M^{+}$, and $\bar{m}=0$ if $x \in M^{-}$.

The cones $C_{T^{m} x}\left(T^{m} x\right)$ and $D_{x} T^{m} C_{x}(x)$ can be written as follows

$$
\begin{aligned}
& C_{T^{m} x}\left(T^{m} x\right)=\left\{u \in T_{T^{m} x} M \backslash\{0\}: a \leq \tau^{-}\left(T^{m} x, u\right) \leq b\right\} \cup\{0\}, \\
& D_{x} T^{m} C_{x}(x)=\left\{u \in T_{T^{m} x} M \backslash\{0\}: \bar{a} \leq \tau^{-}\left(T^{m} x, u\right) \leq \bar{b}\right\} \cup\{0\},
\end{aligned}
$$

where

$$
\begin{aligned}
& a=G_{T^{m} x}^{-}\left(T^{m} x\right), \quad b=+\infty, \\
& \bar{a}=t\left(x, T^{m} x\right)-\sup _{u \in C_{x}(x) \backslash\{0\}} \tau_{\bar{m}}^{+}(x, u), \\
& \bar{b}=t\left(x, T^{m} x\right)-\inf _{u \in C_{x}(x) \backslash\{0\}} \tau_{\bar{m}}^{+}(x, u) .
\end{aligned}
$$

Note that the collisions $\left\{T^{\bar{m}} x, \ldots, T^{m-1} x\right\}$ with the neutral boundary $\Gamma^{0}$ only affect the quantity $t\left(x, T^{m} x\right)$.

By Condition B2, we have

$$
\sup _{u \in C_{x}(x) \backslash\{0\}} \tau_{\frac{+}{m}}^{+}(x, u) \leq \lambda_{j}^{+} .
$$

Next, directly from the definition of the cone $C_{x}(x)$ when $\Gamma_{j}$ is dispersing, and from Part (3) of Theorem 4.1 when $\Gamma_{j}$ is focusing, it follows that

$$
\frac{d\left(T^{\bar{m}} x\right)}{2} \leq \inf _{u \in C_{x}(x) \backslash\{0\}} \tau_{\bar{m}}^{+}(x, u)
$$

Finally, note that $G_{T^{m} x}^{-}\left(T^{m} x\right) \leq \lambda_{k}^{-}$. By the previous observations,

$$
\begin{equation*}
a \leq \lambda_{k}^{-}, \quad t\left(x, T^{m} x\right)-\lambda_{j}^{+} \leq \bar{a}, \quad \bar{b} \leq t\left(x, T^{m} x\right)-\frac{d\left(T^{\bar{m}} x\right)}{2} \tag{4}
\end{equation*}
$$

By Lemma 5.2, $D_{x} T^{m} C_{x}(x) \subset \operatorname{int} C_{T^{m} x}\left(T^{m} x\right) \cup\{0\}$. Hence, the expansion $\widetilde{\sigma}_{x}\left(D_{x} T^{m}\right)$ can be estimated by using a formula proved by Wojtkowski [32, Lemma A. 4 and Appendix B]. We obtain ${ }^{4}$

$$
\begin{equation*}
\widetilde{\sigma}_{x}\left(D_{x} T^{m}\right)=\sqrt{1+w}+\sqrt{w}, \quad \text { where } w=\frac{\bar{a}-a}{\bar{b}-\bar{a}} \tag{5}
\end{equation*}
$$

From (4), it follows that

$$
w \geq \frac{t\left(x, T^{m} x\right)-\lambda_{j}^{+}-\lambda_{k}^{-}}{\lambda_{j}^{+}-d\left(T^{m} x\right) / 2} .
$$

It is easy to see that there exists a constant $c>0$ depending only on the billiard domain $\Omega$ such that

$$
0<\lambda_{j}^{+}-\frac{d\left(T^{\bar{m}} x\right)}{2}<\frac{1}{c} .
$$

Hence,

$$
w \geq c\left(t\left(x, T^{m} x\right)-\lambda_{j}^{+}-\lambda_{k}^{-}\right) .
$$

Now, the wanted conclusion follows from (5) once we have observed that $t\left(x, T^{m} x\right)-\lambda_{j}^{+}-\lambda_{k}^{-} \geq \lambda$ by Condition B2 if $\Gamma_{j}$ or $\Gamma_{k}$ is focusing, and $\lambda_{j}^{+}=\lambda_{k}^{-}=0$ if $\Gamma_{j}$ and $\Gamma_{k}$ are dispersing.

If $x \in M \backslash\left(R_{\infty}^{+} \cup N^{+}\right)$, the positive semi-orbit of $x$ visits $E$ infinitely many times. That is, there exists a strictly increasing sequence $\left\{r_{k}\right\}_{k \in \mathbb{N}}$ of non-negative integers such that for every $k \in \mathbb{N}$, we have $T^{r_{k}} x \in E$, and $T^{i} x \notin E$ for all $r_{k}<i<r_{k+1}$. Note that $r_{1}=0$ if and only if $x \in E$. We call $\left\{r_{k}\right\}_{k \in \mathbb{N}}$ the sequence of the positive return times to $E$ of $x$.

Similarly, if $x \in M \backslash\left(R_{\infty}^{-} \cup N^{-}\right)$, there exists a strictly decreasing sequence of non-positive integers $\left\{r_{k}\right\}_{k \in \mathbb{N}}$ such that for every $k \in \mathbb{N}$, we have $T^{r_{k}} x \in E$ and $T^{i} x \notin E$ for all $r_{k+1}<i<r_{k}$. In this case, we call $\left\{r_{k}\right\}_{k \in \mathbb{N}}$ the sequence of the negative return times to $E$ of $x$.

Proposition 6.2. Suppose that $x \in M \backslash\left(R_{\infty}^{+} \cup N^{+}\right)$, and let $\left\{r_{k}\right\}$ be the sequence of the positive return times to $E$ of $x$. Then

$$
\lim _{k \rightarrow+\infty} \widetilde{\sigma}_{T^{r_{1} x}}^{*}\left(D_{T^{r_{1} x}} T^{r_{k}-r_{1}}\right)=\lim _{k \rightarrow+\infty} \widetilde{\sigma}_{T^{r_{1} x}}\left(D_{T^{r_{1}} x} T^{r_{k}-r_{1}}\right)=+\infty .
$$

Similarly, suppose $x \in M \backslash\left(R_{\infty}^{-} \cup N^{-}\right)$, and let $\left\{r_{k}\right\}$ be the sequence of the negative return times to $E$ of $x$. Then

$$
\lim _{k \rightarrow+\infty} \widetilde{\sigma}_{T^{r_{k x}}}^{*}\left(D_{T^{r_{k}}} T^{r_{1}-r_{k}}\right)=\lim _{k \rightarrow+\infty} \widetilde{\sigma}_{T^{r_{k}}}\left(D_{T^{r_{k}}} T^{r_{1}-r_{k}}\right)=+\infty .
$$

[^4]Proof. We prove the proposition only for $x \in M \backslash\left(R_{\infty}^{+} \cup N^{+}\right)$. For the other case, the proof is similar.

Define $x_{k}=T^{r_{k}} x$ and $m_{k}=r_{k+1}-r_{k}$ for all $k \in \mathbb{N}$. By Lemma 6.1, $D_{x_{1}} T^{r_{2}-r_{1}} C_{x_{1}}\left(x_{1}\right) \subset \operatorname{int} C_{x_{2}}\left(x_{2}\right) \cup\{0\}$, which in turn implies

$$
\begin{equation*}
\tilde{\sigma}_{x_{1}}^{*}\left(D_{x_{1}} T^{r_{2}-r_{1}}\right)>0 . \tag{6}
\end{equation*}
$$

Next, one can easily show that

$$
\begin{equation*}
\widetilde{\sigma}_{x_{1}}^{*}\left(D_{x_{1}} T^{r_{k}-r_{1}}\right) \geq \widetilde{\sigma}_{x_{1}}^{*}\left(D_{x_{1}} T^{r_{2}-r_{1}}\right) \cdot \widetilde{\sigma}_{x_{2}}\left(D_{x_{2}} T^{r_{k}-r_{2}}\right) \tag{7}
\end{equation*}
$$

and

$$
\begin{equation*}
\widetilde{\sigma}_{x_{1}}\left(D_{x_{1}} T^{r_{k}-r_{1}}\right) \geq \prod_{i=1}^{k-1} \widetilde{\sigma}_{x_{i}}\left(D_{x_{i}} T^{m_{i}}\right) \tag{8}
\end{equation*}
$$

It follows from Lemma 6.1 that the sequence $\prod_{i=1}^{k-1} \widetilde{\sigma}_{x_{i}}\left(D_{x_{i}} T^{m_{i}}\right)$ is strictly increasing in $k$. It does not diverge only if $x_{i}$ for every $i$ sufficiently large, and $\lim _{i \rightarrow+\infty} t\left(x_{i}, x_{i+1}\right)=0$. This means that the positive semi-trajectory of $x$ is eventually trapped inside a cusp formed by two dispersing components or a dispersing and a flat component. However, every trajectory entering a cusp leaves it after finitely many collisions [7, Appendix A1.3]. Hence, $\prod_{i=1}^{k-1} \widetilde{\sigma}_{x_{i}}\left(D_{x_{i}} T^{m_{i}}\right) \rightarrow+\infty$ as $k \rightarrow+\infty$. By (8), we conclude that

$$
\lim _{k \rightarrow+\infty} \widetilde{\sigma}_{x_{1}}\left(D_{x_{1}} T^{r_{k}-r_{1}}\right)=+\infty .
$$

The same argument applied to $x_{2}$ shows that $\widetilde{\sigma}_{x_{2}}\left(D_{x_{2}} T^{r_{k}-r_{2}}\right) \rightarrow+\infty$ as $k \rightarrow+\infty$. This together with (6) and (7) gives

$$
\lim _{k \rightarrow+\infty} \widetilde{\sigma}_{x_{1}}^{*}\left(D_{x_{1}} T^{r_{k}-r_{1}}\right)=+\infty
$$

Corollary 6.3. Every element $x \in H$ is sufficient, and has a quadruple $(l, N, O, K)$ such that $O \cup T^{-N} O \subset E$.

Proof. The set $H$ can be decomposed as follows:

$$
H=\left(M \backslash\left(R_{\infty}^{+} \cup N^{+}\right)\right) \cup\left(M \backslash\left(R_{\infty}^{-} \cup N^{-}\right)\right) .
$$

So if $x \in H$, then $x \in M \backslash\left(R_{\infty}^{+} \cup N^{+}\right)$or $x \in M \backslash\left(R_{\infty}^{-} \cup N^{-}\right)$. We prove the corollary only for the case $x \in M \backslash\left(R_{\infty}^{+} \cup N^{+}\right)$. The other case can be proved similarly.

Let $\left\{r_{k}\right\}_{k \in \mathbb{N}}$ be the sequences of the positive return times to $E$ of $x$. Define $x_{k}=T^{r_{k}} x$ for all $k \in \mathbb{N}$. Note that although $x \in S_{1}^{-}$is allowed, nevertheless $x_{k} \notin S_{1}^{-}$for all $k>1$. It follows that $x_{k}$ is an interior point of $E$ for every $k>1$.

Proposition 6.2 applied to $x_{2}$ implies that there exists $k>2$ such that $\widetilde{\sigma}_{x_{2}}\left(D_{x_{2}} T^{r_{k}-r_{2}}\right)>3$. Set $l=r_{k}$ and $N=r_{k}-r_{2}$. We need to find the neighborhood $O$ of $x_{k}$ and the invariant continuous cone field $K$ on $O \cup T^{-N} O$ as in Definition A.1.

Since $T^{N}$ is a local diffeomorphism at $x_{2}$, and the cone fields ( $U_{x_{2}}, C_{x_{2}}$ ) and ( $U_{x_{k}}, C_{x_{k}}$ ) are continuous, the function $y \mapsto \widetilde{\sigma}_{x_{2}}\left(D_{y} T^{N}\right)$ is continuous at $x_{2}$. Hence, there exists a neighborhood $V$ of $x_{2}$ such that i) $V \subset U_{x_{2}} \backslash R_{N}^{+}$and $T^{N} V \subset U_{x_{k}}$, ii) $V \cap T^{N} V=\emptyset$ if $x_{2} \neq x_{k}$, iii) $\widetilde{\sigma}_{x_{2}}\left(D_{y} T^{N}\right)>3$ for every $y \in V$. Since $x_{2}$ and $x_{k}$ are interior points of $E$, we can further choose $V$ so that $V \cup T^{N} V \subset E$. It is easy to see that a similar choice can be made for the case $x \in M \backslash\left(R_{\infty}^{-} \cup N^{-}\right)$.

To construct the cone field $\left(O \cup T^{-N} O, K\right)$, we consider separately the cases $x_{2} \neq x_{k}$ and $x_{2}=x_{k}$. If $x_{2} \neq x_{k}$, then define $O=T^{N} V$ and

$$
K(y)= \begin{cases}C_{x_{2}}(y) & \text { if } y \in T^{-N} O \\ C_{x_{k}}(y) & \text { if } y \in O\end{cases}
$$

Note that $O \cap T^{-N} O=V \cap T^{N} V=\emptyset$, and that $C_{x_{2}}$ and $C_{x_{k}}$ are continuous. This remark and Lemma 5.2 imply that the cone field $K$ is continuous, invariant and satisfies

$$
\sigma_{K}\left(D_{y} T^{N}\right)=\widetilde{\sigma}_{x_{2}}\left(D_{y} T^{N}\right)>3 \quad \text { for every } y \in T^{-N} O
$$

If $x_{2}=x_{k}$ (i.e., $x$ is periodic), then let $O=T^{N} V$. Since $U_{x_{2}}=U_{x_{k}}$, we have $O \cup T^{-N} O=V \cup T^{N} V \subset U_{x_{2}}$. Now, define $K$ to be the restriction of $C_{x_{2}}$ to $O \cup T^{-N} O$. As for the precious case, we see that the cone field $K$ is continuous, invariant and satisfies $\sigma_{K}\left(D_{y} T^{N}\right)=\widetilde{\sigma}_{x_{2}}\left(D_{y} T^{N}\right)>3$ for every $y \in T^{-N} O$.

We can now proceed to prove Conditions L1-L3.
6.2. Proof of Conditions L1-L3. From now on, for every sufficient point $x \in H$, we will always take as a quadruple $(l, N, O, K)$ the one constructed in the proof of Corollary 6.3. The choice of this quadruple is of crucial importance for the proof of Conditions L1-L4, because the cone $\left(O \cup T^{N} O, K\right)$ is obtained directly from $\left\{\left(U_{y}, C_{y}\right)\right\}_{y \in E}$, and the neighborhood $O$ has the property that $O \cup T^{N} O \subset E$.

We proved a number of properties concerning the geometry of the singulars sets $R_{k}^{-}$and $R_{k}^{+}$in [15]. We will use several of these results in the proof of the following propositions.
Proposition 6.4. Condition L1 is satisfied.
Proof. For billiards satisfying Conditions B1 and B2, the regularity of the singular sets $R_{k}^{-}$and $R_{k}^{+}$is proved in [15, Theorem 2.2].
Proposition 6.5. Every element of $H$ satisfies Condition L2.
Proof. This proposition follows from [15, Proposition 6.2]. For the convenience of the reader, we give here a more direct proof. We prove only the first part of Condition L2, because the second one can be proved similarly.

Let $x \in H$, and let $\left(O \cup T^{N} O, K\right)$ be the cone field for $x$ as in the proof of Corollary 6.3. Suppose that $\Sigma$ is a component of $R_{k}^{-} \cap T^{-N} O$
for some $k \in \mathbb{N}$. Without loss of generality, we can assume $\Sigma \not \subset R_{k-1}^{-}$. Let $\Sigma^{\prime}=\Sigma \backslash \partial \Sigma$. Recall that $\partial \Sigma$ denotes the set of the endpoints of $\Sigma$.

The map $T^{-k+1}: \Sigma^{\prime} \rightarrow T^{-k+1} \Sigma^{\prime}$ is a diffeomorphism, and the family of rays emerging from the points of $\mathcal{I}\left(T^{-k} \Sigma^{\prime}\right)$ focuses in linear approximation at a corner of $\partial Q$ or at a point lying on a dispersing component of $\partial Q$. Hence, if $y \in \Sigma^{\prime}$ and $u \in T_{y} \Sigma \backslash\{0\}$, then

$$
\begin{equation*}
\tau^{-}\left(T^{-k+1} y, D_{y} T^{-k+1} u\right)=t\left(\mathcal{I}\left(T^{-k+1} y\right)\right) \tag{9}
\end{equation*}
$$

Let $v \in K^{\prime}(y) \backslash\{0\}$. By construction, we have $K^{\prime}(y)=C_{x_{2}}(y)$ for some $x_{2} \in T^{-N} O$ (see the proof of Corollary 6.3). There exists $1 \leq i<k$ such that $T^{-i} \Sigma^{\prime} \subset E$, and $T^{-j} \Sigma^{\prime} \subset M^{0}$ for every $i<j<k$. Hence, $D_{y} T^{-i} v \in \operatorname{int} C_{T^{-i} y}^{\prime}\left(T^{-i} y\right)$ by Lemma 5.2. Now, Condition B2 implies

$$
\tau^{-}\left(T^{-i} y, D_{y} T^{-i} v\right)<\sum_{j=i}^{k-1} t\left(\mathcal{I}\left(T^{-j} y\right)\right)
$$

To obtain this conclusion, we use Part (2) of Condition B2 whenever the family of rays emerging from the points of $\mathcal{I}\left(T^{-k} \Sigma^{\prime}\right)$ focuses at a corner of $\partial Q$. The previous inequality together with the fact that $T^{-j} \Sigma^{\prime} \subset M^{0}$ for every $i<j<k$ implies

$$
\begin{aligned}
\tau^{-}\left(T^{-k+1} y, D_{y} T^{-k+1} v\right) & =\tau^{-}\left(T^{-i} y, D_{y} T^{-i} v\right)-\sum_{j=i}^{k-2} t\left(\mathcal{I}\left(T^{-j} y\right)\right) \\
& <t\left(\mathcal{I}\left(T^{-k+1} y\right)\right)
\end{aligned}
$$

By comparing the last conclusion with (9), we infer that $u \in \operatorname{int} K(y)$.
Finally, let $y \in \partial \Sigma$, and let $u \in T_{y} \Sigma \backslash\{0\}$. Since $\Sigma$ is a $C^{2}$ arc, there exist two sequence $y_{n} \in \Sigma^{\prime}$ and $u_{n} \in T_{y_{n}} \Sigma$ such that $y_{n} \rightarrow y$ and $u_{n} \rightarrow u$ as $n \rightarrow+\infty$. Since $u_{n} \in \operatorname{int} K(y)$ and $K$ is a continuous cone field, it follows that $u \in K(y)$. This completes the proof.

Proposition 6.6. Every element of $H$ satisfies Condition L3.
Proof. We only proof the part of Condition L3 concerning $S_{1}^{-}$, because the proof of the other part is similar.

By [15, Propositions 6.17-6.19], the set $S_{1}^{-} \cap R_{\infty}^{+}$is at most countable. This combined with Condition B4 gives $m_{-}\left(S_{1}^{-} \cap\left(N^{+} \cup R_{\infty}^{+}\right)\right)=0$. Hence, we first need to prove that each element of $S_{1}^{-} \backslash\left(N^{+} \cup R_{\infty}^{+}\right)$is $u$-essential.

Let $y \in S_{1}^{-} \backslash\left(N^{+} \cup R_{\infty}^{+}\right)$, and let $\left\{r_{k}\right\}_{k \in \mathbb{N}}$ be the sequence of the positive return times to $E$ of $y$, which exists by Proposition 6.2. Define $y_{k}=T^{r_{k}} y$ for every $k \in \mathbb{N}$.

Since $T^{r_{1}}$ is a local diffeomorphism around $y$, there exist a neighborhood $U$ of $y$ and a constant $c>0$ such that

$$
\begin{equation*}
\frac{\left\|D_{y} T^{r_{1}} u\right\|}{\|u\|} \geq c \tag{10}
\end{equation*}
$$

for all $y \in U$ and for all $u \in T_{y} M \backslash\{0\}$. Now, let $\alpha>0$. By Proposition 6.2, there is $k>1$ such that

$$
\begin{equation*}
\tilde{\sigma}_{y_{1}}^{*}\left(D_{y_{1}} T^{r_{k}-r_{1}}\right) \geq \alpha / c . \tag{11}
\end{equation*}
$$

The argument that proves the continuity of $y \mapsto \widetilde{\sigma}_{x_{2}}\left(D_{y} T^{N}\right)$ at $x_{2}$ in the proof of Corollary 6.3 also proves the continuity of $z \mapsto$ $\widetilde{\sigma}_{y_{1}}^{*}\left(D_{z} T^{r_{k}-r_{1}}\right)$ at $y_{1}$. Next, since $y \in S_{1}^{-} \backslash\left(N^{+} \cup R_{\infty}^{+}\right)$, the point $y$ is not periodic, and in particular, $y \neq y_{k}$. These observations imply that there is a neighborhood $V$ of $y_{1}$ such that i) $V \subset U_{y_{1}} \backslash\left(R_{r_{1}}^{-} \cup R_{r_{k}-r_{1}}^{+}\right)$ and $T^{r_{k}-r_{1}} V \subset U_{y_{k}}$, ii) $T^{-r_{1}} V \subset U$, iii) $T^{-r_{1}} V \cap T^{r_{k}-r_{1}} V=\emptyset$, and iv) $\widetilde{\sigma}_{y_{1}}^{*}\left(D_{z} T^{r_{k}-r_{1}}\right)>\alpha / c$ for every $z \in V$.

Now, define $O_{y, \alpha}=T^{-r_{1}} V$ and

$$
K_{y, \alpha}= \begin{cases}D_{T^{r_{1}}} T^{-r_{1}} C_{y_{1}}\left(T^{r_{1}} z\right) & \text { if } z \in O_{y, \alpha} \\ C_{y_{k}}(z) & \text { if } z \in T^{r_{k}} O_{y, \alpha} .\end{cases}
$$

Note that $O_{y, \alpha} \cap T^{r_{k}} O_{y, \alpha}=T^{-r_{1}} V \cap T^{r_{k}-r_{1}} V=\emptyset$, the cone fields $C_{y_{1}}$ and $C_{y_{k}}$ are continuous, and $T^{r_{1}}$ is a diffeomorphism on $O_{y, \alpha}$. From the previous remark, Lemma 5.2 and the construction of $K_{y, \alpha}$, it follows that $K_{y, \alpha}$ is continuous and invariant. Set $n_{y, \alpha}=r_{k}$. For every $z \in$ $O_{y, \alpha}$, we have

$$
\begin{aligned}
\sigma_{K_{y, \alpha}}\left(D_{z} T^{n_{y, \alpha}}\right) & =\inf _{u \in \operatorname{int} K_{y, \alpha}(z)} \frac{\sqrt{Q_{K_{y, \alpha}}\left(T^{n_{y, \alpha}} z, D_{z} T^{n_{y, \alpha} u}\right)}}{\|u\|} \\
& =\inf _{u \in \operatorname{int} K_{y, \alpha}(z)} \frac{\left\|D_{z} T^{r_{1}} u\right\|}{\|u\|} \cdot \frac{\sqrt{Q_{C_{y_{k}}}\left(T^{n_{y, \alpha}} z, D_{z} T^{n_{y, \alpha}} u\right)}}{\left\|D_{z} T^{r_{1}} u\right\|} \\
& \geq \inf _{u \in \operatorname{int} K_{y, \alpha}(z)} \frac{\left\|D_{z} T^{r_{1}} u\right\|}{\|u\|} \cdot \widetilde{\sigma}_{y_{1}}^{*}\left(D_{T^{r_{1}}} T^{r_{k}-r_{1}}\right),
\end{aligned}
$$

where the last inequality is a consequence of $D_{z} T^{r_{1}} K_{y, \alpha}(z)=C_{y_{1}}\left(T^{r_{1}} z\right)$ and $T^{r_{1}} z \subset U_{1}$. By (10) and (11), we obtain

$$
\sigma_{K_{y, \alpha}}\left(D_{z} T^{n_{y, \alpha}}\right) \geq c \cdot \widetilde{\sigma}_{y_{1}}^{*}\left(D_{T^{r_{1}}} T^{r_{k}-r_{1}}\right) \geq c \cdot \frac{\alpha}{c}=\alpha .
$$

We conclude that $y$ is an $u$-essential point (see Definition A.2).
To prove that the cone field $\left(O \cup T^{-N} O, K\right)$ for $x \in H$ and the cone field ( $O_{y, \alpha} \cup T^{n_{y, \alpha}} O_{y, \alpha}, K_{y, \alpha}$ ) just constructed are jointly invariant, we observe that $K$ is a restriction of the cone fields $C_{x_{2}}$ and $C_{x_{k}}$ for some $x_{2}, x_{k} \in E$, and that $K_{y, \alpha}$ is obtained by taking the preimage of $C_{y_{1}}$ and the restriction of $C_{y_{k}}$ for some $y_{1}, y_{k} \in E$. The joint invariance then follows from Lemma 5.2.

## 7. Proof of Condition L4

In Corollary 6.3, we proved that every $x \in H$ is a sufficient point with a quadruple $(l, N, O, K)$ having the property that $O \cup T^{-N} O \subset E$. We recall that $l \in \mathbb{Z}, N \in \mathbb{N}, O$ is a neighborhood of $T^{l} x$, and $K$ is
a continuous invariant cone field on $O \cup T^{-N} O$. We also recall that $\left\{\left(U_{y}, C_{y}\right)\right\}_{y \in E}$ is the family of cone fields introduced in Section 4.3.
In this section, we prove Condition L4 for every $x \in H$. This condition requires the existence of stable and unstable manifolds a.e. on the neighborhood $O$, which is guaranteed by Proposition A.3. We derive L4 from the following property:

Non-contraction: there exists $\beta^{\prime}>0$ such that if $z \in E \backslash R_{m}^{+}$and $T^{m} z \in E \cup \mathcal{I}(E)$ for some $m \in \mathbb{N}$, then

$$
\begin{equation*}
\left\|D_{z} T^{m} v\right\| \geq \beta^{\prime}\|v\| \quad \text { for every } v \in C_{z}(z) \tag{12}
\end{equation*}
$$

Since $E$ is the set of the collisions 'entering' non-flat boundary components, $\mathcal{I}(E)$ is the set of the collisions 'leaving' non-flat boundary components.

To prove this property, we decompose the billiard orbits into special blocks, and study (12) separately for each block. This analysis will be carried out using certain semi-norms defined in terms of transversal Jacobi fields.
7.1. Jacobi fields and semi-norms. Jacobi fields for billiards are special vectors fields defined along trajectories of the billiard flow. For the definitions of a billiard flow and a Jacobi field for billiards, we refer the reader to [10, 20] and [16, 33], respectively. Here, we limit ourselves to recall the main properties of the Jacobi fields for billiards.

The billiard flow $\Phi^{t}$ on the domain $\Omega$ acts on the unit tangent bundle $T_{1} \bar{\Omega} \subset T_{1} \mathbb{R}^{2}$ of the closure of $\Omega$. Let $\pi: T_{1} \bar{\Omega} \rightarrow \bar{\Omega}$ be the natural projection. A billiard trajectory $\xi$ is a piecewise linear curve $\xi:(a, b) \rightarrow$ $\bar{\Omega}$ with $-\infty \leq a<0<b \leq+\infty$ such that $\xi(t)=\pi\left(\Phi^{t}(z)\right)$ for all $t \in(a, b)$ and for some $z \in T_{1} \bar{\Omega}$. We say that $\xi(t)$ is a collision for some $t \in(a, b)$ if $\xi(t) \in \partial \Omega$, or equivalently, if there exists $y \in M$ such that $\xi(t)=q(y)$. A collision is called non-tangential provided that $y \notin A_{2}$.

A transversal Jacobi field $\mathcal{J}$ along a billiard trajectory $\xi$ is a piecewise smooth map $\mathcal{J}:(a, b) \rightarrow \mathbb{R}^{2}$ such that $\mathcal{J}(t)$ and its time time derivative $\mathcal{J}^{\prime}(t)$ are perpendicular to $\xi^{\prime}(t)$ every $t \in(a, b)$, and satisfy further properties for which we refer the reader to [16, 33]. It turns out that for any fixed $t$, the field $\mathcal{J}$ is uniquely determined by the values $\mathcal{J}(t)$ and $\mathcal{J}^{\prime}(t)$. Let $\xi^{\perp}(t)$ denote the unit vector of $\mathbb{R}^{2}$ orthogonal to $\xi^{\prime}(t)$ such that $\left\{\xi^{\prime}(t), \xi^{\perp}(t)\right\}$ is a positively oriented basis of $\mathbb{R}^{2}$. Since the vectors $\mathcal{J}(t)$ and $\mathcal{J}^{\prime}(t)$ are always collinear to $\xi^{\perp}(t)$, we can identify $\mathcal{J}(t)$ and $\mathcal{J}^{\prime}(t)$ with the corresponding orthogonal projections onto $\xi^{\perp}(t)$, and think of $\mathcal{J}(t)$ and $\mathcal{J}^{\prime}(t)$ as real numbers.

Suppose that there exists $0<\delta<b$ such that $\xi(t)$ is not a collision for all $t \in(0, \delta)$, and $\xi(\delta)=q(y)$ is a non-tangential collision for some $y \in M \backslash A_{2}$. If $\mathcal{J}$ is a Jacobi field along $\xi$, then the evolution of the
pair $\left(\mathcal{J}, \mathcal{J}^{\prime}\right)$ is governed by the following equations:

$$
\begin{aligned}
& \binom{\mathcal{J}(t)}{\mathcal{J}^{\prime}(t)}=F(t)\binom{\mathcal{J}(0)}{\mathcal{J}^{\prime}(0)} \quad \text { for } 0 \leq t<\delta, \\
& \binom{\mathcal{J}(\delta)}{\mathcal{J}^{\prime}(\delta)}=R(y) \lim _{t \rightarrow \delta^{-}}\binom{\mathcal{J}(t)}{\mathcal{J}^{\prime}(t)},
\end{aligned}
$$

where

$$
F(t):=\left(\begin{array}{cc}
1 & t  \tag{13}\\
0 & 1
\end{array}\right) \quad \text { and } \quad R(y):=\left(\begin{array}{cc}
-1 & 0 \\
\frac{2}{d(y)} & -1
\end{array}\right) .
$$

For every $y \in M \backslash A_{2}$ and every $u=\left(u_{s}, u_{\theta}\right) \in T_{y} M$, define

$$
\begin{equation*}
J(u)=\sin \theta(y) u_{s}, \quad J^{\prime}(u)=-u_{\theta}-\kappa(y) u_{s}, \tag{14}
\end{equation*}
$$

and observe that the transformation $u \mapsto\left(J(u), J^{\prime}(u)\right)$ is a bijection between $T_{y} M$ and $\mathbb{R}^{2}$. In particular, the pair $\left(J, J^{\prime}\right)$ is a system of coordinates for $T_{y} M$. Note that $J^{\prime}$ is just a symbol and not the time derivative of $J$. Suppose that $\xi(0)=q(y)$ is a non-tangential collision for some $y \in M \backslash A_{2}$, and let $u \in T_{y} M$. Our previous observation combined with the fact that the entire transversal Jacobi field $\mathcal{J}$ along $\xi$ is determined by the choice $\mathcal{J}(0)=J(u)$ and $\mathcal{J}^{\prime}(0)=J^{\prime}(u)$ shows that there is a bijective correspondence between transversal Jacobi fields along $\xi$ and elements of $T_{y} M$, and so we can identify elements of $T_{y} M$ with transversal Jacobi fields. Now, in addition to the previous assumption, suppose that there exists $0<\delta<b$ such that $\xi(0)$ and $\xi(\delta)$ are consecutive non-tangential collisions. This means that $\delta=t(y)$ and $\xi(\delta)=q(T y)$. Moreover, if $u \in T_{y} M$ and $u_{1} \in T_{T y} M$ are the tangent vectors corresponding to $\left(\mathcal{J}(0), \mathcal{J}^{\prime}(0)\right)$ and $\left(\mathcal{J}(\delta), \mathcal{J}^{\prime}(\delta)\right)$, respectively, one can show that $D_{y} T u=u_{1}$. It follows that the matrix of $D_{y} T$ in coordinates $\left(J, J^{\prime}\right)$ is given by

$$
R(T y) F(t(y))=\left(\begin{array}{cc}
-1 & -t(y)  \tag{15}\\
\frac{2}{d(T y)} & -1+\frac{2 t(y)}{d(T y)}
\end{array}\right) .
$$

The matrices (13) and (15) have determinant equal to 1 , and so $D_{y} T$ preserves the standard symplectic form $J \wedge J^{\prime}$. Finally, from (14), it follows immediately that the forward focusing time of a tangent vector $u \in T_{y} M$ in terms of $J(u)$ and $J^{\prime}(u)$ is given by

$$
\tau^{+}(y, u)= \begin{cases}-\frac{J(u)}{J^{\prime}(u)} & \text { if } J^{\prime}(u) \neq 0  \tag{16}\\ \infty & \text { if } J^{\prime}(u)=0\end{cases}
$$

We now introduce two semi-norms, which will help us prove the Non-contraction property.
Definition 7.1. For every $y \in M$ and every $u \in T_{y} M$, define

$$
\|u\|_{J}=\sqrt{J^{2}(u)+J^{\prime 2}(u)} \quad \text { and } \quad|u|_{J^{\prime}}=\left|J^{\prime}(u)\right| .
$$

In the rest of this subsection, we prove that $\|\cdot\|$ and $\|\cdot\|_{J}$ are equivalent on a certain subset of the tangent bundle $T M$.

Lemma 7.2. There exists a constant $\alpha_{1}>1$ depending only on $\Omega$ and the family of cone fields $\left\{\left(U_{y}, C_{y}\right)\right\}_{y \in E^{+}}$such that if $y \in E^{+}, z \in U_{y}$ and $0 \leq k \leq n(z)$, then

$$
\left|D_{z} T^{k} u\right|_{J^{\prime}} \leq\left\|D_{z} T^{k} u\right\|_{J} \leq \alpha_{1}\left|D_{z} T^{k} u\right|_{J^{\prime}} \quad \text { for } u \in C_{y}(z)
$$

Proof. Let $y, z, k, u$ be as in the hypotheses of the lemma. It is clear that $\left|D_{z} T^{k} u\right|_{J^{\prime}} \leq\left\|D_{z} T^{k} u\right\|_{J}$. By Part (3) of Theorem 4.1, there exists a constant $f>0$ depending only on $\left\{\left(U_{y}, C_{y}\right)\right\}_{y \in E}$ such that

$$
\frac{d\left(T^{k} z\right)}{2} \leq \tau_{k}^{+}(z, u) \leq f
$$

By (16), we then have $\left|J\left(D_{z} T^{k} u\right)\right| \leq f\left|J^{\prime}\left(D_{z} T^{k} u\right)\right|$, which gives the remaining inequality with $\alpha_{1}=\left(1+f^{2}\right)^{1 / 2}$.

Lemma 7.3. There exists $\alpha_{2}>1$ such that if $y \in M \backslash \partial M$, then

$$
\|u\|_{J} \leq \alpha_{2}\|u\| \quad \text { for } u \in T_{y} M
$$

Proof. A straightforward computation gives the wanted inequality with $\alpha_{2}=2^{1 / 2}\left(1+\kappa_{1}^{2}\right)^{1 / 2}$, where $\kappa_{1}=\max _{z \in M}|\kappa(z)|$.

Lemma 7.4. There exists $\alpha_{3}>0$ such that if $y \in M^{-}$and $z \in U_{y}$, then

$$
\|u\| \leq \alpha_{3}\|u\|_{J} \quad \text { for } u \in C_{y}(z)
$$

Proof. Let $\kappa_{2}=\min _{z \in M^{-}}|\kappa|>0$. By the definition of $C_{y}(z)$, we have $u_{\theta} / u_{s} \leq \kappa(z)<0$ for all $0 \neq u \in C_{y}(z)$. Hence,

$$
\begin{aligned}
\|u\|_{J}^{2} & \geq|J(u)|^{2}=\kappa^{2}(z) u_{s}^{2}+u_{\theta}^{2}+2 \kappa(z) u_{s} u_{\theta} \\
& \geq \kappa_{2}^{2} u_{s}^{2}+u_{\theta}^{2} \geq \frac{\kappa_{2}^{2}}{1+\kappa_{2}^{2}}\left(u_{s}^{2}+u_{\theta}^{2}\right)=\frac{\kappa_{2}^{2}}{1+\kappa_{2}^{2}}\|u\|^{2} .
\end{aligned}
$$

Lemma 7.5. There exists $\alpha_{4}>0$ such that if $y \in E^{+}, z \in U_{y}$ and $u \in C_{y}(z)$, then

$$
\left\|D_{z} T^{k} u\right\| \leq \alpha_{4}\left\|D_{z} T^{k} u\right\|_{J} \quad \text { for } 0 \leq k \leq n(z)
$$

Proof. Denote by $\kappa_{3}$ the maximum of $\kappa$ on $M^{+}$, and denote by $\bar{m}$ the smallest $m_{i}$ associated to focusing components of $\partial \Omega$ as in Theorem 4.1. Let $0 \neq u \in C_{y}(z)$, and write $\left(u_{k, s}, u_{k, \theta}\right)$ for the vector $D_{z} T^{k} u$. By Theorem 4.1, if $z \in M_{i} \subset M^{+}$, then $-\kappa\left(T^{k} z\right)+m_{i} \leq m\left(u_{k}\right) \leq \kappa\left(T^{k} z\right)$
for $0 \leq k \leq n(z)$. In particular, $\left|u_{k, \theta}\right| \leq \kappa\left(T^{k} z\right)\left|u_{k, s}\right|$. Therefore,

$$
\begin{aligned}
\left\|D_{z} T^{k} u\right\|_{J}^{2} & \geq\left(\kappa\left(T^{k} z\right) u_{k, s}+u_{k, \theta}\right)^{2}=\left(\kappa\left(T^{k} z\right)+m\left(u_{k}\right)\right)^{2} u_{k, s}^{2} \\
& \geq\left(\kappa\left(T^{k} z\right)-\kappa\left(T^{k} z\right)+m_{i}\right)^{2} u_{k, s}^{2} \geq \frac{m_{i}^{2}}{1+\kappa^{2}\left(T^{k} z\right)}\left\|D_{z} T^{k} u\right\|^{2} \\
& \geq \frac{\bar{m}^{2}}{1+\kappa_{3}^{2}}\left\|D_{z} T^{k} u\right\|^{2} .
\end{aligned}
$$

Lemma 7.6. There exist constants $\epsilon_{0}>0$ and $0<\theta_{0}<\pi / 2$ such that $\theta\left(M^{0} \cap S_{1}^{-}\left(\epsilon_{0}\right)\right) \subset\left(\theta_{0}, \pi-\theta_{0}\right)$.

Proof. The lemma is an immediate consequence of the following easy-to-check fact. Consider a flat component $\Gamma_{i}$ of $\partial \Omega$. If the line containing $\Gamma_{i}$ contains also a corner $p$ or is tangent to a dispersing component $\Gamma_{j}$, then no ray emerging from the elements of $M_{i} \cap S_{1}^{+}$contains $p$ or is tangent to $\Gamma_{j}$.

Lemma 7.7. Let $\epsilon_{0}$ be the constant in Lemma 7.6. There exists $\alpha_{5}>0$ such that if $y \in M^{0} \cap S_{1}^{-}\left(\epsilon_{0}\right)$, then

$$
\|u\| \leq \alpha_{5}\|u\|_{J} \quad \text { for } u \in T_{y} M
$$

Proof. Since $\kappa(y)=0$, we have $\|u\|_{J}^{2} \geq \sin ^{2} \theta_{0} u_{s}^{2}+u_{\theta}^{2}>\sin ^{2} \theta_{0}\|u\|^{2}$.
Corollary 7.8. There exist two constants $0<A_{1}<A_{2}$ such that if $y \in E, z \in U_{y}, u \in C_{y}(z)$ or $y \in M^{0} \cap S_{1}^{-}\left(\epsilon_{0}\right), u \in T_{y} M$, then

$$
A_{1}\|u\|_{J} \leq\|u\| \leq A_{2}\|u\|_{J} .
$$

Proof. The claim follows from Lemmas 7.3-7.7.
Remark 7.9. It can be easily proved that Corollary 7.8 remains valid if $\|\cdot\|$ is replaced by $|\cdot|_{J^{\prime}}$, thus showing that the semi-norms $\|\cdot\|,\|\cdot\|_{J}$, $|\cdot|_{J^{\prime}}$ are equivalent in the sense specified in the corollary. However, this general equivalence is not needed for the proof of L4.
7.2. Block decomposition. The Non-contraction property is just a slight modification of a condition with the same name introduced in [22]. Roughly speaking, this property states that the expansion of the iterations of vectors from the cones in $\left\{U_{y}, C_{y}\right\}_{y \in E}$ is uniformly bounded from below along orbits starting at $E$ and ending at $E \cup \mathcal{I}(E)$. This property implies that there is no loss of expansion along arbitrarily long sequences of consecutive collisions with focusing components or between flat components.

We now show that every sequence of collisions $\left\{z, \ldots, T^{m} z\right\}$ as in the non-contraction property can be decomposed into a finite number of special subsequences called blocks.

Definition 7.10. Let $z \in M \backslash R_{m}^{+}$with $m>0$. A sequence of consecutive collisions $\left\{z, \ldots, T^{m} z\right\}$ is called $a$ block of type $i \in\{1, \ldots, 4\}$ if the corresponding condition (i) below is satisfied:
(1) $z \in E^{+},\left\{T^{n(z)+1} z, \ldots, T^{m-1} z\right\} \subset M^{0}$ and $T^{m} z \in M^{-}$,
(2) $z \in M^{-},\left\{T z, \ldots, T^{m-1} z\right\} \subset M^{0} \cup M^{-}$and $T^{m} z \in M^{-}$,
(3) $z \in M^{-},\left\{T z, \ldots, T^{m-1} z\right\} \subset M^{0}$ and $T^{m} z \in E^{+}$,
(4) $z$ and $T^{m} z$ belong to $E^{+}$.

Definition 7.11. A block is called minimal if it does not contain any other block of the same type. A block included in a sequence of consecutive collisions $\varphi$ is called maximal in $\varphi$ if it is not contained in any other block of the same type in $\varphi$.

We observe that blocks of type 1 and 3 are always maximal and minimal. Also, we observe that every block of type 2 and 4 is a concatenation of finitely many minimal blocks of the type 2 and 4 , respectively.

Proposition 7.12. Let $\varphi=\left\{z, \ldots, T^{m} z\right\}$ be a sequence of collisions such that both $z$ and $T^{m} z$ belong to $E$. Then $\varphi$ is the concatenation of $\varphi_{1}, \ldots, \varphi_{n}$ with $1 \leq n \leq 5$ and $\varphi_{1}, \ldots, \varphi_{n}$ being maximal blocks of type 1-4. Moreover, this decomposition is unique.

Proof. First, suppose that $\varphi$ does not contain blocks of type 2 and 4. Then we see that either $\varphi$ is a block of type 1 or 3 , or $\varphi$ is the concatenation of two blocks: a block of type 3 followed by a block of type 1 .

Now, suppose that $\varphi$ contains blocks of type 2 , but does not contain blocks of type 4. A maximal block of type 2 can only be preceded by a block of type 1 and followed by a block of type 3 . Hence, $\varphi$ can contain at most two maximal blocks of type 2 . Moreover, if $\varphi$ contains exactly two maximal blocks of type 2 , then $\varphi$ is the concatenation of a maximal block of type 2 , a block of type 1 , a block of type 3 and a maximal block of type 2, following one another in this order. Instead, if $\varphi_{1}$ contains only a single maximal block of type 2 , then there are three possibilities: $\varphi$ is a block of type $2, \varphi$ is the concatenation of a maximal block of type 2 and a block of type 1 or 3 , and $\varphi$ is the concatenation of a block of type 1, a maximal block of type 2 and a block of type 3 , following one another in this order.

Finally, suppose that $\varphi$ contains blocks of type 4. From the definition of these blocks, we see that there is only one single maximal block of type 4 in $\varphi$. Such a block can only be preceded by a block of type 3 and followed by a block of type 1. Moreover, blocks of type 1 and 3 in $\varphi$ can only be adjacent to blocks of type 2 or 4 . Since there is only one maximal block of type 4 in $\varphi$, besides this block, $\varphi$ can contain at most a block of type 1, at most a block of type 3 and at most two maximal blocks of type 2 . The blocks of type 1 and 3 are concatenated to the block of type 4 , whereas one block of type 2 is concatenated to
the block of type 1, and the second block of type two is concatenated to the block of type 3 .

It is not difficult to see that the decomposition of $\varphi$ into blocks of type 1-4 we have just obtained is unique, because the blocks forming it are maximal.
7.3. Proving the Non-contraction property. In view of Proposition 7.12, it suffices to show that (12) holds along each block of type $1-4$ with $\beta^{\prime}$ depending only on the block type. We start with some preliminary results.
Lemma 7.13. Let $0 \leq m_{1}<m_{2}$, and suppose that $z \in E \backslash R_{m_{2}}^{+}$and $v \in C_{z}(z) \backslash\{0\}$. Then

$$
\frac{\left|D_{z} T^{m_{2}} v\right|_{J^{\prime}}}{\left|D_{z} T^{m_{1}} v\right|_{J^{\prime}}}=\prod_{k=m_{1}+1}^{m_{2}}\left|\frac{\tau_{k}^{-}(z, v)}{\tau_{k}^{+}(z, v)}\right|
$$

Proof. Define $z_{k}=T^{k} z$ and $v_{k}=D_{z} T^{k} v$ for $0 \leq k \leq n(z)$. By the definition of $C_{z}$ and its invariance, it is not difficult to see using Theorem 4.1 that $J^{\prime}\left(v_{k}\right) \neq 0$ for $0 \leq k \leq n(z)$. Using (15) and Condition B2, one can further show that $J\left(v_{k}\right) \neq 0$ for $1 \leq k \leq n(z)$. Therefore,

$$
\frac{\left|v_{m}\right|_{J^{\prime}}}{\left|v_{m_{1}}\right|_{J^{\prime}}}=\prod_{k=m_{1}+1}^{m}\left|\frac{J^{\prime}\left(v_{k}\right)}{J\left(v_{k}\right)}\right| \cdot\left|\frac{J\left(v_{k}\right)}{J^{\prime}\left(v_{k-1}\right)}\right| .
$$

Now, the wanted equality follows from $\tau^{+}\left(z_{k}, v_{k}\right)=-J\left(v_{k}\right) / J^{\prime}\left(v_{k}\right)$ and $\tau^{-}\left(z_{k}, v_{k}\right)=J\left(v_{k}\right) / J^{\prime}\left(v_{k-1}\right)$. The first expression for $\tau^{+}\left(z_{k}, v_{k}\right)$ is just (16), whereas the one for $\tau^{-}\left(z_{k}, v_{k}\right)$ can be easily derived from (15) and (16).

The following proposition provides an upper bound on the number of consecutive collisions $n(x)$ along a Donnay arc provided that the initial angle $\theta(x)$ is sufficiently close to 0 or $\pi$. For its proof see [16, Corollary 5.3 and Part (1) of Proposition 6.1].

Proposition 7.14. Suppose that $\Gamma_{i}$ is a Donnay arc, and let $\theta_{i}$ be the corresponding constant as in Theorem 4.1. There exist a real number $c_{i}>0$ and a function $N_{i}:(0, \pi / 2) \rightarrow \mathbb{N}$ such that if $x \in M_{i} \backslash A_{2}$, then $n(x) \leq c_{i} / \min \{\theta(x), \pi-\theta(x)\}$ whenever $\theta(x) \in\left(0, \theta_{i}\right) \cup\left(\pi-\theta_{i}, \pi\right)$, and $n(x)<N_{i}(\theta(x))$ whenever $\theta(x) \in\left[\theta_{i}, \pi-\theta_{i}\right]$.
Proposition 7.15. There exists $\gamma_{1}>0$ such that if $z \in E^{+}$and $0 \leq$ $m_{1}<m_{2} \leq n(z)$, then

$$
\left\|D_{z} T^{m_{2}} v\right\| \geq \gamma_{1}\left\|D_{z} T^{m_{1}} v\right\| \quad \text { for } v \in C_{z}(z)
$$

Proof. In virtue of Lemmas 7.2, 7.3 and 7.5 , we can prove the proposition with the norm $\|\cdot\|$ replaced by the semi-norm $|\cdot|_{J^{\prime}}$. Let $z \in E_{i} \subset E^{+}$for some $i$, and pick $v \in C_{z}(z)$. Of course, it is enough to prove the proposition with $\gamma_{1}$ depending on the focusing component
$\Gamma_{i}$. By taking the minimum of such $\gamma_{1}$ 's over all focusing components, we obtain the proposition in its general form. Let $\theta_{i}, a_{i}, c_{i}, N_{i}$ be as in Theorem 4.1 and Proposition 7.14. We define $z_{k}=T^{k} z, v_{k}=D_{z} T^{k} v$ for $0 \leq k \leq n(z)$.

Suppose first that $\theta=\theta(z) \in\left(0, \theta_{i}\right) \cup\left(\pi-\theta_{i}, \pi\right)$. We consider only the case when $\theta \in\left(0, \theta_{i}\right)$, because the argument for $\theta \in\left(\pi-\theta_{i}, \pi\right)$ is similar. By Theorem 4.1, we have $m\left(v_{k}\right) \geq-a_{i} \theta$. Let $R_{i}=\max _{y \in M_{i}} 1 / \kappa(y)$. Using (1), we easily obtain

$$
\frac{\tau^{-}\left(z_{k}, v_{k}\right)}{\tau^{+}\left(z_{k}, v_{k}\right)} \geq \frac{1-R_{i} a_{i} \theta}{1+R_{i} a_{i} \theta}
$$

By Proposition 7.14 , we have $n(z) \leq c_{i} / \theta$, and so Lemma 7.13 implies that

$$
\frac{\left|v_{m_{2}}\right|_{J^{\prime}}}{\left|v_{m_{1}}\right|_{J^{\prime}}} \geq\left(\frac{1-R_{i} a_{i} \theta}{1+R_{i} a_{i} \theta}\right)^{\frac{c_{i}}{\theta}}
$$

Since the right hand-side of the previous inequality converges to $e^{-2 a_{i} c_{i} R_{i}}>$ 0 as $\theta \rightarrow 0^{+}$, there exists $0<\bar{\theta}<\theta_{i}$ such that

$$
\begin{equation*}
\frac{\left|v_{m_{2}}\right|_{J^{\prime}}}{\left|v_{m_{1}}\right|_{J^{\prime}}} \geq \frac{1}{2} e^{-2 a_{i} c_{i} R_{i}} \quad \text { for } \theta \in(0, \bar{\theta}) . \tag{17}
\end{equation*}
$$

Now, suppose that $\theta(z) \in[\bar{\theta}, \pi-\bar{\theta}]$. By Part (3) of Theorem 4.1, there exists a constant $f_{i}>0$ such that $d\left(z_{k}\right) / 2 \leq \tau^{ \pm}\left(z_{k}, v_{k}\right) \leq f_{i}$ for $1 \leq k \leq m_{2}$. Let $r_{i}=\min _{y \in M_{i}} 1 / \kappa(y)$, and let $d_{i}=r_{i} \sin \theta_{i}$. Then

$$
\frac{\tau^{-}\left(z_{k}, v_{k}\right)}{\tau^{+}\left(z_{k}, v_{k}\right)} \geq \frac{d\left(z_{k}\right)}{2 f_{i}} \geq \frac{d_{i}}{2 f_{i}} \quad \text { for } 1 \leq k \leq m_{2} .
$$

Using Lemma 7.13 and Proposition 7.14, we conclude that

$$
\begin{equation*}
\frac{\left|v_{m_{2}}\right|_{J^{\prime}}}{\left|v_{m_{1}}\right|_{J^{\prime}}} \geq \min \left\{1,\left(\frac{d_{i}}{2 f_{i}}\right)^{N_{i}(\bar{\theta})}\right\} . \tag{18}
\end{equation*}
$$

The wanted conclusion now follows from (17) and (18).
Let $\lambda$ be as in B2, and let $f_{i}$ be as in the proof of Proposition 7.15. Define $f$ to be the maximum of all $f_{i}$ 's.
Lemma 7.16. Consider a sequence of collisions $\left\{z, \ldots, T^{m} z\right\}$ with $m>0$ such that $z, T^{m} z \in E$, and $T^{k} z \in M^{0}$ for $n(z)<k<m$. Then, for every $v \in C_{z}(z) \backslash\{0\}$ and every $n(z) \leq k<m$, we have

$$
\left|D_{z} T^{m} v\right|_{J^{\prime}}>\delta\left|D_{z} T^{k} v\right|_{J^{\prime}}
$$

where $\delta=\lambda / f$ if $T^{m} z \in M^{+}$, and $\delta=1$ if $T^{m} z \in M^{-}$.
Proof. Let $v \in C_{z}(z) \backslash\{0\}$, and define $z_{k}=T^{k} z$ and $v_{k}=D_{z} T^{k} v$ for $0 \leq k \leq m$. From $\left\{z_{n(z)+1}, \ldots, z_{m-1}\right\} \subset M^{0}$, it follows that $\left|v_{k}\right|_{J^{\prime}}=$ $\left|v_{m-1}\right|_{J^{\prime}}$ for all $n(z) \leq k<m$. Then, by Lemma 7.13, we have

$$
\frac{\left|v_{m}\right|_{J^{\prime}}}{\left|v_{k}\right|_{J^{\prime}}}=\frac{\left|v_{m}\right|_{J^{\prime}}}{\left|v_{m-1}\right|_{J^{\prime}}}=\frac{\tau^{-}\left(z_{m}, v_{m}\right)}{\tau^{+}\left(z_{m}, v_{m}\right)} \quad \text { for } n(z) \leq k<m \text {. }
$$

If $z_{m} \in M^{+}$, then $\tau^{-}\left(z_{m}, v_{m}\right)=t\left(z_{n(z)}, z_{m}\right)-\tau^{+}\left(z_{n(z)}, v_{n(z)}\right)>\lambda$ by Condition B2. Since $0<\tau^{+}\left(z_{m}, v_{m}\right) \leq f$ (see the proof of Proposition 7.15), we obtain $\tau^{-}\left(z_{m}, v_{m}\right) / \tau^{+}\left(z_{m}, v_{m}\right)>\lambda / f$. If $z_{m} \in M^{-}$, then we use the Mirror Formula (2) to relate the focusing times $\tau^{-}\left(z_{m}, v_{m}\right)$ and $\tau^{+}\left(z_{m}, v_{m}\right)$. A simple computation using $d\left(z_{m}\right)<0$ shows that $\tau^{-}\left(z_{m}, v_{m}\right)<\tau^{+}\left(z_{m}, v_{m}\right)<0$, and so $\tau^{-}\left(z_{m}, v_{m}\right) / \tau^{+}\left(z_{m}, v_{m}\right)>1$.
Lemma 7.17. Consider a sequence of collisions $\left\{z, \ldots, T^{m} z\right\}$ with $m>0$ such that $z \in M^{-}, T^{m} z \in E$ and $T^{k} z \in M^{0}$ for $0<k<m$. Then, for every $v \in C_{z}(z) \backslash\{0\}$ and every $0 \leq k<m$, we have

$$
\left|J\left(D_{z} T^{m} v\right)\right|>\left|J\left(D_{z} T^{k} v\right)\right| .
$$

Proof. Let $v \in C_{z}(z) \backslash\{0\}$. By the definition of $C_{z}$ (see Subsection 4.3), it is easy to see that $J(v) J^{\prime}(v) \geq 0$ and $J^{\prime}(v) \neq 0$. Since $\left\{T z, \ldots, T^{m-1} z\right\} \subset M^{0}$, we have $\left|J^{\prime}\left(D_{z} T^{k} v\right)\right|=\left|J^{\prime}(v)\right|$ for $0<k<m$. Using (15), we obtain

$$
\left|J\left(D_{z} T^{k} v\right)\right|=|J(v)|+l\left(z, T^{k} z\right)\left|J^{\prime}(v)\right| \quad \text { for } 0 \leq k \leq m,
$$

where $t\left(z, T^{k} z\right)$ is the length of the piece of trajectory starting at $z$ and ending at $T^{k} z$ (see the beginning of Subsection 6.1). The wanted conclusion follows immediately from the previous equality.
Lemma 7.18. There exists a constant $\delta_{1}>0$ such that for every sequence of consecutive collisions $\left\{z, \ldots, T^{m} z\right\}$ with $z \in E^{+}, T^{m} z \in E$ and $T^{k} z \in M^{0}$ for $n(z)<k<m$, we have

$$
\left\|D_{z} T^{m} v\right\| \geq \delta_{1}\left\|D_{z} T^{n(z)} v\right\| \quad \text { for } v \in C_{z}(z)
$$

Proof. If we replace $\|\cdot\|$ with $|\cdot|_{J^{\prime}}$, then the wanted conclusion with $\delta_{1}=\delta$ follows from Lemma 7.16. To obtain the actual conclusion, use the obvious fact that $\|\cdot\|_{J} \geq|\cdot|_{J^{\prime}}$, apply Lemma 7.3 to $\left\|D_{z} T^{m} v\right\|_{J}$, and finally apply Lemmas 7.2 and 7.5 to $\left|D_{z} T^{n(z)} v\right|_{J^{\prime}}$ and $\left\|D_{z} T^{n(z)} v\right\|_{J}$, respectively.
Lemma 7.19. There exists $\beta_{1}^{\prime}>0$ such that every sequence of consecutive collisions $\left\{z, \ldots, T^{m} z\right\}$ such that $z \in E^{+},\left\{T^{n(z)+1} z, \ldots, T^{m-1} z\right\} \subset$ $M^{0}$ and $T^{m} z \in E$ satisfies (12) with $\beta^{\prime}=\beta_{1}^{\prime}$. In particular, the previous conclusion is true for every block of type 1 .

Proof. Let $\varphi=\left\{z, \ldots, T^{m} z\right\}$ be a block of type 1 , and choose $0 \neq v \in$ $C_{z}(z)$. By Proposition 7.15 , we have $\left\|D_{z} T^{n(z)} v\right\| \geq \gamma_{1}\|v\|$. To complete the proof, use Lemma 7.18.
Lemma 7.20. There exists $\beta_{2}^{\prime}>0$ such that every block of type 2 satisfies (12) with $\beta^{\prime}=\beta_{2}^{\prime}$.
Proof. Note first that every block of type 2 consists of finitely many minimal blocks of type 2. Next, suppose that $\left\{z, \ldots, T^{m} z\right\}$ is a minimal block of type 2, and let $v \in C_{z}(z) \backslash\{0\}$. In this case, we have $T^{k} z \in M^{0}$ for $1 \leq k \leq m-1$. It follows that we have $\left|J^{\prime}\left(D_{z} T^{m} v\right)\right|>\left|J^{\prime}(v)\right|$
by Lemma 7.16, and $\left|J\left(D_{z} T^{m} v\right)\right|>|J(v)|$ by Lemma 7.17. Therefore, $\left\|D_{z} T^{m} v\right\|_{J}>\|v\|_{J}$. Of course, the same conclusion extends to a general block of type 2. To complete the proof, apply Lemmas 7.3 and 7.4 to $\left\|D_{z} T^{m} v\right\|_{J}$ and $\|v\|_{J}$, respectively.
Lemma 7.21. There exists $\beta_{3}^{\prime}>0$ such that every block of type 3 satisfies (12) with $\beta^{\prime}=\beta_{3}^{\prime}$.
Proof. Suppose that $\left\{z, \ldots, T^{m} z\right\}$ is a block of type 3, and let $v \in$ $C_{z}(z) \backslash\{0\}$. Since $T^{k} z \in M^{0}$ for $1 \leq k<m$, we have $\left|J^{\prime}\left(D_{z} T^{m} v\right)\right|>$ $\left|J^{\prime}(v)\right| \lambda / f$ by Lemma 7.16, and $\left|J\left(D_{z} T^{m} v\right)\right|>|J(v)|$ by Lemma 7.17. Hence, $\left\|D_{z} T^{m} v\right\|_{J}>\min \{\lambda / f, 1\}\|v\|_{J}$. To complete the proof, apply Lemmas 7.3 and 7.4 to $\left\|D_{z} T^{m} v\right\|_{J}$ and $\|v\|_{J}$, respectively.

Let $\left\{z, \ldots, T^{m} z\right\}$ be a block of type 4 . As before, we define $z_{j}=T^{j} z$ for all $0 \leq j \leq m$. We will use this notation throughout the rest of this section. Note that every block of type 4 consists of finitely many minimal blocks of type 4 . In other words, there exist $N>0$ integers $0=i_{0}<\cdots<i_{N}=m$ such that $\left\{z_{i_{k}}, \ldots, z_{i_{k+1}}\right\}$ is a minimal block of type 4 for each $0 \leq k \leq N-1$ (see Fig. 1). Now, recall that $n\left(z_{i_{k}}\right)$ denotes the number of consecutive collisions of $z_{i_{k}}$ with the focusing component $\Gamma_{i}$ before leaving it (see Definition 3.2). Define $j_{k}=n\left(z_{i_{k}}\right)+i_{k}$. Then

$$
D_{z} T^{m}=D_{z_{j_{N-1}}} T^{i_{N}-j_{N-1}} \circ D_{z_{j_{0}}} T^{j_{N-1}-j_{0}} \circ D_{z_{0}} T^{j_{0}}
$$

Proposition 7.22. If $N>1$, then the matrix of $D_{z_{0}} T^{j_{N-1}-j_{0}}$ in coordinates $\left(J, J^{\prime}\right)$ is given by

$$
D_{z_{j}} T^{j_{N-1}-j_{0}}=F_{N-1}^{-1} A_{N-1} B_{N-2} \cdots B_{1} A_{1} B_{0} F_{0}
$$

where $A_{k}, B_{k}, F_{k}$ are matrices of the form

$$
A_{k}= \pm\left(\begin{array}{cc}
1 / \zeta_{k} & 0  \tag{19}\\
\eta_{k} & \zeta_{k}
\end{array}\right), \quad B_{k}= \pm\left(\begin{array}{cc}
1+a_{k} & b_{k} \\
c_{k} & 1+d_{k}
\end{array}\right), \quad F_{k}=\left(\begin{array}{cc}
1 & f_{k} \\
0 & 1
\end{array}\right)
$$

with $\zeta_{k}, \eta_{k}, f_{k}>0, a_{k}, c_{k}, d_{k} \geq 0$ and $b_{k}>\lambda$.
Proof. For every $0 \leq k \leq N-1$, define

$$
s_{k}^{-}=G_{z_{i_{k}}}^{-}\left(z_{i_{k}}\right) \quad \text { and } \quad s_{k}^{+}=\sup \left\{\tau_{j_{k}-i_{k}}^{+}\left(z_{i_{k}}, v\right): 0 \neq v \in C_{z_{i_{k}}}\left(z_{i_{k}}\right)\right\}
$$

where $G_{z_{i_{k}}}^{ \pm}$are defined in Remark 4.2. Recall the definitions of the matrices $F$ and $R$ in (13). If we define

$$
\begin{aligned}
& A_{k}=F\left(s_{k}^{+}\right) D_{z_{i_{k}}} T^{j_{k}-i_{k}} R\left(z_{i_{k}}\right) F\left(s_{k}^{-}\right) \\
& B_{k}=F^{-1}\left(s_{k}^{-}\right) R^{-1}\left(z_{i_{k+1}}\right) D_{z_{j_{k}}} T^{i_{k+1}-j_{k}} F^{-1}\left(s_{k}^{+}\right) \\
& F_{k}=F\left(s_{k}^{+}\right)
\end{aligned}
$$

then we easily see that

$$
D_{z_{j_{k}}} T^{j_{k+1}-j_{k}}=F_{k+1}^{-1} A_{k+1} B_{k} F_{k} .
$$



Figure 1. Decomposition of a block of type 4. The collisions $z_{i_{k}} \in E^{+}$and $z_{j_{k}} \in \mathcal{I}\left(E^{+}\right)$are the first and the last in a sequence of consecutive collisions with a focusing arc, respectively. The sequence of collisions $\left\{z_{i_{k}}, \ldots, z_{i_{k+1}}\right\}$ is a minimal block of type 4 . The dashed piece of trajectory between the points $q_{1}^{+}$and $q_{N}^{-}$represents a sequence of collisions with finitely many boundary components of the billiard domain.

Therefore,

$$
D_{z_{j_{0}}} T^{j_{N-1}-j_{0}}=\prod_{k=0}^{N-2} D_{z_{j_{k}}} T^{j_{k+1}-j_{k}}=F_{N-1}^{-1} A_{N-1} B_{N-2} \cdots B_{1} A_{1} B_{0} F_{0}
$$

The form of the matrices $F_{k}$ can be immediately derived from (13). Since $f_{k}=s_{k}^{+}$, the positivity of $f_{k}$ follows from $s_{k}^{+}>0$. We now determine the form of the matrices $A_{k}$ and $B_{k}$. Let $v_{k} \in T_{z_{i_{k}}} M$ such that $\tau^{+}\left(z_{i_{k}}, v_{k}\right)=G_{z_{i_{k}}}^{+}\left(z_{i_{k}}\right)$, and denote by $q_{k}^{-}$and $q_{k}^{+}$the points along the trajectory of $z_{i_{k}}$ where the vector $v_{k}$ focuses backward, and the vector $D_{z_{i_{k}}} T^{j_{k}-i_{k}} v_{k}$ focuses forward, respectively. The matrix $A_{k}$ governs the dynamics of transversal Jacobi fields along the trajectory starting at $q_{k}^{-}$and ending at $q_{k}^{+}$(see Fig. 1). In other words, if a transversal Jacobi field is equal to $\left(J, J^{\prime}\right)$ at $q_{k}^{-}$, then the same field is equal to $A_{k}\left(J, J^{\prime}\right)^{T}$ at $q_{k}^{+}$.

By construction of the billiard cone field (see Theorem 4.1), every transversal Jacobi field focusing at $q_{k}^{-}$focuses again at $q_{k}^{+}$. Since $\tau^{+}$is a strictly decreasing function of $\tau^{-}$by the Mirror Formula, it follows that every transversal Jacobi field obtained from a parallel variation of the billiard trajectory passing through $q_{k}^{-}$focuses before reaching $q_{k}^{+}$. Thus, there exist $\xi_{k}, \zeta_{k}, \eta_{k}>0$ such that

$$
A_{k}\binom{1}{0}= \pm\binom{\xi_{k}}{\eta_{k}}, \quad A_{k}\binom{0}{1}= \pm\binom{ 0}{\zeta_{k}}
$$

Since $A_{k}$ is product of finitely many matrices as in (13), we have $\operatorname{det} A_{k}=1$, and so $\xi_{k}=\zeta_{k}^{-1}$. We conclude that

$$
A_{k}= \pm\left(\begin{array}{cc}
1 / \zeta_{k} & 0 \\
\eta_{k} & \zeta_{k}
\end{array}\right)
$$

To derive the form of $B_{k}$, we argue similarly. The matrix $B_{k}$ governs the dynamics of transversal Jacobi fields along the piece of the billiard trajectory starting at $q_{k}^{+}$and ending at $q_{k+1}^{-}$(see Fig. 1). By the definition of $q_{k}^{ \pm}$, we see that along this piece of trajectory, there are only collisions with flat or dispersing components. Hence, it easily follows from (15) that $B_{k}$ is a product of finitely many matrices of the form

$$
P= \pm\left(\begin{array}{cc}
1+a & b  \tag{20}\\
c & 1+d
\end{array}\right)
$$

with $a, b, c, d \geq 0$ and $\operatorname{det} P=1$. All the matrices $P$ forms a semigroup with respect to the standard matrix multiplication. For this reason, there are $a_{k}, b_{k}, c_{k}, d_{k} \geq 0$ such that

$$
B_{k}= \pm\left(\begin{array}{cc}
1+a_{k} & b_{k} \\
c_{k} & 1+d_{k}
\end{array}\right)
$$

Since $F$ and $R$ are upper and lower triangular, respectively, we see that the entry $b_{k}$ cannot be smaller than the length of the piece of the trajectory connecting $q_{k}^{+}$and $q_{k+1}^{-}$. This observation combined with Condition B2 gives $b_{k}>\lambda$.

The notation used in the following proposition is as in Proposition 7.22 and the paragraph before it.

Proposition 7.23. There exists $\beta_{4}^{\prime \prime}>0$ such that

$$
\left|F_{N-1} D_{z} T^{j_{N-1}} v\right|_{J^{\prime}} \geq \beta_{4}^{\prime \prime}\left|F_{0} D_{z} T^{j_{0}} v\right|_{J^{\prime}} \quad \text { for } v \in C_{z}(z)
$$

Proof. The inequality holds trivially if $N=1$. Therefore, we assume that $N>1$. Given a matrix $L$, denote by $|L|$ the matrix obtained by replacing each entry of $L$ with its absolute value. Given two square matrices $L_{1}$ and $L_{2}$ of the same order, we write $L_{1} \geq L_{2}$ if each entry of $L_{1}$ is greater than or equal to the corresponding entry of $L_{2}$. Let $\zeta=\zeta_{1} \cdots \zeta_{N-1}>0$. By the properties the matrices $A_{k}$ and $B_{k}$, it follows that

$$
\begin{aligned}
\left|A_{N-1} B_{N-2} \cdots A_{1} B_{0}\right| & =\left|A_{N-1}\right|\left|B_{N-2}\right| \cdots\left|A_{1}\right|\left|B_{0}\right| \\
& \geq\left|A_{N-1}\right|\left|A_{N-2}\right| \cdots\left|A_{1}\right|\left|B_{0}\right| \\
& \geq\left(\begin{array}{cc}
1 / \zeta_{N-1} & 0 \\
0 & \zeta_{N-1}
\end{array}\right) \cdots\left(\begin{array}{cc}
1 / \zeta_{1} & 0 \\
0 & \zeta_{1}
\end{array}\right)\left(\begin{array}{ll}
1 & \lambda \\
0 & 1
\end{array}\right) \\
& =\left(\begin{array}{cc}
1 / \zeta & \lambda / \zeta \\
0 & \zeta
\end{array}\right) .
\end{aligned}
$$

Let $v \in C_{z}(z)$. Also, define $v_{k}=D_{z} T^{k} v$ for $0 \leq k \leq m, w_{0}=F_{0} v_{j_{0}}$ and $w_{1}=F_{N-1} v_{j_{N-1}}$. By Proposition 7.22, we have

$$
w_{1}=A_{N-1} B_{N-2} \cdots B_{1} A_{1} B_{0} w_{0}
$$

From the construction of the cone field $\left\{\left(U_{y}, C_{y}\right)\right\}_{y \in E}$ and the definition of $F_{k}$, we easily see that $0 \leq J\left(w_{i}\right) / J^{\prime}\left(w_{i}\right) \leq f$ for $i=0,1$, where $f$ is the positive constant in the proof of Lemma 7.2. The fact that $J\left(w_{i}\right)$ and $J^{\prime}\left(w_{i}\right)$ have the same sign implies that

$$
\begin{equation*}
\left|J\left(w_{1}\right)\right| \geq \frac{1}{\zeta}\left|J\left(w_{0}\right)\right|+\frac{\lambda}{\zeta}\left|J^{\prime}\left(w_{0}\right)\right| \quad \text { and } \quad\left|J^{\prime}\left(w_{1}\right)\right| \geq \zeta\left|J^{\prime}\left(w_{0}\right)\right| \tag{21}
\end{equation*}
$$

From $0 \leq J\left(w_{i}\right) / J^{\prime}\left(w_{i}\right) \leq f$ and the inequality on the left hand-side of (21), we obtain

$$
\left|J^{\prime}\left(w_{1}\right)\right| \geq \frac{1}{f}\left|J\left(w_{1}\right)\right| \geq \frac{\lambda}{\zeta f}\left|J^{\prime}\left(w_{0}\right)\right|
$$

which together with the inequality on the right-hand sides of (21) gives

$$
\left|J^{\prime}\left(w_{1}\right)\right| \geq \frac{1}{2}\left(\zeta+\frac{\lambda}{\zeta f}\right)\left|J^{\prime}\left(w_{0}\right)\right|
$$

But $x+c^{2} / x \geq 2 c$ for every $x>0$ and every $c \in \mathbb{R}$, and so

$$
\left|J^{\prime}\left(w_{1}\right)\right| \geq\left(\frac{\lambda}{f}\right)^{1 / 2}\left|J^{\prime}\left(w_{0}\right)\right|
$$

The wanted inequality holds with $\beta_{4}^{\prime \prime}=(\lambda / f)^{1 / 2}$.
Let $z, \ldots$ as before Proposition 7.22.
Lemma 7.24. Let $\beta_{4}^{\prime \prime}>0$ be the constant in Proposition 7.23. Then

$$
\left\|D_{z} T^{j_{N-1}} v\right\| \geq \frac{\beta_{4}^{\prime \prime}}{\alpha_{1} \alpha_{2} \alpha_{4}}\left\|D_{z} T^{j_{0}} v\right\| \quad \text { for } v \in C_{z}(z)
$$

Proof. By Lemmas 7.2, 7.3 and 7.5, it suffices to prove the inequality with $\|\cdot\|$ replaced by $|\cdot|_{J^{\prime}}$. From the form of the matrices $F_{k}$, we see at once that $\left|v_{j_{0}}\right|_{J^{\prime}}=\left|w_{0}\right|_{J^{\prime}}$ and $\left|v_{j_{N-1}}\right|_{J^{\prime}}=\left|w_{1}\right|_{J^{\prime}}$. The notation here is as in the proof of Proposition 7.23. The wanted conclusion now follows from Proposition 7.23.
Proposition 7.25. There exists $\beta_{4}^{\prime}>0$ such that every block of type 4 satisfies (12) with $\beta^{\prime}=\beta_{4}^{\prime}$.
Proof. The notation is an in the proof of Proposition 7.22 and the paragraph before it. Accordingly, a block $\varphi$ of type 4 is given by $\varphi=$ $\left\{z_{i_{0}}, \ldots, z_{i_{N}}\right\}$. We write $\varphi=\varphi_{1} \cup \varphi_{2}$, where $\varphi_{1}=\left\{z_{i_{0}}, \ldots, z_{j_{N-1}}\right\}$ and $\varphi_{2}=\left\{z_{j_{N-1}}, \ldots, z_{i_{N}}\right\}$. It is easy to see that there are 2 integers $j_{N-1}<$ $k_{1} \leq k_{2} \leq i_{N}$ such that $\varphi_{2}=\psi_{1} \cup \psi_{2} \cup \psi_{3}$ with $\psi_{1}=\left\{z_{j_{N-1}}, \ldots, z_{k_{1}}\right\}$ being an orbit starting at $z_{j_{N-1}} \in \mathcal{I}\left(E^{+}\right)$and ending at $M^{-}$after a sequence of collisions with $M^{0}, \psi_{2}=\left\{z_{k_{1}}, \ldots, z_{k_{2}}\right\}$ being a block of type 2 , and $\psi_{3}=\left\{z_{k_{2}}, \ldots, z_{i_{N}}\right\}$ being a block of type 3 .

Proposition 7.15 and Lemma 7.24 give $\left\|v_{j_{0}}\right\| \geq \gamma_{1}\left\|v_{i_{0}}\right\|$ and $\left\|v_{j_{N-1}}\right\| \geq$ $\beta_{4}^{\prime \prime}\left\|v_{j_{0}}\right\|$, respectively. By Lemma 7.18, we obtain $\left\|v_{k_{1}}\right\| \geq \delta_{1}\left\|v_{j_{N-1}}\right\|$ with $\delta_{1}$ independent of $\psi_{1}$ and $v_{0}$. Since $\psi_{2}$ and $\psi_{3}$ are blocks of type 2 and 3, we have $\left\|v_{k_{2}}\right\| \geq \beta_{2}^{\prime}\left\|v_{k_{1}}\right\|$ and $\left\|v_{i_{N}}\right\| \geq \beta_{3}^{\prime}\left\|v_{k_{2}}\right\|$ by Lemmas 7.20 and 7.21. Putting all together, we obtain the wanted conclusion with $\beta_{4}^{\prime}=\beta_{2}^{\prime} \beta_{3}^{\prime} \beta_{4}^{\prime \prime} \gamma_{1} \delta_{1} /\left(\alpha_{1} \alpha_{2} \alpha_{4}\right)$.
Corollary 7.26. If $T^{m} z \in E$, then (12) is satisfied.
Proof. If $T^{m} x \in E$, then (12) follows from Lemmas 7.19-7.21 and Proposition 7.25.

Proposition 7.27. There exists a constant $\gamma_{2}>0$ such that if $z \in$ $E \backslash R_{m}^{+}$with $m>1,\left\{T^{n(z)+1} z, \ldots, T^{m-1} z\right\} \subset M^{0}, T^{j} z \in S_{1}^{-}\left(\epsilon_{0}\right)$ for some $n(z)<j<m$, and $T^{m} z \in E$, then

$$
\left\|D_{z} T^{m} v\right\| \geq \gamma_{2}\left\|D_{z} T^{j} v\right\| \quad \text { for } v \in C_{z}(z)
$$

Proof. Let $0 \neq v \in C_{z}(z)$, and define $z_{k}$ and $v_{k}$ for $0 \leq k \leq m$ as in the proof of Lemma 7.13. We study separately the two cases: $J\left(v_{j}\right) J^{\prime}\left(v_{j}\right) \geq$ 0 and $J\left(v_{j}\right) J^{\prime}\left(v_{j}\right)<0$. By Lemmas 7.3 and 7.7, it is enough to prove the desired inequality with $\|\cdot\|$ replaced by $\|\cdot\|_{J}$.
Suppose that $J\left(v_{j}\right) J^{\prime}\left(v_{j}\right) \geq 0$. By Lemma 7.16, we have $\left|J^{\prime}\left(v_{m}\right)\right|>$ $\delta\left|J^{\prime}\left(v_{j}\right)\right|$, and by arguing as in the proof of Lemma 7.17, one can easily prove that $\left|J\left(v_{m}\right)\right| \geq\left|J\left(v_{j}\right)\right|$. It follows at once that there exists $c>0$ independent of the sequence $\left\{z, \ldots, T^{m} z\right\}$ as in the hypotheses of the proposition and $v \in C_{z}(z)$ such that $\left\|v_{m}\right\|_{J} \geq c\left\|v_{j}\right\|_{J}$.

Suppose now that $J\left(v_{j}\right) J^{\prime}\left(v_{j}\right)<0$. This means that $v_{j}$ is focusing $\left(\tau^{+}\left(z_{j}, v_{j}\right)>0\right)$, and so $z \in M^{+}$. By Theorem 4.1, we have $\tau^{+}\left(z_{n(z)}, v_{n(z)}\right) \leq f$. Recall that $f=\max _{i} f_{i}$ (see the paragraph before Lemma 7.16). From $z_{k} \in M^{0}$ for $n(z)<k \leq j$, it follows that $\left|J^{\prime}\left(v_{j}\right)\right|=\left|J^{\prime}\left(v_{n(z)}\right)\right|$, and it is not difficult to see that $\left|J\left(v_{j}\right)\right| \leq$ $\left|J\left(v_{n(z)}\right)\right| \leq f\left|J^{\prime}\left(v_{j}\right)\right|$. Thus $\left\|v_{j}\right\|_{J} \leq \delta_{1}\left|v_{j}\right|_{J^{\prime}}$, where $\delta_{1}=\left(1+f^{2}\right)^{1 / 2}$. On the other hand, by applying Lemma 7.16 to $\left\{z, \ldots, T^{m} z\right\}$, we obtain $\left|v_{m}\right|_{J^{\prime}} \geq \delta\left|v_{j}\right|_{J^{\prime}}$. Therefore, using the trivial fact $\left\|v_{m}\right\|_{J} \geq\left|v_{m}\right|_{J^{\prime}}$, we obtain $\left\|v_{m}\right\|_{J} \geq \delta \delta_{1}^{-1}\left\|v_{j}\right\|_{J}$. This completes the proof, because $\delta$ and $\delta_{1}$ do not depend on the sequence $\left\{z, \ldots, T^{m} z\right\}$ and $v \in C_{z}(z)$.

Proposition 7.28. The Non-contraction property is satisfied.
Proof. Let $\varphi=\left\{z, \ldots, T^{m} z\right\}$ with $z \in E$ and $T^{m} z \in \mathcal{I}(E)$. Since $M^{-} \subset E$ and $\mathcal{I}\left(M^{-}\right)=M^{-}$, it is enough to prove the lemma for the case $T^{m} z \in \mathcal{I}\left(E^{+}\right)$. If $T^{m} z \in E^{+} \cap \mathcal{I}\left(E^{+}\right)$, then the Non-contraction property is proved in Corollary 7.26. So, suppose that $T^{m} z \notin E^{+}$. There are two cases: $\varphi \subset M_{i} \subset M^{+}$for some $i$, and $\varphi \not \subset M_{i}$ for every $i$ such that $M_{i} \subset M^{+}$. In the first case, we have $z \in E_{i}$ and $m=n(z)$. By Proposition 7.15, it follows that (12) is satisfied with $\beta^{\prime}=\gamma_{1}$. In the second case, we can write $\varphi=\varphi_{1} \cup \varphi_{2}$, where $\varphi_{1}=\left\{z, \ldots, T^{k} z\right\}$ with $z, T^{k} z \in E$, and $\varphi_{2}=\left\{T^{k} z, \ldots, T^{m} z\right\}$ with $T^{k} z \in E^{+}$and $T^{m} z \in$
$\mathcal{I}\left(E^{+}\right)$. Since $\varphi_{2}$ is an orbit of the type considered in the first case, we have $\left\|D_{T^{k} z} T^{m-k} v\right\| \geq \gamma_{1}\|v\|$ for $v \in C_{T^{k} z}\left(T^{k} z\right)$. Note that $\varphi_{1}$ is a piece of orbit as in (12) so that by Corollary 7.26, there exists $\beta_{5}^{\prime}>0$ independent of $\varphi_{1}$ such that $\left\|D_{z} T^{k} v\right\| \geq \beta_{5}^{\prime}\|v\|$ for $v \in C_{z}(z)$. Hence, we see again that (12) is satisfied with constant $\beta_{5}^{\prime} \gamma_{1}$. This together with the fact that $\beta_{5}^{\prime}$ and $\gamma_{1}$ do not depend on $\varphi$ completes the proof.
7.4. Conclusion of the proof of Condition L4. Since the billiard map $T$ is time-reversible, it is easy to check that the stable part of $L 4$ for $x$ is equivalent to the unstable part of L 4 but with $O$ replaced by $\mathcal{I}(O)$. Also, note that by Proposition 5.3, the invariant set $\Lambda \subset M$ where $T$ admits a local stable and an unstable manifold has full measure. For the definition of these manifolds, see Proposition A.3. The local stable manifold (resp. local unstable manifold) of $y \in \Lambda$ is denoted by $V^{s}(y)$ (resp. $\left.v^{u}(y)\right)$.

In view of the time-reversibility of $T$, we have $\mathcal{I}\left(V^{s}(y)\right)=V^{u}(\mathcal{I}(y))$ and $\mathcal{I}\left(V^{u}(y)\right)=V^{s}(\mathcal{I}(y))$ for every $y \in \Lambda$. Hence $\Lambda=\mathcal{I}(\Lambda)$. From these considerations, it follows that the unstable part of L4 with $O$ replaced by $E \cup \mathcal{I}(E)$, and $\Lambda_{x}$ replaced by $\Lambda$ implies the full L4 (stable and unstable parts).

We will also need the following technical lemma. Its proof is exactly as the one of $\left[14\right.$, Lemma 3.6]. The sets $W^{s}(y)$ and $W^{u}(y)$ are defined in Definition A. 5 .

Lemma 7.29. The set $\Lambda$ can be chosen so that it satisfies the following property: if $y \in \Lambda \cap E$ and $z \in W^{u}(y)$ (resp. $z \in W^{s}(y)$ ), then $T_{z} W^{u}(y) \subset C_{y}(y)\left(\right.$ resp. $\left.T_{z} W^{u}(y) \subset C_{y}^{\prime}(y)\right)$, where $\left\{\left(U_{y}, C_{y}\right)\right\}_{y \in E}$ is the family of cone fields defined in Subsection 4.3.

Proposition 7.30. Every point $x \in H$ satisfies Condition L4.
Proof. Suppose that $x \in H$, and let $(l, N, O, K)$ be the quadruple of $x$ as specified at the beginning of Subsection 6.2. Also, let $\epsilon_{0}$ as in Lemma 7.6. We will prove the unstable part of L 4 with $O$ replaced by $E \cup \mathcal{I}(E), \Lambda_{x}$ replaced by $\Lambda$ and $\epsilon=\epsilon_{0}$. As explained at the beginning of this subsection, this implies L4.

Suppose that $y \in \Lambda \cap(E \cup \mathcal{I}(E)), z \in(E \cup \mathcal{I}(E)) \cap W^{u}(y) \cap T^{k} S_{1}^{-}\left(\epsilon_{0}\right)$ for some $k>0$. We study separately the cases $T^{-k} z \in E$ and $T^{-k} z \notin$ $E$. For simplifying the notation, we will write $D_{i, j}$ for the restriction of $D_{T^{i} z} T^{j}$ to the tangent subspace of $W^{u}\left(T^{i} y\right)$ at $T^{i} z$ with $i, j \in \mathbb{Z}$.

Before continuing with the proof, we need to make a remark. First, that the cone field $(O, K)$ is equal to $\left(U_{y}, C_{y}\right)$ for a proper $y \in E$. Then, note that by Lemma 7.29 and the invariance of the cone fields $\left\{\left(U_{y}, C_{y}\right): y \in E\right\}$, the tangent space of any unstable manifold considered in this proof at a point $z \in E$ is always contained in the cone $C_{z}(z)$. This property is essential for applying Propositions 7.27 and
7.28 , and will be used implicitly in this proof every time that one of the two propositions is applied.

We can now resume the proof. First, suppose that $T^{-k} z \in E$. Using Proposition 7.28 , we immediately obtain that $\left\|D_{0}^{-k}\right\| \leq 1 / \beta^{\prime}$, where $\beta^{\prime}$ is the constant appearing in (12).

Now, suppose that $T^{-k} z \notin E$. Then, there are 2 possibilities, either $T^{-k} z \in M^{0} \cap S_{1}^{-}\left(\epsilon_{0}\right)$ or $T^{-k} z \in M^{+} \backslash E$. We analyze each possibility individually. First, define $m_{\mp}=\inf \left\{i>0: T^{-k \mp i} z \in E\right\}$. Assume that $T^{-k} z \in M^{0} \cap S_{1}^{-}\left(\epsilon_{0}\right)$. Applying Proposition 7.27 to $T^{-k-m_{-} z}$ with $m=m_{-}+m_{+}$and $j=m_{-}$, we obtain $\left\|D_{-k+m_{+},-m_{+}}\right\| \leq 1 / \gamma_{2}$. Also, note that $\left\|D_{0,-k-m_{+}}\right\|$is trivially equal to 1 if $m_{+}=k$, and is not larger than $1 / \beta^{\prime}$ if $m_{+}<k$ by Proposition 7.28. Combining the previous observations, we conclude that $\left\|D_{0,-k}\right\| \leq 1 /\left(\gamma_{2} \min \left\{1, \beta^{\prime}\right\}\right)$.

Finally, assume that $T^{-k} z \in M^{+} \backslash E$. Let $n=n\left(T^{-k} z\right)$. By breaking the sequence $\{-k, \ldots, 0\}$ into $\{-k, \ldots,-k+n\},\left\{-k+n, \ldots,-k+m_{+}\right\}$ and $\left\{-k+m_{+}, \ldots, 0\right\}$, we see that

$$
D_{0,-k}=D_{-k+n,-n} \circ D_{-k+m_{+},-m_{+}+n} \circ D_{0,-k+m_{+}} .
$$

By Proposition 7.28, Lemma 7.18 and Proposition 7.15, we then obtain $\left\|D_{0}^{-k+m_{+}}\right\| \leq 1 / \beta^{\prime},\left\|D_{-k+m_{+},-m_{+}+n}\right\| \leq 1 / \delta_{1}$ and $\left\|D_{-k+n,-n}\right\| \leq 1 / \gamma_{1}$. Combining these inequalities, we conclude that $\left\|D_{0,-k}\right\| \leq 1 /\left(\beta^{\prime} \gamma_{1} \delta_{1}\right)$.
To complete the proof, we observe that the upper bounds of $\left\|D_{0,-k}\right\|$ for the different cases studied above do not depend on the data $y, z, k$ of Condition L4.

## 8. Donnay billiards and their generalizations

We now apply Theorem 5.6 to a large class of hyperbolic billiards. The purpose of this section is to provide concrete examples of new billiards with the Bernoulli property. For the definition of Bernoulli component, see Appendix A.

Definition 8.1. A polygon with pockets and bumps is a planar domain $\Omega$ obtained from a polygon P by replacing a sufficiently small neighborhood of each vertex with internal angle less than $\pi$ with a focusing arc (pocket) or a dispersing arc (bump), and of each vertex with internal angle greater than $\pi$ by a dispersing arc so that the vertex lies outside the closure of $\Omega$ (see Fig. 2).

Stadium-like domains with either parallel or non-parallel segments can be considered degenerate polygons with pockets. These domains correspond to quadrilaterals with a pocket or a bump replacing two vertexes instead of one. The results proved in this subsection apply also to billiards in stadium-like domains.

Note that polygons with pockets and bumps satisfy B1 by definition. In the rest of this section, we will assume that billiards in polygons


Figure 2. Polygon with pockets and bumps (solid curve) and original polygon (dashed curve).
with pockets and bumps satisfies Condition B2. This can achieved, for example, by using sufficiently short focusing arcs as pockets.

Proposition 8.2. If a billiard in a polygon with pockets and bumps satisfies Condition B2, then it satisfies Conditions B3 and B4 as well. In particular, $T$ has non-zero Lyapunov exponents a.e. on $M$.

Proof. A general result for polygonal billiards states that every semiorbit of a polygon billiard is either periodic or accumulates at least at one vertex of the polygon [18]. From this result and the assumption on the geometry of $\Omega$, it follows that if $x \in N^{-} \cup N^{+}$, then $x$ has to be periodic. Hence, $N^{-}=N^{+}=N$ consists of periodic orbits. In a polygonal billiard, every periodic orbit is contained in a family of parallel orbits having the same period. This family consists of finitely many segments contained in horizontal segments $(\theta=$ const $)$ with endpoint belonging to $R$. Since the number of distinct strips in a polygonal billiard is countable, we immediately obtain $\mu(N)=0$. Hence, B3 is satisfied. The fact that $N$ consists of periodic orbits implies also that $N^{\prime}=\emptyset$. Thus, $S_{1}^{-} \cap N^{+} \subset N^{\prime}=\emptyset$, and so Condition B4 is satisfied as well.
The second part of the proposition follows from Proposition 5.3.
Chernov and Troubetzkoy proved that a billiard in convex polygons with pockets $\Omega$ is ergodic if the pockets are arcs of circles with the full circles contained in $\Omega$ [11]. Here, we consider the more general situation of polygons with pockets and bumps. The condition in Definition 8.1 that each vertex with internal angle greater than $\pi$ is replaced by a dispersing arc so that the vertex lies outside the closure of the polygon plays a crucial role in the proof of the next theorem.

Theorem 8.3. The map $T$ of a billiard in a polygon with pockets and bumps that satisfies Condition B2 has the Bernoulli property.

Proof. Conditions B3 and B4 are satisfied by Proposition 8.2. By Theorem 5.6, every point of $H$ has neighborhood contained $(\bmod 0)$ in one

Bernoulli component. Now, suppose that $M_{i} \subset M^{-} \cup M^{+}$. Since $R$ is countable (see [15, Propositions 6.17-6.19]), $N^{\prime}=\emptyset$ and $N \cap M_{i}=\emptyset$, the set $H \cap M_{i}$ is connected. Hence, every set $M_{i} \subset M^{-} \cup M^{+}$is contained $(\bmod 0)$ in one Bernoulli component. We cannot claim the same for sets $M_{i} \subset M^{0}$, because in this case ${ }^{5}$, $N$ may disconnect $H \cap M_{i}$. Thus, to prove that $T$ is Bernoulli, we cannot use Corollary 5.8.

The alternative approach that we take is quite simple: we prove that the entire set $M^{-} \cup M^{+}$is contained $(\bmod 0)$ in one Bernoulli component. This together with $\mu(N)=0$ implies that $T$ is Bernoulli. Indeed, by Theorem A.6, the number of Bernoulli components of $T$ is equal to the minimum integer $n>0$ such that $T^{n} B=B$ for some Bernoulli component $B$. By the geometry of $\Omega$, it is always possible to find two distinct curved components $\Gamma_{i}$ and $\Gamma_{j}$ such that $\mu\left(M_{j} \cap T M_{i}\right)>$ 0 . Thus, if $B$ is the Bernoulli component containing $M^{-} \cup M^{+}$, then $T B=B$, and so $n=1$.

To prove that $M^{-} \cup M^{+}$is contained $(\bmod 0)$ in a single Bernoulli component, we start by observing that the polygon $P$ decomposes into finitely many triangles with pairwise disjoint interiors. The fact that this decomposition is not unique is irrelevant for our purposes. Let $\Delta$ be one of triangles of the decomposition. By construction of $\Omega$, every vertex with internal angle greater than $\pi$ lies outside $\Omega$. Then, we claim that the union of the sets $M_{i}$ corresponding to the pockets and bumps attached to the vertexes of $\Delta$ are contained $(\bmod 0)$ in the same Bernoulli component. In fact, let $\Gamma_{i}, \Gamma_{j}, \Gamma_{k}$ be the pockets or bumps attached at the vertexes of $\Delta$. If $B$ is the Bernoulli component containing $M_{i}$, then by Theorem A. 6 states that there exists a Bernoulli component $B^{\prime}$ such that $T B=B^{\prime}$. It is easy to see that $\mu\left(M_{j} \cap\right.$ $\left.T M_{i}\right)>0$ and $\mu\left(M_{k} \cap T M_{i}\right)>0$. Thus, $M_{j}$ and $M_{k}$ must belong to $B^{\prime}$. By symmetry of the configuration, also $M_{i}$ and $M_{j}$ belong the same Bernoulli component. We can then conclude that $M_{i}, M_{j}, M_{k}$ belong to the same Bernoulli component. In a stadium-like domain, the previous argument does not work, because there are only 2 curved components $\Gamma_{i}$ and $\Gamma_{j}$ attached to the 3 vertexes of $\Delta$. However, in this case, we have i) $\mu\left(M_{j} \cap T M_{i}\right)>0$ and $\mu\left(M_{j} \cap T M_{i}\right)>0$, and ii) $\mu\left(M_{j} \cap T^{2} M_{i}\right)>0$. Claim i) is obvious. Claim ii) follows from the fact that there is a positive measure set of orbits starting at $M_{i}$ and reaching $M_{j}$ after one collision with flat components of $\partial \Omega$. Since each $M_{i}$ and $M_{j}$ is contained in a Bernoulli component, and all the Bernoulli components are cyclically permuted with the same period $n>0$ by Theorem A. 6 , claim i) implies that $n \leq 2$, whereas claim ii) implies that $n$ is either 3 or one of its divisors. Hence $n=1$. The same conclusion is clearly true for all the sets $M_{i}$ corresponding to pockets and bumps attached to vertexes of two adjacent triangles $\Delta_{1}$

[^5]and $\Delta_{2}$. Hence, the entire set $M^{-} \cup M^{+}$belongs $(\bmod 0)$ to a Bernoulli component.

Donnay billiards are billiards in convex polygons with pockets [16]. The next result then follows directly from Theorem 8.3.
Corollary 8.4. Donnay billiards have the Bernoulli property.
The conclusion of Theorem 8.3 is robust for sufficiently small perturbations of the pockets and bumps.
Definition 8.5. Two polygons with pockets and bumps $\Omega_{1}$ and $\Omega_{2}$ are $\epsilon$-close if $\Omega_{2}$ is obtained from $\Omega_{1}$ by replacing each pocket (resp. bump) with another pocket (resp. bump) of the same length, $\epsilon$-close in the $C^{6}$-topology (resp. $C^{3}$-topology) and $\epsilon$-close in the Hausdorff distance.

Proposition 8.6. Suppose that the billiard in a polygon with pockets and bumps $\Omega_{1}$ satisfies B2. Then there exists $\epsilon>0$ such that if $\Omega_{2}$ is $\epsilon$-close to $\Omega_{1}$, then the billiard in $\Omega_{2}$ is Bernoulli.
Proof. Condition B2 is an open condition. Thus, if $\epsilon$ is sufficiently small, then the billiard in $\Omega_{2}$ satisfies B 2 , and so it is Bernoulli by Theorem 8.3.

We conclude this section with a result concerning hyperbolic billiards with domains for which $\Gamma^{0}=\emptyset$, i.e., without straight boundary components. For these billiards, the Bernoulli property follows quite directly from Corollary 5.8. We observe that Sinai billiards [28] - those with domains bounded only by strictly convex outwards arcs - belong to this class of billiards. For them, the Bernoulli property was first proved in [17].
Theorem 8.7. Let $\Omega$ be a billiard domain without straight boundary components, and suppose that its map $T$ satisfies Conditions B1-B4. Then $T$ is Bernoulli.

Proof. For billiards satisfying the hypotheses of the theorem, the set $R$ is countable (see [15, Propositions 6.17-6.19]), and we trivially have $N=N R=\emptyset$. Thus, each set $H \cap M_{i}$ is connected, and B5 is satisfied. The wanted conclusion now follows from Corollary 5.8.

## Appendix A. Local Ergodic Theorem

We state the Local Ergodic Theorem proved in [14], and recall the relevant definitions. The formulation of this theorem in its general form requires a series of technical definitions, which are not needed for 2-dimensional billiards. Thus, to avoid unnecessary technicalities, we specialize the presentation of the Local ergodic Theorem to 2-dimensional billiards. Accordingly, $T$ will denote the billiard map for some planar domain $\Omega$ throughout this appendix.

The definition of a cone field and related notions are given in Section 4.

Definition A.1. A point $x \in M \backslash \partial M$ is called sufficient if there exist
i) an integer $l$ such that $T^{l}$ is a local diffeomorphism at $x$,
ii) a neighborhood $O$ of $T^{l} x$ and an integer $N>0$ such that $O$ and $R_{N}^{-}$are disjoint,
iii) an invariant continuous cone field $K$ on $O \cup T^{-N} O$ such that $\sigma_{K}\left(D_{y} T^{N}\right)>3$ for every $y \in T^{-N} O$.
Every time we need to emphasize the role of the data l, N, O,K in this definition, we will write that $x$ is a sufficient point with quadruple $(l, N, O, K)$.
Definition A.2. A point $x \in M \backslash \partial M$ is called $u$-essential if for every $\alpha>0$, there exist $n_{x, \alpha} \in \mathbb{N}$, a neighborhood $O_{x, \alpha}$ of $x$ with $O_{x, \alpha} \cap$ $R_{n_{x, \alpha}}^{+}=\emptyset$ and a continuous invariant cone field ( $O_{x, \alpha} \cup T^{n_{x, \alpha}} O_{x, \alpha}, K_{x, \alpha}$ ) such that $\sigma_{K_{x, \alpha}}^{*}\left(D_{y} T^{n_{x, \alpha}}\right)>\alpha$ for every $y \in O_{x, \alpha}$. Analogously, a point $x \in M \backslash \partial M$ is called $s$-essential if, in the previous definition, $T$ and $R_{n_{x, \alpha}}^{+}$are replaced by $T^{-1}$ and $R_{n_{x, \alpha}}^{-}$, respectively.

The cone field $(O, K)$ in Definition A. 1 is eventually strictly invariant. By a well-known result [23, 31, 32], it follows that all the Lyapunov exponents of $T$ are non-zero a.e. on the set $\bigcup_{k \in \mathbb{Z}} T^{k} O$. This fact combined with the Katok-Strelcyn theory [20] gives Proposition A. 3 below (Part (3) is proved in [14, Proposition 5.3]). For the definition of absolute continuity of a foliation, we refer the reader to [10, 20].
Proposition A.3. Let $x \in M \backslash \partial M$ be a sufficient point with quadruple $(l, N, K, O)$. Then there exist an invariant set $\Lambda_{x} \subset \bigcup_{k \in \mathbb{Z}} T^{k} O$ with $\mu\left(\Lambda_{x}\right)=\mu\left(\bigcup_{k \in \mathbb{Z}} T^{k} O\right)>0$ and two families $V^{s}=\left\{V^{s}(y)\right\}_{y \in \Lambda_{x}}$ and $V^{u}=\left\{V^{u}(y)\right\}_{y \in \Lambda_{x}}$ consisting of $C^{2}$ submanifolds such that for every $y \in \Lambda_{x}$, we have
(1) $V^{s}(y) \cap V^{u}(y)=\{y\}$,
(2) $V^{s}(y)$ and $V^{u}(y)$ are embedded open intervals,
(3) $T_{y} V^{s}(y) \subset K^{\prime}(y)$ and $T_{y} V^{u}(y) \subset K(y)$ provided that $y \in O \cup$ $T^{-N} O$,
(4) $T V^{s}(y) \subset V^{s}(T y)$ and $T^{-1} V^{u}(y) \subset V^{u}\left(T^{-1} y\right)$,
(5) $d\left(T^{n} y, T^{n} z\right) \rightarrow 0$ exponentially as $n \rightarrow+\infty$ for every $z \in$ $V^{s}(y)$, and the same is true as $n \rightarrow-\infty$ for every $z \in V^{u}(y)$,
(6) $V^{s}(y)$ and $V^{u}(y)$ vary measurably with $y \in \Lambda_{x}$,
(7) the families $V^{s}$ and $V^{u}$ have the absolute continuity property.

Definition A.4. The submanifolds forming the families $V^{s}$ and $V^{u}$ are called local stable manifolds and local unstable manifolds, respectively.

Definition A.5. Let $x$ be a sufficient point of $M \backslash \partial M$, and let $\Lambda_{x}$ be the set in Proposition A.3. For every $y \in \Lambda_{x}$, we denote by $W^{u}(y)$ the connected component of

$$
\bigcup_{k \geq 0} T^{k} V^{u}\left(T^{-k} y\right)
$$

containing $y$. Analogously, denote by $W^{s}(y)$ the set obtained by replacing $T$ with $T^{-1}$ and $V^{u}$ with $V^{s}$ in the definition of $W^{u}(y)$.

We now recall the Spectral Decomposition Theorem. It applies to a larger class of hyperbolic system with singularities than billiards, but here we formulate it only for billiards. For its proof, one has to combine two results: [20, Theorem 13.1, Part II], which extends Pesin'result for smooth systems [27] to systems with singularities, and [9, Theorem 3.1]. See also [26], for results similar to those of [9].

Theorem A.6. Suppose that the billiard map $T$ has non-vanishing Lyapunov exponents a.e. on $M$. Then there exist at most countably many pairwise disjoint measurable subsets $E_{0}, E_{1}, \ldots$ of $M$ such that
(1) $M=\bigcup_{i} E_{i}$,
(2) $\mu\left(E_{0}\right)=0$, and $\mu\left(E_{i}\right)>0$ for every $i$,
(3) $T E_{i}=E_{i}$, and $\left(\left.T\right|_{E_{i}},\left.\mu\right|_{E_{i}}\right)$ is ergodic for every $i$,
(4) for every $i$, there exist $m_{i} \in \mathbb{N}$ pairwise disjoint measurable subsets $B_{i, 1}, \ldots, B_{i, m_{i}}, B_{i, m_{i}+1}=B_{i, 1}$ of $M$ such that $E_{i}=$ $\bigcup_{j=1}^{m_{i}} B_{i, j}, T B_{i, j}=B_{i, j+1}$ and $\left(\left.T^{m_{i}}\right|_{B_{i, j}},\left.\mu\right|_{B_{i, j}}\right)$ is Bernoulli for every $j=1, \ldots, m_{i}$.

The sets $E_{i}$ and $B_{i, j}$ are called an ergodic component of $T$ and a Bernoulli component of $T$, respectively. These sets are uniquely defined up to a set of zero measure.

We need one last definition before formulating the Local Ergodic Theorem.

Definition A.7. Let $\left(O_{1}, C_{1}\right)$ and $\left(O_{2}, C_{2}\right)$ be two cone fields. We say that $\left(O_{1}, C_{1}\right)$ and $\left(O_{2}, C_{2}\right)$ are jointly invariant if $D_{x} T^{k} C_{1}(x) \subset$ $C_{2}\left(T^{k} x\right)$ for every $x \in O_{1}$ and $k>0$ such that $T^{k} x \in O_{2}$, and $D_{x} T^{k} C_{2}(x) \subset C_{1}\left(T^{k} x\right)$ for every $x \in O_{2}$ and $k>0$ such that $T^{k} x \in O_{1}$.

Note that in the previous definition, we neither require that the sets $O_{1}$ and $O_{2}$ are disjoint nor that the cone fields $C_{1}$ and $C_{2}$ are invariant. However, it is easy to see that $C_{1}$ and $C_{2}$ are invariant in the following sense: if $x \in O_{1}$ and $k_{2}>k_{1}>0$ such that $T^{k_{1}} x \in O_{2}$ and $T^{k_{2}} x \in O_{1}$, then $D_{x} T^{k_{2}} C_{1}(x) \subset C_{1}\left(T^{k_{2}} x\right)$. The same is true for $C_{2}$, once $O_{1}$ has been replaced by $O_{2}$.

Definition A.8. A subset $\Sigma \subset M$ is called regular if it is a union of finitely many arcs $\Sigma_{1}, \ldots, \Sigma_{k}$ of class $C^{2}$ that can only intersect at their boundaries. The arcs $\Sigma_{1}, \ldots, \Sigma_{k}$ are called the components of $\Sigma$.

Definition A. 9 (Regularity). We say that $T$ satisfies Condition L1 if the singular sets $R_{k}^{+}$and $R_{k}^{-}$are regular for every $k>0$.

In the rest of this subsection, we assume that $x \in M \backslash \partial M$ is a sufficient point with quadruple ( $l, N, O, K$ ). Let $\Lambda_{x}$ be the subset associated to $x$ as in Proposition A.3.

Definition A. 10 (Alignment). We say that $x$ satisfies Condition L2 if the sets $O \cap R_{k}^{+}$and $O \cap R_{k}^{-}$are regular for every $k>0$, and the tangent subspace ${ }^{6} T_{y} \Sigma$ is contained in $K(y)$ (resp. $K^{\prime}(y)$ ) for every $k>0$, every component $\Sigma$ of $O \cap R_{k}^{-}$(resp. $O \cap R_{k}^{+}$) and every $y \in \Sigma \cap T^{-N} O$ (resp. $\Sigma \cap O)$.
Let $m_{+}$and $m_{-}$be the 1-dimensional Riemannian volume on $S_{1}^{+}$and $S_{1}^{-}$), respectively.

Definition A. 11 (Ansatz). We say that $x$ satisfies Condition L3 if the set of u-essential points of $S_{1}^{-}$(resp. s-essential points of $S_{1}^{+}$) has full $m_{-}$-measure (resp. $m_{+}$-measure), and if $y$ is any of such points, then the cone fields $(O, K)$ and $\left(O_{y, \alpha}, K_{y, \alpha}\right)$ are jointly invariant for every $\alpha>0$.

Given $A \subset M$ and $\epsilon>0$, we call the set $A(\epsilon)=\{x \in M: d(x, A)<$ $\epsilon\}$ the $\epsilon$-neighborhood of $A$.

Definition A. 12 (Contraction). We say that $x$ satisfies Condition L4 if there exist $\beta>0$ and $\epsilon>0$ such that

$$
\left\|\left.D_{z} T^{-k}\right|_{T_{z} W^{u}(y)}\right\| \leq \beta \quad\left(\text { resp } . \quad\left\|\left.D_{z} T^{k}\right|_{T_{z} W^{s}(y)}\right\| \leq \beta\right)
$$

for every $y \in O \cap \Lambda_{x}$ and every $z \in O \cap W^{u}(y) \cap T^{k} S_{1}^{-}(\epsilon)$ (resp. $\left.O \cap W^{s}(y) \cap T^{-k} S_{1}^{+}(\epsilon)\right)$ with $k>0$.
Theorem A. 13 (Local Ergodic Theorem). Suppose that L1 is satisfied, and that $x \in M \backslash \partial M$ is a sufficient point satisfying L2-L4. Then there exists a neighborhood of $x$ contained up to a set of zero $\mu$-measure in a Bernoulli component of $T$.

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[^1]:    ${ }^{1}$ There is no one-to-one correspondence between the elements of $M$ and the unit vectors of $\mathbb{R}^{2}$ attached to $\partial \Omega$. Indeed, two distinct elements $x_{1}=\left(s_{1}, \theta_{1}\right) \in \Gamma_{i}$ and $x_{2}=\left(s_{2}, \theta_{2}\right) \in \Gamma_{j}$ of $M$ may correspond to the same unit vector of $\mathbb{R}^{2}$. However, this can happen only if $\gamma_{i}\left(s_{1}\right)=\gamma_{j}\left(s_{2}\right)$ is a corner point of $\partial \Omega$.

[^2]:    ${ }^{2}$ More precisely, for billiards satisfying Conditions B1 and B2 in Section 5.

[^3]:    ${ }^{3} \mathrm{~A}$ single continuous cone field on $E$ exists for Bunimovich and Wojtkowski billiards $[1,31,32]$, but these form a subset of the billiards considered here.

[^4]:    ${ }^{4}$ In general, we have $\rho=\frac{b-\bar{b}}{b-a} \cdot \frac{\bar{a}-a}{b-\bar{a}}$. Here, we have used $b=+\infty$.

[^5]:    ${ }^{5}$ This is indeed the case whenever the polygon $P$ has two parallel sides facing each other.

[^6]:    ${ }^{6}$ In the general version of the theorem, the tangent space $T_{y} \Sigma$ has to be replaced by its Lagrangian skew-orthogonal complement with respect to the symplectic form $\omega$.

