# Non-Uniformly Hyperbolic Horseshoes Arising from Bifurcations of Poincaré Heteroclinic Cycles 

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#### Abstract

The purpose of this paper is to advance the knowledge of the dynamics arising from the creation and subsequent bifurcation of Poincaré heteroclinic cycles. The problem is central to dynamics: it has to be addressed if, for instance, one aims at describing the typical orbit behaviour of a typical system, thus providing a global scenario for the ensemble of dynamical systems - see the Introduction and [P1, P2]. Here, we shall consider smooth, i.e. $C^{\infty}$, one-parameter families of dissipative, meaning non-conservative, surface diffeomorphisms. An hetereoclinic cycle may appear when the parameter evolves and an orbit of tangency, say quadratic, is created between stable and unstable manifolds (lines) of periodic orbits that belong to a basic hyperbolic set. The key novelty is to allow this basic set, a horseshoe, to have Hausdorff dimension bigger than one. In the present paper we do assume such a dimension to be beyond one, but in a limited way, as explicitly indicated in the Introduction. [A mild non-degeneracy condition on the family of maps is assumed: at the orbit of tangency the invariant lines, stable and unstable, cross each other with positive relative speed]. We then prove that most diffeomorphisms, corresponding to parameter values near the bifurcating one, are non-uniformly hyperbolic in a neighborhood of the horseshoe and the orbit of tangency; such diffeomorphisms display no attractors nor repellors in such a neighborhood. A first precise formulation of our main theorem is at the Introduction and a more encompassing version at the end of the paper. These results were announced in [PY3].


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## 1 Introduction

### 1.1 The Context

Since Poincaré referred to a property valid "pour la plupart des coefficients" of a polynomial (analytic) dynamical system, the outstanding problem of describing at large the orbits of a "typical" system became the source of much creative work in dynamics.

A stumbling block in a possible program to solve this question is the understanding of dynamics arising from bifurcations of homoclinic or heteroclinic cycles. Such cycles were defined by Poincaré himself: they involve stable and unstable manifolds of invariant sets (typically periodic orbits) that successively intersect each other. In his classic book on Celestial Mechanics, he prophetically stated: "Rien n'est plus propre à nous donner une idée de la complication de tous les problèmes de Dynamique".

In fact, the creation and unfolding of cycles, in particular homoclinic tangencies for surface diffeomorphisms and their unfolding, led Newhouse to show the non-denseness of hyperbolic dynamics, thus contradicting Smale's remarkable conjecture of the early 60's. That is, there are systems that cannot be approximated by one with a hyperbolic limit set. On the way, Newhouse showed that in this context always appear surface diffeomorphisms displaying infinitely many simultaneous sinks (periodic attractors) or sources (periodic repellors).

Abundance of other more intricate kind of attractors, the so called Hénon-like ones, was proved to be also present in the unfolding of such cycles. This was another striking fact. It resulted from the works of Benedicks-Carleson, Mora-Viana and Colli. Attractors here mean invariant sets that attract future orbits of points of a positive Lebesgue measure set in the phase space (space of events).

In view of all these intricacies inherent to homoclinic and heteroclinic bifurcations, a new global conjecture has been proposed in [P1] (see also [P2]) concerning a typical dynamical system: In particular, systems with finitely many attractors should be dense in the universe of dynamics, i.e. $C^{r}$ flows, diffeomorphisms and maps, with $r \geq 1$. Also, their basins of attraction should cover the whole phase space, except for a Lebesgue zero measure set.


Figure 1

An example of the creation of a heteroclinic cycle associated to a (hyperbolic) horseshoe is indicated in Figure 1. Initially we have a classic Smale's horseshoe map (diffeomorphism) on the two-sphere $S^{2}$ with two saddle fixed points $p_{s}, p_{u}$ with positive eigenvalues, a fixed point repellor outside the figure and two fixed points attractors $s_{1}$ and $s_{2}$. The rectangle inside the figure is sent by the map to the snake-shaped piece, while the bigger top half-disk is sent to the small one around $s_{2}$ and the lower bigger half disk is sent to the small one around $s_{1}$. At the right hand side of the figure, we show how to move a small neighborhood of a point in the stable line of $p_{u}$ so to create a tangency with the unstable manifold of $p_{s}$. This is done through a one-parameter family of diffeomorphisms;
until we create such a tangency the corresponding map remains hyperbolic, i.e. having a hyperbolic limit set with no cycles among its basic sets.

Concerning Smale's conjecture on the denseness of hyperbolicity, it is to be remarked that the question is still open for rational maps of the Riemann sphere. On the other hand, it was known to be true in the early 60's for flows on orientable surfaces (Peixoto), preceeded by the case of the two-disk (Andronov-Pontryagin). Much more recently, it was proved to be true for quadratic maps of the interval (Graczyk-Swiatek and Lyubich) and then, more generally, unimodal maps (Kozlovski) and, finally, multimodal maps (Kozlovski-Shen-Van Strien).

We point out that comprehensive accounts of the results mentioned above and many other related ones are in $[\mathrm{BDV}]$, $[\mathrm{P} 1]$ and $[\mathrm{P} 2]$.

The present paper represents a contribution to the understanding of the dynamics arising from bifurcating a cycle of a $C^{\infty}$ surface diffeomorphism. The notion of "most" used here requires us to consider one parameter families of diffeomorphisms $g_{t}$ strictly containing the initial bifurcating one, say $g_{0}$ at parameter value $t=0$. We assume that $g_{t}$ is hyperbolic for $t<0$ and $|t|$ small. We suppose that the cycle is formed by a (hyperbolic) horseshoe $K$ and an orbit of tangency $o(q)$ between stable and unstable manifolds of different periodic orbits of $K$. We assume the maximal invariant set in a small neighborhood of $K \cup o(q)$ to consist precisely of $K \cup o(q)$. A main novelty is that we allow the Hausdorff dimension of $K$ to be larger than one, but not to far from it. We show that right after the bifurcation, i.e. for $t>0$ small, most diffeomorphisms still display no attractors nor repellors in some neighborhood of $K \cup o(q)$. This means that the parameter values corresponding to diffeomorphisms displaying no attractors nor repellors should have total density at $t=0$. The concept is again discussed in the next subsection.

Our results considerably extend those in $[\mathrm{PT}],[\mathrm{NP}]$ obtained for the case when the Hausdorff dimension $H D(K)$ is smaller than one. They were announced in [PY3].

Of course, we expect the same to be true for all cases $0<H D(K)<2$. For that, it seems to us that our methods need to be considerably sharpened: we have to study deeper the dynamical recurrence of points near tangencies of higher order (cubic, quartic ...) between stable and unstable curves. We also hope that the ideas introduced in the present paper might be useful in broader contexts. In the horizon lies the famous question whether for the standard family of area preserving maps, one can find sets of positive Lebesgue probability in parameter space such that: the corresponding maps display non-zero Lyapunov exponents in sets of positive Lebesgue probability in phase space. Finally, we expect our results to be true in higher dimensions (see [MPV]).

We wish to thank W. de Melo e M. Viana for fruitful conversations.

### 1.2 The Setting and a First Formulation of the Main Result

Let $f$ be a smooth, i.e. $C^{\infty}$ diffeomorphism of a smooth surface $M$.

Recall that a basic set is a compact hyperbolic transitive locally maximal invariant set. A basic set is a horseshoe if it is infinite and is neither an attractor nor a repellor.

We assume that there exists a basic set $K$ for $f$, points $p_{s}, p_{u} \in K, q \in M-K$ such that the following properties hold:
$\left(\mathbf{H 1 )} p_{s}\right.$ and $p_{u}$ are periodic points and belong to distinct periodic orbits;
(H2) $W^{s}\left(p_{s}\right)$ and $W^{u}\left(p_{u}\right)$ have a quadratic tangency at $q$;
(H3) there exists a neighbourhood $U$ of $K$, a neighbourhood $V$ of the orbit $\mathcal{O}(q)$ of $q$, such that $K \cup \mathcal{O}(q)$ is the maximal invariant set in $U \cup V$.

We would like to understand, when $U, V$ are appropriately small and $g$ is $C^{\infty}$ close to $f$, the maximal invariant set

$$
\begin{equation*}
\Lambda_{g}=\bigcap_{\mathbb{Z}} g^{-n}(U \cup V) \tag{1.1}
\end{equation*}
$$

Observe that the smaller set

$$
\begin{equation*}
K_{g}=\bigcap_{\mathbb{Z}} g^{-n}(U) \tag{1.2}
\end{equation*}
$$

is a horseshoe which is the hyperbolic continuation of $K$.
Let $\mathcal{U}$ be an appropriately small neighbourhood of $f$ in $\operatorname{Diff}^{\infty}(M)$. We still denote by $p_{s}, p_{u}$ the continuation of these hyperbolic periodic points in $\mathcal{U}$. The condition that $W^{s}\left(p_{s}\right), W^{u}\left(p_{u}\right)$ have a quadratic tangency near $q$ defines a codimension 1 hypersurface $\mathcal{U}_{0}$ through $f$ in $\mathcal{U}$. It divides $\mathcal{U}$ into regions $\mathcal{U}_{+}, \mathcal{U}_{-}$such that, for $g \in \mathcal{U}_{-}, W^{s}\left(p_{s}\right)$ and $W^{u}\left(p_{u}\right)$ do not intersect near $q$ while, for $g \in \mathcal{U}_{+}$, $W^{s}\left(p_{s}\right)$ and $W^{u}\left(p_{u}\right)$ have two transverse intersection points near $q$ (for obvious dynamical reasons, the intersection is actually infinite in this case; we are really considering here the intersection derived from the continuation of large compact curves contained in $W^{s}\left(p_{s}\right)$ and $W^{u}\left(p_{u}\right)$.

When $g \in \mathcal{U}_{-}$, we clearly have

$$
\begin{equation*}
\Lambda_{g}=K_{g} \tag{1.3}
\end{equation*}
$$

When $g \in \mathcal{U}_{0}$, we have

$$
\begin{equation*}
\Lambda_{g}=K_{g} \cup \mathcal{O}\left(q_{g}\right) \tag{1.4}
\end{equation*}
$$

where $q_{g}$ is the tangency point close to $q$ given by the definition of $\mathcal{U}_{0}$.

The interesting case is therefore $g \in \mathcal{U}_{+}$.

It is actually not realistic to try to understand $\Lambda_{g}$ for all $g \in \mathcal{U}_{+}$. One of the reasons is the so-called Newhouse's phenomenon $[\mathrm{N}]$ : there exists an open set $\mathcal{N} \subset \mathcal{U}_{+}$, with $\mathcal{U}_{0} \subset \overline{\mathcal{N}}$, such that, residually in $\mathcal{N}, \Lambda_{g}$ has infinitely many periodic sinks or sources and so its full dynamical description appears to be beyond reach. See also [BC], [MV], [C] for similar results involving Hénon-like attractors.

Still, we can and shall consider most $g \in \mathcal{U}_{+}$in the following sense.

We will say that a subset $\mathcal{P} \subset \mathcal{U}_{+}$contains most $g \in \mathcal{U}_{+}$if, for any smooth 1-parameter family $\left(g_{t}\right)_{t \in\left(-t_{0}, t_{0}\right)}$ which is transverse to $\mathcal{U}_{0}$ at $t=0$ (with $g_{t} \in \mathcal{U}_{+}$for $t>0$ ), we have

$$
\begin{equation*}
\lim _{t \rightarrow 0} \frac{1}{t} \operatorname{Leb}\left(s \in(0, t], g_{s} \in \mathcal{P}\right)=1 \tag{1.5}
\end{equation*}
$$

Denote by $W^{s}(K)$ (resp. $\left.W^{u}(K)\right)$ the stable set (resp. unstable set) of $K$ for $f$. This is a partial foliation with a $C^{1+\alpha}$ Cantor transverse structure; denote by $d_{s}^{0}$ (resp. $d_{u}^{0}$ ) the transverse Hausdorff dimension of $W^{s}(K)$ (resp. $W^{u}(K)$ ). The Hausdorff dimension of $K$ is equal to $d_{s}^{0}+d_{u}^{0}$. We then have, in some contrast to Newhouse's phenomenon:

Theorem. [PT], [NP] Assume that $d_{s}^{0}+d_{u}^{0}<1$. Then, for most $g \in \mathcal{U}_{+}, \Lambda_{g}$ is a horseshoe.
On the other hand, by [PY1], the same conclusion does not hold when $d_{s}^{0}+d_{u}^{0}>1$. The paper [MY] gives substantially more geometric information in this case, specially concerning tangencies between stable and unstable manifolds (lines) in the hyperbolic continuation $K_{g}$ of $K$. These results have been extended to higher dimensions, as announced in [MPV] and complete proofs to appear in the near future.

In the present work, we investigate the maximal invariant set $\Lambda_{g}$, for most $g \in \mathcal{U}_{+}$, provided that the dimensions $d_{s}^{0}, d_{u}^{0}$ satisfy (see figure 2)

$$
\begin{equation*}
\left(d_{s}^{0}+d_{u}^{0}\right)^{2}+\left(\max \left(d_{s}^{0}, d_{u}^{0}\right)\right)^{2}<d_{s}^{0}+d_{u}^{0}+\max d_{s}^{0}, d_{u}^{0} \tag{H4}
\end{equation*}
$$

Our results can essentially be summarized as:

Main Theorem. Assume that (H1), (H2), (H3), (H4) hold. Then, for most $g \in \mathcal{U}_{+}, \Lambda_{g}$ is a non-uniformly hyperbolic horseshoe.

The meaning of a non-uniformly hyperbolic horseshoe in the present context will be explained somewhat in the next section and more completely in the rest of the paper. We can, however, comment that, for most $g \in \mathcal{U}_{+}, \Lambda_{g}$ will be a saddle-like object in the sense that both the stable set $W^{s}\left(\Lambda_{g}\right)$ and the unstable set $W^{u}\left(\Lambda_{g}\right)$ have Lebesgue measure zero and, so, it carries no attractors nor repellors. It will be (non-uniformly) hyperbolic in the sense that we will construct geometric
invariant measures, à la Sinai-Ruelle-Bowen $[\mathrm{Si}, \mathrm{Ru}, \mathrm{BR}]$, on $\Lambda_{g} \subset W^{s}\left(\Lambda_{g}\right)$ and $\Lambda_{g} \subset W^{u}\left(\Lambda_{g}\right)$ with non-zero Lyapunov exponents. Such properties of the invariant set $\Lambda_{g}$ are made especially precise in Section 10 and 11, the last ones in the paper. They yield some rephrasing of the main result in these terms, which is presented at the end of Section 11.


Figure 2
Remark. In the case when $d_{s}^{0}+d_{u}^{0}<1$, mentioned above and studied in [PT], [NP], it is not necessary to assume that $p_{s}, p_{u}$ belong to distinct periodic orbits. It is probably not necessary in our case either, at least as far as the qualitative statements are concerned. But, this assumption seems to make the technicalities significantly easier in what is already a very long construction.

### 1.3 A Summary of the Next Sections of the Paper

Sections 2-4 consist mainly of preparatory work.
In Section 2, we introduce a Markov partition by smooth disjoint rectangles $\left(R_{a}\right)_{a \in a}$ for the horseshoe $K$. The dynamics in the neighbourhood $U$ of $K$ is given by the transition maps from one rectangle to another, which enjoy a nice hyperbolic behaviour. To understand the dynamics in the larger set $U \cup V$, we need to control the dynamics along a finite part of the orbit of $q$, stretching from the moment this orbit goes out of $R:=\sqcup R_{a}$ until it comes back to $R$. The region of exit of $R$ and the region of entry into $R$ are two parabolic tongues $L_{u}$ and $L_{s}$, and the transition map

$$
G=g^{N_{0}}: L_{u} \rightarrow L_{s}
$$

is a folding map which share many features with the Henon quadratic polynomial diffeomorphisms of the plane.

Section 3 is essentially a summary of our previous work [PY2], (which was written having the present paper in mind). The important concept of affine-like map is introduced. The basic idea, which goes back to the early stages of the hyperbolic theory, is to describe maps that present hyperbolic features in an implicit way exhibiting preference for coordinates with a macroscopic range. Concretely, if a two-dimensional diffeomorphism contracts the vertical coordinate $y$ and expands the horizontal coordinate $x$, we use $y_{0}$ and $x_{1}$ as independent variables associated with a point $\left(x_{0}, y_{0}\right)$ and its image $\left(x_{1}, y_{1}\right)$, writing $x_{0}$ and $y_{1}$ as functions of $y_{0}$ and $x_{1}$.

Cone conditions are easy to formulate in this setting. A nice feature of this implicit representation of the dynamics is that it is time-symmetric: the map and its inverse satisfy symmetric formulas. Another even more important feature is that this formalism is well-adapted to the right concept of distortion (for 2-dimensional maps), yielding appropriate control on the partial derivatives of order two.

Composition of two affine-like maps which satisfy the same cone condition is also affine-like, and the distortion is only slightly bigger than the distortion of the two maps. Besides this "simple" composition, we study "parabolic" compositions of the form $F_{1} \circ G \circ F_{0}$, where $F_{0}, F_{1}$ are affine-like and $G$ is the folding map of Section 2. When the relative positions of the parabolic strip $G\left(Q_{0}\right)$ (where $Q_{0}$ is the image of $F_{0}$ ) and $P_{1}$ (the domain of $F_{1}$ ) are appropriate, the domain of $F_{1} \circ G \circ F_{0}$ has two connected components and the restrictions $F^{ \pm}$of $F_{1} \circ G \circ F_{0}$ to each component is affine-like. A control of the distortion of $F^{+}$and $F^{-}$is also obtained.

In Section 4, the general structure of parameter space is introduced. The parameter coordinate is normalized by the relative speed at the quadratic tangency of the tips of the stable and unstable manifolds. Then, with $\varepsilon_{0}$ very small, the starting interval $I_{0}:=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]$ for the parameter selection process is introduced. A small parameter $\tau$ (with $\tau \ll 1$ but still $\varepsilon_{0}^{\tau} \ll 1$ ) determines a sequence of scales $\left(\varepsilon_{k}\right)_{k \geq 0}$ in parameter space through the formula $\varepsilon_{k+1}=\varepsilon_{k}^{1+\tau}$. At level $k$, we have disjoint parameter intervals of length $\varepsilon_{k}$ (starting from level 0 with $I_{0}$ ). Each parameter interval of level $k$ that has been selected is divided into $\left[\varepsilon_{k}^{-\tau}\right]$ disjoint candidates of length $\varepsilon_{k+1}$. These candidates will pass a test to decide whether they are selected at level $k+1$.

The test takes two forms. First, in Section 5, a property of the parameter interval called regularity (see below) will be introduced; candidates which do not possess this property are discarded. Such a property is sufficient to develop in Sections $5-8$ some basic combinatorial and quantitative properties, but it is not well-adapted to an inductive scheme. Hence, in Section 9, a stronger property called strong regularity is introduced, and candidates have to satisfy this property in order to be selected.

Sections 5-7 constitute in some sense a single logical step: in Section 5, certain classes of restrictions of iterates of $g_{t}$ are inductively defined, and the definition is only possible because of properties that are inductively proved in Sections 6 and 7.

In Section 5, the goal is to define, for each parameter interval $I$ which is a candidate (i.e. its parent interval at the immediately upper level has been tested as regular), a class $R(I)$ of $I$-persistent affine-like iterates. An $I$-persistent affine-like iterate is a triple $(P, Q, n)$ where $P$ is a vertical-like strip in some rectangle $R_{a}$ depending on $t \in I, Q$ is a horizontal-like strip in some rectangle $R_{a^{\prime}}$, depending on $t \in I$, and the restriction of $g_{t}^{n}$ to $P$ is a diffeomorphism onto $Q$ which is affine-like.

However, we do not want to have in $\mathcal{R}(I)$ all $I$-persistent affine-like iterates: we will argue about them by induction (on $n$, for instance) and in order to do this, we want to obtain them in some explicit constructive way. Therefore, a number of Axioms, (R1)-(R7), are introduced and together they completely determine the class $\mathcal{R}(I)$. The most important feature of these Axioms is the following: every element of $\mathcal{R}(I)$ consisting of more than one iteration of $g_{t}$ can be obtained from simpler elements of $\mathcal{R}(I)$ by simple or parabolic composition; in this context, the notions of parent and simple or non-simple child introduced here, play a relevant role; simple composition is allowed in $\mathcal{R}(I)$ whenever it makes sense; and parabolic compositions of elements of $\mathcal{R}(I)$ is allowed if and only if a certain transversality relation is satisfied.

Thus, the definition of $\mathcal{R}(I)$ is reduced to the definition of this transversality relation, which is presented in Subsection 5.4. The intuitive notion behind the formal definition is the following: an element $(P, Q, n)$ should be $I$-transverse to an element $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ if the distance $\delta\left(Q, P^{\prime}\right)$ between the tip of the parabolic-like strip $G(Q)$ and $P^{\prime}$ satisfies

$$
\delta\left(Q, P^{\prime}\right) \geq|Q|^{1-\eta}+\left|P^{\prime}\right|^{1-\eta}
$$

for all $t \in I$, where $|Q|,\left|P^{\prime}\right|$ are the widths of the strips $Q$ and $P^{\prime}$. Here $\eta$ is a small positive constant, fixed once and for all. However, a number of properties, presented in Section 6, are very helpful, and they require a definition of the transversality relation that may seem quite complicated. In Appendix C, we explain why this seemed complication is rather necessary.

For the starting interval $I_{0}$, it follows from the formal definition that the transversality relation is never satisfied; therefore, parabolic composition is not allowed and the class $\mathcal{R}\left(I_{0}\right)$ is exactly the one associated with the symbolic dynamics given by the Markov partition. We conclude Section 5 with the definition of regularity. First, one says that a strip $P$ (from an element $(P, Q, n) \in \mathcal{R}(I)$ ) is $I$-transverse if one can find finitely many $Q_{\alpha}$, whose union contains the unstable set of $\Lambda$, such that $Q_{\alpha}$ and $P$ are $I$-transverse for every $\alpha$; otherwise one says that $P$ is $I$-critical. Then, given a constant $\beta>1$, one says that the parameter interval is $\beta$-regular if any $(P, Q, n) \in \mathcal{R}(I)$ such that both $P$ and $Q$ are $I$-critical satisfies $|P|<|I|^{\beta},|Q|<|I|^{\beta}$ for all $t \in I$. Intuitively, this means that no short return to the critical set is allowed.

In Section 6 , we prove a number of properties of the transversality relation and the classes $\mathcal{R}(I)$.

Amongst the most important is the following: children are born from their parents. Let us explain what it means. Let $(P, Q, n) \in \mathcal{R}(I)$, and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ the element of $\mathcal{R}(I)$ such that $P \subset \widetilde{P}$, $P \neq \widetilde{P}$ and $\widetilde{P}$ is the thinnest rectangle with this property; one says that $P$ is a child of $\widetilde{P}$ and that $\widetilde{P}$ is the parent of $P$. There are two cases; either $n=\widetilde{n}+1$ and one says that $P$ is a simple child; $(P, Q, n)$ is obtained by simple composition of $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ with an element of length 1 ; or $n>\widetilde{n}+1$ and one says that $P$ is a non-simple child; one then proves that $(P, Q, n)$ is obtained by parabolic composition of $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ with some element $\left(P_{1}, Q_{1}, n_{1}\right)$.

The most important result in Section 6 is a structure theorem for new rectangles in Subsection 6.7. One considers an element $(P, Q, n)$ which belongs to $\mathcal{R}(I)$ but not to $\mathcal{R}(\widetilde{I})$, where $\widetilde{I}$ is the parameter interval containing $I$ of the level immediately inferior (one says that $\widetilde{I}$ is the parent of $I$ ). Then there is a unique way to write $(P, Q, n)$ as the result of a sequence of $k$ parabolic compositions, possible in $\mathcal{R}(I)$ but not in $\mathcal{R}(\widetilde{I})$, of elements $\left(P_{0}, Q_{0}, n_{0}\right), \cdots,\left(P_{k}, Q_{k}, n_{k}\right)$. This fundamental result has several useful corollaries.

We have grouped in Section 7 a number of calculations and estimates related to the definitions of Section 5, and that are necessary for the inductive construction of the classes $\mathcal{R}(I)$. The first result relates the length $n$ of an element $(P, Q, n)$ to the width $|P|,|Q|$ of the strips. While there is no uniform exponential estimate as in the uniformly hyperbolic case, we are still able to prove a stretched-exponential uniform estimate.

We prove next that a uniform cone condition is satisfied by the affine-like iterates that we consider, and also that they have uniformly bounded distortion. After a technical estimate related to parabolic composition, we deal in Subsection 7.6 with the relative speed of the strips when the parameter varies; this is clearly of capital importance if we are to succeed. A point which is worth mentioning is that we are not able to obtain estimates for all pairs of strips (actually, it is easy to see that such estimates do not exist); we have to restrict ourselves to strips satisfying certain criticality conditions, that, fortunately, will be satisfied every time we need some information on these speeds. In the last Subsection 7.7, we investigate the oscillation of the widths of the strips with the parameter. While it is just not true that the relative oscillation is bounded (in the sense that the maximum over a parameter interval is no greater than a constant times the minimum), the result that we get will allow us to argue as if it was.

At the end of Section 7, the construction of the classes $\mathcal{R}(I)$ is complete, for every parameter interval $I$ whose parent $\widetilde{I}$ is regular. But we still don't know whether a single interval $I$ is regular.

In Section 8, we develop several quantitative estimates that will turn out to be crucial both in the parameter selection process of Section 9 and in the analysis of the dynamics for strongly regular parameters in Sections 10 and 11. We first investigate, for a given element $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$, the number of elements $(P, Q, n)$ such that $P$ is a non-simple child of $\widetilde{P}$; we show that, for every $\varepsilon>0$, there are at most $\varepsilon^{-C \eta}$ such non-simple children with width $|P|$ larger than $\varepsilon|\widetilde{P}|$. The constant $\eta$ here is small
and related to the definition of the transversality relation. The meaning of this estimate is that the presence of non-simple children is not too significant from the point of view of Hausdorff (or box) dimension, as it is made clear in Subsection 8.2. In Subsection 8.3, we transfer this information to parameter space, combining it with the result on relative speed of strips in Subsection 7.6.

Section 9 is the longest one in the paper and deals with the parameter selection process. The concept of regularity is very useful to develop a number of properties of the classes $R(I)$, as we did in Sections 5-8. Unfortunately, we are not able to prove (and it is probably false) that, given a $\beta$-regular interval $\widetilde{I}$, most candidates $I \subset \widetilde{I}$ at the next level are $\beta$-regular. [It is a consequence of the structure theorem of Section 6 that all candidates are $\bar{\beta}$-regular, where $\bar{\beta}=\beta(1+\tau)^{-1}$ is very close to $\beta$; this allow us to obtain all qualitative consequences of regularity for all candidates; but obviously we cannot repeat this at many successive levels of parameter intervals, because we need to keep $\beta>1$.] The problem with the concept of regularity is that it is dealing with only one scale $|\widetilde{I}|^{\beta}$; it could happen a priori that for a regular parameter interval $\widetilde{I}$ we have many $\widetilde{I}$-bicritical $(P, Q, n) \in \mathcal{R}(I)$ with $|P|$ or $|Q|$ only slightly below the threshold $|\widetilde{I}|^{\beta}$ (and therefore above the next threshold $|I|^{\beta}$ for candidates $\left.I \subset \widetilde{I}\right)$; for each such $(P, Q, n)$, we have to eliminate candidates $I$ such that $(P, Q, n)$ is $I$-bicritical, and no candidate will survive this selection process if there are too many $(P, Q, n)$.

The solution to this difficulty is to introduce the condition of strong regularity, which implies regularity and gives a quantitative control at all scales. Actually, the strong regularity condition involves two parts. In the first, one controls the size of the critical locus (in several slightly different ways) by a series of eight inequalities which all amount to say that the "dimension" of the critical locus is not much larger than $d_{s}^{0}+d_{u}^{0}-1$. In this case, the parameter selection process is based on the result mentioned above in the last part of Section 8. The second part of the strong regularity condition, by far the most subtle one, is a quantitative estimate for the number of bicritical elements at all scales. Because of the inductive nature of the argument, which relies in an essential way on the structure theorem of Section 6, we need to control the number of elements $(P, Q, n) \in \mathcal{R}(I)$ such that $P$ is $I_{\alpha}$-critical, $Q$ is $I_{\omega}$-critical and $|P| \geqslant x$ for some $t \in I$. Here, $I_{\alpha}$ and $I_{\omega}$ are parameter intervals containing $I$, and the control will depend on $I_{\alpha}, I_{\omega}$ and $x$. The formulas in Subsection 9.4 present a phase transition with respect to the width parameter $x$. Discussing this phase transition leads naturally to the hypothesis (H4) on the transverse dimensions $d_{s}^{0}, d_{u}^{0}$ : a small calculation shows that (H4) is exactly what one needs to obtain $\beta$-regularity with $\beta>1$.

Having stated the strong regularity condition, the goal in the rest of Section 9 is to prove that, given a strongly regular parameter interval $\widetilde{I}$, most candidates $I \subset \widetilde{I}$ at the next level are also strongly regular (the proportion of failed candidates turns out to be not larger than $C|I|^{\tau^{2}}$ ). This requires the control of two things. First, is to bound the number of "new" bicritical elements $(P, Q, n) \in \mathcal{R}(I)$ which did not belong to $\mathcal{R}(\widetilde{I})$; this is based on the structure theorem of Section 6 and leads to a long but straightforward calculation. Second, is to estimate which proportion of bicritical elements for $\widetilde{I}$ are still bicritical for $I$; this is only necessary when $I_{\alpha}$ or $I_{\omega}$ is equal to
$I$; when only one of the two intervals $I_{\alpha}, I_{\omega}$ is equal to $I$, the idea is simply to estimate what is the mean proportion (over all candidates), and to discard candidates for which the proportion is much above the mean. To compute the mean proportion, we rely again on the result at the end of Section 8. The case where $I=I_{\alpha}=I_{\omega}$ is the most important and the most difficult. When $x$ is "large", the same argument than when $I=I_{\alpha} \neq I_{\omega}$ still applies; but when $x$ is "small", the phase transition of the estimate means than the argument is not sufficient any more. A more complicated strategy is required, which is explained in Subsection 9.8 and carried out in 9.9 through 9.13.

It is worth mentioning that up to the end of Section 9, we never consider the dynamics for a single parameter, only for parameter intervals. In the last two sections, we study the dynamics for a strongly regular parameter value, i.e. the intersection of a decreasing sequence $\left(I_{m}\right)$ of strongly regular parameter intervals.

In Section 10, we study the dynamics on the set of stable curves. A stable curve $\omega$ is the decreasing intersection of a sequence of vertical-like strips $P_{k}$, where $\left(P_{k}, Q_{k}, n_{k}\right) \in \mathcal{R}=\bigcup_{m} \mathcal{R}\left(I_{m}\right)$. The set of stable curves is denoted by $\mathcal{R}_{+}^{\infty}$, their union by $\widetilde{\mathcal{R}}_{+}^{\infty}$. In order to define a map on $\widetilde{\mathcal{R}}_{+}^{\infty}$ (which is not invariant under $g$ ), we introduce the concept of prime element in $R$, i.e. one which cannot be written as the simple composition of shorter elements. Let then $\omega$ be a stable curve which is not contained in infinitely many prime elements $P_{k}$, and let $(P, Q, n)$ be such that $P$ is the thinnest prime element containing $\omega$. The image $g^{n}(\omega)$ is contained in a stable curve $\omega^{\prime}$ and we set $T^{+}(\omega)=\omega^{\prime}, \widetilde{T}^{+} / \omega=g^{n} / \omega$. This defines a map $T^{+}$from a subset $\mathcal{D}_{+}$of $\mathcal{R}_{+}^{\infty}$ onto $\mathcal{R}_{+}^{\infty}$ which lifts to a map $\widetilde{T}^{+}$from the union $\widetilde{\mathcal{D}}_{+}$of curves in $\mathcal{D}_{+}$to $\widetilde{\mathcal{R}}_{+}^{\infty}$.

The map $T^{+}$is Bernoulli in the following sense: its domain $\mathcal{D}_{+}$splits into countably many pieces $\mathcal{R}_{+}^{\infty}(P)$ indexed by prime elements, and each piece is sent homeomorphically by $T^{+}$onto the intersection of $\mathcal{R}_{+}^{\infty}$ with some rectangle $R_{a}$ of the Markov partition.

The map $T^{+}$is uniformly expanding (with countably many branches) and we introduce a one parameter family of weighted transfer operators in the spirit of classical uniformly hyperbolic maps. One has only to be careful because the presence of countably many branches is the source of some problems, which are dealt with in Subsection 10.3 using the estimates of Section 8 on the number of children.

As expected, the transfer operators $L_{d}$, considered in the appropriate function space, turn out to have a positive eigenfunction $h_{d}$ associated with a dominant eingevalue $\lambda_{d}>0$. There is a unique value $d_{s}$ such that $\lambda_{d_{s}}=1$. This value turns out to be, unsurprisingly, the transverse Hausdorff dimension of the partial foliation $\widetilde{\mathcal{R}}_{+}^{\infty}$ (which is proved in Subsection 10.5 to be transversally Lipschitzian). The transfer operator also allows us to identify, as usual, a measure $\mu_{d}$ with prescribed Jacobian and an invariant measure $\nu_{d}=h_{d} \mu_{d}$. For $d=d_{s}$, the $\mu_{d}$-measure (or $\nu_{d}$-measure) of the set of stable curves contained in any vertical-like strip $P$ is proportional to $|P|^{d s}$.

The set $\widetilde{\mathcal{R}}_{+}^{\infty}-\widetilde{\mathcal{D}}_{+}$where $\widetilde{T}^{+}$is not defined, has transverse dimension smaller than $d_{s}$, hence is negligible in a geometrical sense. One can lift the $T^{+}$-invariant measure $\nu=\nu_{d_{s}}$ to a $\widetilde{T}^{+}$-invariant measure $\widetilde{\nu}$ which is ergodic and then spread it to a $g$-invariant measure on $\Lambda$.

In Section 11, the last in the paper, we pursue the study of the dynamics of $g_{t}$ on $\Lambda=\Lambda_{g_{t}}$ for a strongly regular parameter $t$, looking now beyond the well-behaved set $\widetilde{\mathcal{R}}_{+}^{\infty}$ which was studied in Section 10. In the first part (Subsections 11.1-11.5), we study the intersection of the invariant set $\Lambda$ with an unstable curve $\omega^{*}$ (defined as a stable curve, exchanging $P$ 's and $Q$ 's). The main part of this intersection is a countable disjoint union of dynamical copies of the set $\mathcal{R}_{+}^{\infty}$ studied in Section 10. There are also at most countably many critical points, corresponding to quadratic tangencies between stable curves and images under $G$ of unstable curves. And, finally, there is an exceptional set (formed by points which come very close to the critical locus infinitely many times); but this exceptional set is small; its Hausdorff dimension is explicitly controlled by a value much smaller than the dimension $d_{s}$ of $\omega^{*} \cap \Lambda$.

In the second part of Section 11, we prove that the invariant set $\Lambda$ is a saddle-like object in the metric sense: both its stable set $W^{s}(\Lambda)$ and its unstable set $W^{u}(\Lambda)$ have Lebesgue measure 0 . So, no attractors nor repellors are present on $\Lambda$. One actually expects more: certainly the Hausdorff dimension of $W^{s}(\Lambda)$ should be strictly less than 2 , probably it is close to $1+d_{s}$, and perhaps even equal to $1+d_{s}$. However, we stick to the simpler, but still very meaningful result: it implies that $\Lambda_{g}$ carries no attractor nor a repellor for most $g$.

One has a nice combinatorial decomposition of the restricted stable set $W^{s}(\Lambda, R)$, but to compute Lebesgue measure (or Hausdorff dimension), one has to transport the pieces of this decomposition by affine-like iterates of $g$ of high order. This is easy to do as far as Lebesgue measure is concerned, because bounded distortion of affine-like maps mean also bounded distortion of measure (bounded relative oscillation of Jacobians). This is much more delicate with respect to Hausdorff dimension: the geometry of the pieces after iteration can get very distorted.

In Appendix A, we recall all formulas related to the implicit representation of affine-like maps; many of them can already be found in [PY2], but we have also to consider the derivatives with respect to parameter, a setting which was not considered in [PY2].

In Appendix B, we perform some calculations related to proposition 40 in Subsection 10.5, which generates the transversally Lipschitz regularity of the partial foliation $\widetilde{\mathcal{R}}_{+}^{\infty}$.

In Appendix C, we give some justification for what seems to be a convoluted definition of the transversality relation in Subsection 5.4.

## 2 Markov Partition and Folding Map

### 2.1 Markov Partition and Related Charts

We will choose once and for all a finite system of smooth charts

$$
I_{a}^{s} \times I_{a}^{u} \stackrel{\approx}{\longrightarrow} R_{a} \subset M, \quad a \in a
$$

indexed by a finite alphabet $a$. Each chart depends smoothly on $g \in \mathcal{U}$; the intervals $I_{a}^{s}, I_{a}^{u}$ are compact; the rectangles $R_{a}$ are disjoint.

Let $R=\bigcup_{a} R_{a}$. We choose the charts in order to have:
(MP1) for each $g \in \mathcal{U}, K_{g}$ is the maximal invariant set in int $R$; for each $g \in \mathcal{U}, a \in a$, one has

$$
\begin{gather*}
g\left(\partial I_{a}^{s} \times I_{a}^{u}\right) \cap R=\emptyset  \tag{2.1}\\
g^{-1}\left(I_{a}^{s} \times \partial I_{a}^{u}\right) \cap R=\emptyset \tag{2.2}
\end{gather*}
$$

(MP2) for each $g \in \mathcal{U}$, the family $\left(R_{a} \cap K_{g}\right)_{a \in}$ induces a Markov partition for the horseshoe $K_{g}$. Let

$$
\begin{equation*}
\mathcal{B}=\left\{\left(a, a^{\prime}\right) \in a^{2}, \quad f\left(R_{a}\right) \cap R_{a^{\prime}} \neq \emptyset\right\} \tag{2.3}
\end{equation*}
$$

The Markov partition provides a coding which is a topological conjugacy between the horseshoe $K_{g}$ and the subshift of finite type of $a^{\mathbb{Z}}$ defined by $\mathcal{B}$.

### 2.2 The Parabolic Tongues $L_{u}, L_{s}$

Denote by $a_{s}, a_{u} \in a$ the letters such that $p_{s} \in R_{a_{s}}, p_{u} \in R_{a_{u}}$. We choose the corresponding charts in order to have:
(MP3) for each $g \in \mathcal{U}$, the equation of the local stable manifold $W_{\text {loc }}^{s}\left(p_{s}\right)$ is $\left\{x_{a_{s}}=0\right\}$, the equation of the local unstable manifold $W_{\text {loc }}^{u}\left(p_{u}\right)$ is $\left\{y_{a_{u}}=0\right\}$.

We have written $x_{a}$ (resp. $y_{a}$ ) for the coordinate in $I_{a}^{s}$ (resp. $I_{a}^{u}$ ). We also choose the rectangles $R_{a}$ in order to have, for some integer $N_{0} \geqslant 2$ :
(MP4) for each $g \in \mathcal{U}_{0}$, there are points $q_{s}, q_{u}$ in the orbit of $q$ such that

- for $n \geq 0, g^{n}\left(q_{s}\right)$ and $g^{n}\left(p_{s}\right)$ belong to the interior of the same rectangle;
- for $n \leq 0, g^{n}\left(q_{u}\right)$ and $g^{n}\left(p_{u}\right)$ belong to the interior of the same rectangle;
- $q_{s}=g^{N_{0}}\left(q_{u}\right)$ and $g^{i}\left(q_{u}\right)$ does not belong to $R$ for $0<i<N_{0}$.

Consider small pieces of $W^{s}\left(p_{s}\right), W^{u}\left(p_{u}\right)$ which are tangent at $q_{u}$ for $g \in \mathcal{U}_{0}$. When $g \in \mathcal{U}_{+}$, these pieces will meet in two points and bound a compact lenticular region $L_{u} \subset$ int $R_{a_{u}}$. Taking the image under $g^{N_{0}}$, we get another lenticular region $L_{s} \subset$ int $R_{a_{s}}$. These regions are called parabolic tongues. See figure 3.


Figure 3

Define then, for $g \in \mathcal{U}_{+}$

$$
\begin{equation*}
\widehat{R}=R \bigcup_{0<i<N_{0}} g^{i}\left(L_{u}\right) \tag{2.4}
\end{equation*}
$$

The maximal invariant set we are interested in is

$$
\begin{equation*}
\Lambda_{g}=\bigcap_{\mathbb{Z}} g^{-n}(\widehat{R}) \tag{2.5}
\end{equation*}
$$

We also define

$$
\begin{align*}
W^{s}\left(\Lambda_{g}, \widehat{R}\right) & =\bigcap_{n \geqslant 0} g^{-n}(\widehat{R})  \tag{2.6}\\
W^{u}\left(\Lambda_{g}, \widehat{R}\right) & =\bigcap_{n \leqslant 0} g^{-n}(\widehat{R}) \tag{2.7}
\end{align*}
$$

The dynamics in $\widehat{R}$ are generated by

- the transition maps related to the Markov partition:

$$
g: R_{a} \cap g^{-1}\left(R_{a^{\prime}}\right) \rightarrow g\left(R_{a}\right) \cap R_{a^{\prime}}, \text { for }\left(a, a^{\prime}\right) \in \mathcal{B}
$$

- the folding map $G:=g^{N_{0}}$ from $L_{u}$ onto $L_{s}$.


### 2.3 The Folding Map $G$

For simplicity, we write $\left(x_{s}, y_{s}\right),\left(x_{u}, y_{u}\right)$ for the coordinates in $R_{a_{s}} \supset L_{s}, R_{a_{u}} \supset L_{u}$.

The folding map $G$ is parabolic in the sense of [PY2]; let us recall this definition.
Consider the graph $\Gamma_{G}$ of the restriction $G$ of $g^{N_{0}}$ to the component of $R_{a_{u}} \cap g^{-N_{0}}\left(R_{a_{s}}\right)$ which contains $L_{u}$ (for $g \in \mathcal{U}_{+}$; we then follow this component in the rest of $\mathcal{U}$ ). Using the corresponding charts, we can view $\Gamma_{G}$ as a surface in $I_{a_{u}}^{s} \times I_{a_{u}}^{u} \times I_{a_{s}}^{s} \times I_{a_{s}}^{u}$. Denote by $\pi$ the projection from $I_{a_{u}}^{s} \times I_{a_{u}}^{u} \times I_{a_{s}}^{s} \times I_{a_{s}}^{u}$ onto $I_{a_{u}}^{u} \times I_{a_{s}}^{s}$. For $\mathcal{U}$ small enough, from the quadratic tangency at $q$ and (MP3) we deduce that:
(P1) the restriction of $\pi$ to $\Gamma_{G}$ is a fold map (in the sense of singulary theory).

Denote by $\Gamma_{0} \subset I_{a_{u}}^{u} \times I_{a_{s}}^{s}$ the smooth curve which is the image of the critical locus of this fold map. It divides $I_{a_{u}}^{u} \times I_{a_{s}}^{s}$ into two regions $\Gamma_{+}, \Gamma_{-}$such that $\Gamma_{+} \cup \Gamma_{0}$ is the image of the fold map. We can reformulate (P1) as:
( $\left.\mathbf{P}^{\prime} \mathbf{1}\right)$ (i) for $\left(y^{0}, x^{0}\right) \in \Gamma_{0}$, the image $G\left(\left\{y_{u}=y^{0}\right\}\right)$ meets $\left\{x_{s}=x^{0}\right\}$ in a single point, interior to both curves, at which the curves have a quadratic tangency;
(ii) for $\left(y^{0}, x^{0}\right) \in \Gamma_{-}$, the curves $G\left(\left\{y_{u}=y^{0}\right\}\right)$ and $\left\{x_{s}=x^{0}\right\}$ do not intersect;
(iii) for $\left(y^{0}, x^{0}\right) \in \Gamma_{+}$, the curves $G\left(\left\{y_{u}=y_{0}\right\}\right)$ and $\left\{x_{s}=x^{0}\right\}$ intersect transversally in two points.

As $G$ is a diffeomorphism, the tangents to $\Gamma_{0}$ are never vertical or horizontal. Therefore, we can and will choose a smooth function $\theta$ on $I_{a_{u}}^{u} \times I_{a_{s}}^{s}$ such that
(P2) $\theta \equiv 0$ on $\Gamma_{0}, \theta>0$ on $\Gamma_{+}, \theta<0$ on $\Gamma_{-}$;
(P3) the partial derivatives $\theta_{y}, \theta_{x}$ of $\theta$ do not vanish on $I_{a_{u}}^{u} \times I_{a_{s}}^{s}$.
Remark. The choice of $\theta$ is far from unique. One could for instance choose $\theta$ of the form

$$
\begin{equation*}
\theta\left(y_{u}, x_{s}\right)=\varepsilon_{u} y_{u}+\varepsilon_{s} \chi\left(x_{s}\right), \tag{2.8}
\end{equation*}
$$

with $\varepsilon_{s}, \varepsilon_{u} \in\{-1,+1\}$ and $\chi$ monotone increasing. We prefer not to specify a particular choice in order to keep a time-symmetric setting between positive and negative iterations.

From $\theta$, we define a smooth function $w$ on $\Gamma_{G}$ by
$(\mathbf{P} 4) w^{2}=\theta \circ \pi$
(there are two choices for $w$; the other is $-w$ ).

Then, from (P3) we obtain smooth maps $Y_{u}, X_{s}$ implicitly defined by
(P5)

$$
\begin{aligned}
w^{2} & =\theta\left(Y_{u}\left(w, x_{s}\right), x_{s}\right) \\
& =\theta\left(y_{u}, X_{s}\left(w, y_{u}\right)\right.
\end{aligned}
$$

On the graph $\Gamma_{G}$, we can use either $\left(x_{u}, y_{u}\right)$ or $\left(x_{s}, y_{s}\right)$ or $\left(w, y_{u}\right)$ or $\left(x_{s}, w\right)$ as coordinates; therefore we can factorize $G$ as $G_{+} \circ G_{0} \circ G_{-}$:

$$
\begin{equation*}
\left(x_{u}, y_{u}\right) \xrightarrow{G_{-}}\left(w, y_{u}\right) \xrightarrow{G_{0}}\left(x_{s}, w\right) \xrightarrow{G_{+}}\left(x_{s}, y_{s}\right) \tag{2.9}
\end{equation*}
$$

with
(P6)

$$
\begin{aligned}
& G_{0}\left(w, y_{u}\right)=\left(X_{s}\left(w, y_{u}\right), w\right), \\
& G_{0}^{-1}\left(x_{s}, w\right)=\left(w, Y_{u}\left(w, x_{s}\right)\right), \\
& G_{+}\left(x_{s}, w\right)=\left(x_{s}, Y_{s}\left(w, x_{s}\right)\right), \\
& G_{-}^{-1}\left(w, y_{u}\right)=\left(X_{u}\left(w, y_{u}\right), y_{u}\right) .
\end{aligned}
$$

The last two formulas define smooth maps $Y_{s}, X_{u}$ and the partial derivatives $Y_{s, w}, X_{u, w}$ do not vanish as $G_{+}, G_{-}$are diffeomorphisms. Observe that the map $G_{0}$ is very similar to a quadratic Hénon-like map.

## 3 Affine-like Maps

This section is essentially a summary of [PY2].

### 3.1 Definition and Implicit Representation

Let $I_{0}^{s}, I_{0}^{u}, I_{1}^{s}, I_{1}^{u}$ be non trivial compact intervals, $x_{0}, y_{0}, x_{1}, y_{1}$ the corresponding coordinates. Consider a smooth diffeomorphisms $F$ whose domain is a vertical strip

$$
P=\left\{\varphi^{-}\left(y_{0}\right) \leqslant x_{0} \leqslant \varphi^{+}\left(y_{0}\right)\right\} \subset I_{0}^{s} \times I_{0}^{u}
$$

and whose image is a horizontal strip

$$
Q=\left\{\psi^{-}\left(x_{1}\right) \leqslant y_{1} \leqslant \psi^{+}\left(x_{1}\right)\right\} \subset I_{1}^{s} \times I_{1}^{u} .
$$

We say that $F$ is affine-like if
(AL1) the restriction to the graph of $F$ of the projection onto $I_{0}^{u} \times I_{1}^{s}$ is a diffeomorphism onto $I_{0}^{u} \times I_{1}^{s}$.

This allow us to define smooth maps $A, B$ on $I_{0}^{u} \times I_{1}^{s}$ such that

$$
F\left(x_{0}, y_{0}\right)=\left(x_{1}, y_{1}\right) \Longleftrightarrow\left\{\begin{array}{l}
x_{0}=A\left(y_{0}, x_{1}\right)  \tag{3.1}\\
y_{1}=B\left(y_{0}, x_{1}\right) .
\end{array}\right.
$$

The pair $(A, B)$ is the implicit representation (or definition) of the affine-like map $F$. See figure 4 . In the formulas below, we shall most of the time omit the arguments of the functions considered, which should be obvious from the context. We will write $A_{x}, A_{y}, A_{x x}, B_{x}, B_{y} \cdots$ for partial derivatives.


Figure 4

On the graph of $F$, we have

$$
\begin{align*}
d x_{0} & =A_{y} d y_{0}+A_{x} d x_{1},  \tag{3.2}\\
d y_{1} & =B_{y} d y_{0}+B_{x} d x_{1},
\end{align*}
$$

which leads to

$$
\begin{align*}
D F & =A_{x}^{-1}\left(\begin{array}{cc}
1 & -A_{y} \\
B_{x} & A_{x} B_{y}-A_{y} B_{x},
\end{array}\right)  \tag{3.3}\\
D F^{-1} & =B_{y}^{-1}\left(\begin{array}{cc}
A_{x} B_{y}-A_{y} B_{x} & A_{y} \\
-B_{x} & 1,
\end{array}\right)  \tag{3.4}\\
\operatorname{det} D F & =A_{x}^{-1} B_{y} . \tag{3.5}
\end{align*}
$$

The main advantage of the implicit representation is the symmetry between positive and negative iteration.

### 3.2 Cone Condition and Distortion

Let $\lambda, u, v>0$ satisfy

$$
\begin{equation*}
1<u v \leqslant \lambda^{2} . \tag{3.6}
\end{equation*}
$$

Let $\left(X_{0}, Y_{0}\right)$ be a tangent vector at some point in the domain of $F$, and let $\left(X_{1}, Y_{1}\right)$ be its image under $T F$. The usual cone condition with parameters $(\lambda, u, v)$ is:
(AL2) (i) if $\left|Y_{0}\right| \leqslant u\left|X_{0}\right|$, then $\left|Y_{1}\right| \leqslant v^{-1}\left|X_{1}\right|$ and $\left|X_{1}\right| \geqslant \lambda\left|X_{0}\right|$;
(ii) if $\left|X_{1}\right| \leqslant v\left|Y_{1}\right|$, then $\left|X_{0}\right| \leqslant u^{-1}\left|Y_{0}\right|$ and $\left|Y_{0}\right| \geqslant \lambda\left|Y_{1}\right|$.

This is readily seen to be equivalent to
( $\mathrm{AL}^{\prime} \mathbf{2}$ )

$$
\begin{aligned}
& \lambda\left|A_{x}\right|+u\left|A_{y}\right| \leqslant 1, \\
& \lambda\left|B_{y}\right|+v\left|B_{x}\right| \leqslant 1,
\end{aligned}
$$

everywhere on $I_{0}^{u} \times I_{1}^{s}$.
We will also need to control partial derivatives of second order of $A, B$. By (3.5), the partial derivatives $A_{x}, B_{y}$ do not vanish on $I_{0}^{u} \times I_{1}^{s}$. It turns out that the right way to look at partial derivatives of second order is to consider the six functions

$$
\begin{aligned}
& \partial_{x} \log \left|A_{x}\right|, \partial_{y} \log \left|A_{x}\right|, A_{y y}, \\
& \partial_{y} \log \left|B_{y}\right|, \partial_{x} \log \left|B_{y}\right|, B_{x x} .
\end{aligned}
$$

We define the distortion of an affine-like map $F$, and denote by $D(F)$, the maximal absolute value attained by any one of these six functions on $I_{0}^{u} \times I_{1}^{s}$.

We also define the width of the domain $P$ of $F$ by

$$
\begin{equation*}
|P|:=\max \left|A_{x}\right|, \tag{3.7}
\end{equation*}
$$

and the width of the image $Q$ by

$$
\begin{equation*}
|Q|:=\max \left|B_{y}\right| . \tag{3.8}
\end{equation*}
$$

### 3.3 Simple Composition

The composition of two affine-like maps is not always affine-like. However, the composition of two affine-like maps which also satisfy the same cone condition (AL2) will again be affine-like and satisfy the same cone condition (actually a better one).

More precisely, let $I_{0}^{s}, I_{0}^{u}, I_{1}^{s}, I_{1}^{u}, I_{2}^{s}, I_{2}^{u}$ be compact intervals. Let $F: P \rightarrow Q$ and $F^{\prime}: P^{\prime} \rightarrow Q^{\prime}$ be affine-like maps with domains $P \subset I_{0}^{s} \times I_{0}^{u}, P^{\prime} \subset I_{1}^{s} \times I_{1}^{u}$ and images $Q \subset I_{1}^{s} \times I_{1}^{u}, Q^{\prime} \subset I_{2}^{s} \times I_{2}^{u}$. We assume that both $F$ and $F^{\prime}$ satisfy (AL2) (or (AL'2)) with parameters $\lambda, u, v$. The composition $F^{\prime \prime}=F^{\prime} \circ F$ has domain $P^{\prime \prime}=P \cap F^{-1}\left(P^{\prime}\right)$ and image $Q^{\prime \prime}=Q^{\prime} \cap F^{\prime}(Q)$. It satisfies (AL1) and (AL2) with parameters $\lambda^{2}, u, v$ (cf. [PY2]). See figure 5.


Figure 5

Let $(A, B),\left(A^{\prime}, B^{\prime}\right),\left(A^{\prime \prime}, B^{\prime \prime}\right)$ be the implicit representations of $F, F^{\prime}, F^{\prime \prime}$ respectively. Define

$$
\begin{equation*}
\Delta:=1-A_{y}^{\prime} B_{x}>1-u^{-1} v^{-1}>0 . \tag{3.9}
\end{equation*}
$$

The partial derivatives of first order of $A^{\prime \prime}, B^{\prime \prime}$ are given by

$$
\begin{align*}
& A_{x}^{\prime \prime}=A_{x} A_{x}^{\prime} \Delta^{-1},  \tag{3.10}\\
& B_{y}^{\prime \prime}=B_{y} B_{y}^{\prime} \Delta^{-1},
\end{align*}
$$

$$
\begin{align*}
& A_{y}^{\prime \prime}=A_{y}+A_{y}^{\prime} A_{x} B_{y} \Delta^{-1},  \tag{3.11}\\
& B_{x}^{\prime \prime}=B_{x}^{\prime}+B_{x} A_{x}^{\prime} B_{y}^{\prime} \Delta^{-1} .
\end{align*}
$$

From (3.10), we get

$$
\begin{align*}
& C^{-1} \leqslant \frac{\left|P^{\prime \prime}\right|}{|P|\left|P^{\prime}\right|} \leqslant C,  \tag{3.12}\\
& C^{-1} \leqslant \frac{\left|Q^{\prime \prime}\right|}{|Q|\left|Q^{\prime}\right|} \leqslant C,
\end{align*}
$$

where the constants are uniform once $u, v$ are fixed and the distortions are uniformly bounded.
The formulas for the partial derivatives of second order are derived in [PY2] and recalled in Appendix A. They lead to the following estimate for the distortion:

$$
\begin{equation*}
D\left(F^{\prime \prime}\right) \leqslant \max \left\{D(F)+C|Q|\left(D(F)+D\left(F^{\prime}\right)\right), D\left(F^{\prime}\right)+C\left|P^{\prime}\right|\left(D(F)+D\left(F^{\prime}\right)\right)\right\}, \tag{3.13}
\end{equation*}
$$

where $C$ depends only on $u, v$.

### 3.4 Properties of the Markov Partition

We choose charts for the Markov partition discussed in Subsection 2.1 in order to have the following property, for some $\lambda, u, v$ satisfying (3.6):
(MP5) for any $\left(a, a^{\prime}\right) \in \mathcal{B}$, any $g \in \mathcal{U}$, the transition map $g_{a, a^{\prime}}$ from $P_{a a^{\prime}}=R_{a} \cap g^{-1}\left(R_{a^{\prime}}\right)$ onto $Q_{a a^{\prime}}=R_{a^{\prime}} \cap g\left(R_{a}\right)$ is affine-like and also satisfies the cone condition (AL2).

These values of ( $\lambda, u, v$ ) will be fixed in what follows.

To any finite word $\underline{a}=\left(a_{0}, \ldots, a_{n}\right)$ with transitions in $\mathcal{B}$, we have a composition

$$
g_{\underline{a}}=g_{a_{n-1} a_{n}} \circ \ldots \circ g_{a_{0} a_{1}}
$$

which satisfies also (AL1) and (AL2).
Moreover, as the widths decrease exponentially with the number of iterations, it follows from (3.13) that there exists $D_{0}>0$ such that all $g_{\underline{a}}$ satisfy
$(\mathrm{MP} 6) \quad D\left(g_{\underline{a}}\right) \leq D_{0}$.

### 3.5 Parabolic Composition

Let $G$ be the folding map of Subsection 2.3, satisfying properties (P1)-(P6).

Let also $I_{0}^{s}, I_{0}^{u}, I_{1}^{s}, I_{1}^{u}$ be compact intervals; let $F_{0}$ be an affine-like map from a vertical strip $P_{0} \subset I_{0}^{s} \times I_{0}^{u}$ to a horizontal strip $Q_{0} \subset I_{a_{u}}^{s} \times I_{a_{u}}^{u}$; let $F_{1}$ be an affine-like map from a vertical strip $P_{1} \subset I_{a_{s}}^{s} \times I_{a_{s}}^{u}$ to a horizontal strip $Q_{1} \subset I_{1}^{s} \times I_{1}^{u}$.

We recall from [PY2] how, under appropriate hypotheses, the composition $F_{1} \circ G \circ F_{0}$ defines two affine-like maps $F^{ \pm}$with domains $P^{ \pm} \subset P_{0}$ and image $Q^{ \pm} \subset Q_{1}$. See figure 6 .


Figure 6

Let $\left(A_{0}, B_{0}\right),\left(A_{1}, B_{1}\right)$ be implicit representations of $F_{0}, F_{1}$, respectively. We assume that

$$
\begin{array}{ll}
\left|A_{1, y}\right|<b, & \left|A_{1, y y}\right|<b  \tag{PC1}\\
\left|B_{0, x}\right|<b, & \left|B_{0, x x}\right|<b
\end{array}
$$

with $b \ll 1$. In the system

$$
\begin{align*}
x_{u} & =X_{u}\left(w, y_{u}\right)  \tag{3.14}\\
y_{u} & =B_{0}\left(y_{0}, x_{u}\right)
\end{align*}
$$

we can, as $\left|B_{0, x}\right| \ll 1$, eliminate $y_{u}$ and solve for $x_{u}$ to define

$$
\begin{equation*}
x_{u}=X\left(w, y_{0}\right) \tag{3.15}
\end{equation*}
$$

Similarly, in the system

$$
\begin{align*}
y_{s} & =Y_{s}\left(w, x_{s}\right),  \tag{3.16}\\
x_{s} & =A_{1}\left(y_{s}, x_{1}\right),
\end{align*}
$$

we eliminate $x_{s}$, solve for $y_{s}$ to get

$$
\begin{equation*}
y_{s}=Y\left(w, x_{1}\right) . \tag{3.17}
\end{equation*}
$$

The next step is to define

$$
\begin{equation*}
C\left(w, y_{0}, x_{1}\right):=w^{2}-\theta\left(B_{0}\left(y_{0}, X\left(w, y_{0}\right)\right), A_{1}\left(Y\left(w, x_{1}\right), x_{1}\right)\right) . \tag{3.18}
\end{equation*}
$$

This quantity has the following geometrical interpretation. Fix values $y_{0}^{*}, x_{1}^{*}$ for $y_{0}, x_{1}$. The image $G_{-} \circ F_{0}\left(\left\{y_{0}=y_{0}^{*}\right\}\right)$ is the graph

$$
\begin{equation*}
\gamma_{0}=\left\{y_{u}=B_{0}\left(y_{0}^{*}, X\left(w, y_{0}^{*}\right)\right)\right\} ; \tag{3.19}
\end{equation*}
$$

symmetrically, $G_{+}^{-1} \circ F_{1}^{-1}\left(\left\{x_{1}=x_{1}^{*}\right\}\right)$ is the graph

$$
\begin{equation*}
\gamma_{1}=\left\{x_{s}=A_{1}\left(Y\left(w, x_{1}^{*}\right), x_{1}^{*}\right)\right\} . \tag{3.20}
\end{equation*}
$$

Then, $C\left(w, y_{0}^{*}, x_{1}^{*}\right)$ gives the relative position of the two curves $\gamma_{0}$ and $G_{0}^{-1}\left(\gamma_{1}\right)$ (or equivalently $G_{0}\left(\gamma_{0}\right)$ and $\left.\gamma_{1}\right)$. More precisely, it is positive for all $w$ if the two curves do not intersect; it vanishes at the intersection points and is negative between the intersection points.

It follows from (PC1) just above that

$$
\begin{align*}
\mid C_{w}-2 w & \ll 1,  \tag{3.21}\\
\left|C_{w w}-2\right| & \ll 1 . \tag{3.22}
\end{align*}
$$

Therefore, for fixed values of $y_{0}$ and $x_{1}, C$ has a unique minimum as a function of $w$; we denote by $\bar{C}\left(y_{0}, x_{1}\right)$ the corresponding minimum value. We have $\bar{C}\left(y_{0}^{*}, x_{1}^{*}\right)>0$ (resp. $=0$, resp. $<0$ ) if and only if the curves $\gamma_{0}$ and $G_{0}^{-1}\left(\gamma_{1}\right)$ do not intersect (resp. are tangent, resp. have two transverse intersection points).

In order to consider parabolic compositions, we shall require that $\bar{C}\left(y_{0}, x_{1}\right)<0$ everywhere on $I_{0}^{u} \times I_{1}^{s}$. Setting

$$
\begin{equation*}
\delta=\min _{y_{0}, x_{1}}-\bar{C}\left(y_{0}, x_{1}\right) \tag{3.23}
\end{equation*}
$$

we actually want to have
(PC2)

$$
\delta>b^{-1}\left(\left|P_{1}\right|+\left|Q_{0}\right|\right) .
$$

The geometric interpretation of this requirement is clear: the displacement of one of the rectangles and the image of the other should be much bigger than the sum of their widths. In other words, the distance between the tip of the parabolic strip $G_{0}^{-1}\left(P_{1}\right)$ and the horizontal strip $Q_{0}$ should be much bigger than the widths of these strips.

Assume now that (PC1) and (PC2) are satisfied; the equation $C\left(w, y_{0}, x_{1}\right)=0$ defines two smooth functions

$$
\begin{equation*}
w=W^{ \pm}\left(y_{0}, x_{1}\right) \tag{3.24}
\end{equation*}
$$

with $w^{+}>w^{-}$. One then defines

$$
\begin{align*}
A^{ \pm}\left(y_{0}, x_{1}\right) & :=A_{0}\left(y_{0}, X\left(W^{ \pm}\left(y_{0}, x_{1}\right), y_{0}\right)\right)  \tag{3.25}\\
B^{ \pm}\left(y_{0}, x_{1}\right) & :=B_{1}\left(Y\left(W^{ \pm}\left(y_{0}, x_{1}\right), x_{1}\right), x_{1}\right) \tag{3.26}
\end{align*}
$$

As shown in [PY2], the pair $\left(A^{+}, B^{+}\right)$(resp. $\left(A^{-}, B^{-}\right)$) implicitly defines an affine-like map $F^{+}$ (resp. $F^{-}$).

Denote by $P^{+}$(resp. $P^{-}$) the domain of $F^{+}$and by $Q^{+}$(resp. $Q^{-}$) the domain of $F^{-}$. Then $P^{+}$ and $P^{-}$are the two components of $P_{0} \cap\left(G \circ F_{0}\right)^{-1}\left(P_{1}\right), Q^{+}$and $Q^{-}$are the two components of $Q_{1} \cap\left(F_{1} \circ G\right)\left(Q_{0}\right) ; F^{+}\left(\right.$resp. $\left.F^{-}\right)$is the restriction of $F_{1} \circ G \circ F_{0}$ to $P^{+}\left(\right.$resp. $\left.P^{-}\right)$.

The formulas for the partial derivatives of $A^{ \pm}, B^{ \pm}$are derived in [PY2] and recalled in Appendix A. They provide the following estimate for the widths:

$$
\begin{align*}
& C^{-1} \leqslant \frac{\left|P^{ \pm}\right|}{\left|P_{0}\right|\left|P_{1}\right| \delta^{-\frac{1}{2}}} \leqslant C  \tag{3.27}\\
& C^{-1} \leqslant \frac{\left|Q^{ \pm}\right|}{\left|Q_{0}\right|\left|Q_{1}\right| \delta^{-\frac{1}{2}}} \leqslant C \tag{3.28}
\end{align*}
$$

where the constants are uniform once $b$ is fixed and the distortions are uniformly bounded.

From [PY2, Theorem 3.7], we also have the following estimate for the distortion of $F^{ \pm}$: assuming that $b$ is small enough (in terms of the partial derivatives of first order of $X_{u}, Y_{s}, \theta$ ), we have

$$
\begin{equation*}
D\left(F^{ \pm}\right) \leq \max \left\{D\left(F_{0}\right)+C\left|Q_{0}\right| \delta^{-1}, D\left(F_{1}\right)+C\left|P_{1}\right| \delta^{-1}\right\} \tag{3.29}
\end{equation*}
$$

provided that $D\left(F_{0}\right)+D\left(F_{1}\right) \leq \delta^{-\frac{1}{2}}$. The constant $C$ in (3.29) depends only on the partial derivatives of first order of $X_{s}, Y_{u}, \theta$.

We also recall from [PY2, formula 3.50] the estimate:

$$
\begin{align*}
& \left|A_{y}^{ \pm}-A_{0, y}\right| \leq C\left|P_{0}\right|\left|Q_{0}\right| \delta^{-\frac{1}{2}}  \tag{3.30}\\
& \left|B_{x}^{ \pm}-B_{1, x}\right| \leq C\left|P_{1}\right|\left|Q_{1}\right| \delta^{-\frac{1}{2}} \tag{3.31}
\end{align*}
$$

where the right-hand terms must be small by (PC2).
As a concluding remark for this section, let us observe that, while conditions (PC1), (PC2) are necessary in order to consider parabolic composition, they are not sufficient: in Section 5, the requirement for parabolic composition will be much more restrictive than (PC2).

## 4 Structure of Parameter Space

### 4.1 Some Important Constants

Throughout the rest of the paper, we will use four main constants $\varepsilon_{0}, \eta, \tau, \beta$ which satisfy

$$
\begin{equation*}
0<\varepsilon_{0} \ll \eta \ll \tau \ll \beta-1<1 \tag{4.1}
\end{equation*}
$$

We roughly explain the meaning of each constant:

- $\varepsilon_{0}$ is the maximal width of the parabolic tongues $L_{u}, L_{s}$. It is also the size of the parameter interval we start with.
- $\eta$ is involved in the transversality relation (defined in Section 5) which allows parabolic composition: instead of the condition (PC2) of Subsection 3.5 , roughly speaking we will ask that

$$
\begin{equation*}
\delta \geqslant\left(\left|P_{1}\right|+\left|Q_{0}\right|\right)^{1-\eta} \tag{4.2}
\end{equation*}
$$

- $\tau$ relates the successive scales of the parameter intervals we will consider through the formula $\varepsilon_{k+1}=\varepsilon_{k}^{1+\tau}$.
- $\beta$ will actually be given in Section 9 by an explicit formula in terms of $d_{s}^{0}$, $d_{u}^{0}$; the condition (H4) involving $d_{s}^{0}, d_{u}^{0}$ in Section 1 is required because we need $\beta>1$. It appears in the definition of regularity in Section 5, which controls the recurrence of the "critical locus".


### 4.2 One-Parameter Families

From now on, we fix a one-parameter family $\left(g_{t}\right)_{t \in\left(-t_{0}, t_{0}\right)}$ in $\mathcal{U}$. We assume that the family is transverse to $\mathcal{U}_{0}$ at $t=0$, with $g_{t} \in \mathcal{U}_{+}$for $t>0$ and $g_{t} \in \mathcal{U}_{-}$for $t<0$.

Observe that $g_{0}$ satisfies exactly the same assumptions as $f$, provided $\mathcal{U}$ is small enough. Therefore, we may and shall, assume that $g_{0}=f$.

We will first reparametrize the family in order to make some computations simpler. Consider the folding map $G_{t}=g_{t}^{N_{0}}$ of Subsection 2.3. If $t_{0}$ is small enough, $G_{t}$ is a fold map for all values of $t \in\left(-t_{0}, t_{0}\right)$. Moreover, we can in properties (P2), (P3) of Subsection 2.3 choose a function $\theta$ which depends smoothly on $t$.

From (MP3), Subsection 2.2, the values $y_{u}=0, x_{s}=0$ of the arguments of $\theta$ correspond to $W_{\text {loc }}^{u}\left(p_{u}\right)$ and $W_{\text {loc }}^{s}\left(p_{s}\right)$ respectively. Therefore, the transversality of the family to $\mathcal{U}_{0}$ is equivalent to

$$
\begin{equation*}
\left.\frac{\partial}{\partial t} \theta_{t}(0,0)\right|_{t=0}>0 \tag{4.3}
\end{equation*}
$$

Taking $t_{0}$ small enough, we can therefore reparametrize our family in order to have

$$
\begin{equation*}
\theta_{t}(0,0) \equiv t, \quad t \in\left(-t_{0}, t_{0}\right) \tag{4.4}
\end{equation*}
$$

### 4.3 Parameter Intervals

The starting parameter interval will be

$$
\begin{equation*}
I_{0}:=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right], \tag{4.5}
\end{equation*}
$$

where, as explained above, $\varepsilon_{0}$ will be taken very small. This is the only parameter interval at level 0 .

At level $k$, we will deal with parameter intervals of length $\varepsilon_{k}$, where the sequence of scales $\varepsilon_{k}$ is defined inductively by

$$
\begin{equation*}
\varepsilon_{k+1}=\varepsilon_{k}^{1+\tau} . \tag{4.6}
\end{equation*}
$$

The constant $\tau$ is small, but $\varepsilon_{0}$ is much smaller and in particular we will have $\varepsilon_{0}^{\tau^{2}} \ll 1$. Every parameter interval of level $k$ is divided into $\left[\varepsilon_{k}^{-\tau}\right]$ parameter intervals of level $k+1$.

The remaining part, if any, is discarded; it is of length $<\varepsilon_{k+1}$; the total length discarded in this way is smaller than $\varepsilon_{1} \ll \varepsilon_{0}$.

Let $\widetilde{I}$ be a parameter interval of level $k$ and $I$ be a parameter interval of level $k+1$ contained in $\widetilde{I}$. We say that $\widetilde{I}$ is the parent of $I$ and that $I$ is a child of $\widetilde{I}$.

### 4.4 The Selection Process

In Section 5, we will define what it means for a parameter interval to be regular. The starting interval $I_{0}$ will be regular.

Given a regular parameter interval $\widetilde{I}$ of level $k$, we divide it into its children: these parameter intervals of level $k+1$ are the candidates. We then test each candidate for regularity and discard those which are not regular. We then proceed to level $k+1$ with each surviving candidate.

The regular parameters are those which are the intersection of a decreasing sequence of regular parameter intervals. For such parameters, we are able to carry out some analysis of the maximal invariant set $\Lambda_{g_{t}}$.

### 4.5 Strongly Regular Parameters

The regularity property is, in some sense, the minimal requirement that is needed to keep control on the geometry and dynamics of the maximal invariant set. However, this requirement is of an essentially qualitative character and this leads in particular to the following difficulty: we are not able to estimate which proportion of the children of a regular parameter interval are also regular.

To circumvent this problem, we define in Section 9 a stronger property for parameter intervals, called strong regularity. It implies regularity, and is better adapted to the inductive selection process. It also gives additional geometric information on the maximal invariant set.

When $\widetilde{I}$ is a strongly regular parameter interval of level $k$, we will show in Section 9 that most candidates of level $k+1$ contained in $\widetilde{I}$ are also strongly regular. The proportion of discarded candidates is less than $\alpha_{k}$, with

$$
\begin{equation*}
\sum_{k \geqslant 0} \alpha_{k} \ll 1, \tag{4.7}
\end{equation*}
$$

the $\ll$ sign means that the sum gets arbitrarily small as $\varepsilon_{0}$ goes to zero. Then we can conclude that most parameters are strongly regular in the sense that they are equal to the intersection of decreasing sequences of strongly regular parameter intervals.

The non-uniformly hyperbolic horseshoes that are the subject of our study are exactly the maximal invariant set $\Lambda_{g}$ for strongly regular $g \in \mathcal{U}_{+}$.

## 5 Classes of Affine-like Iterates and the Transversality Relation

### 5.1 Affine-like Iterates

Let $I$ be a parameter interval of some level.

Definition. An I-persistent affine-like iterate is a triple $(P, Q, n)$ such that

- $P$ is a vertical strip in some $R_{a}$, depending smoothly on $t \in I$;
- $Q$ is a horizontal strip in some $R_{a^{\prime}}$, depending smoothly on $t \in I$;
- $n$ is a nonnegative integer;
- for each $t \in I$, the restriction of $g_{t}^{n}$ to $P_{t}$ is an affine-like map onto $Q_{t}$, i.e. property (AL1) of Subsection 3.1 holds;
- for each $t \in I$, each $m \in[0, n]$, we have $g_{t}^{m}\left(P_{t}\right) \subset \widehat{R}$.


## Examples.

1. For $n=0$, the $I$-persistent affine-like iterates are the $\left(R_{a}, R_{a}, 0\right), a \in a$.
2. For $n=1$, the $I$-persistent affine-like iterates are the $\left(P_{a a^{\prime}}, Q_{a a^{\prime}}, 1\right),\left(a, a^{\prime}\right) \in \mathcal{B}$.
3. More generally, for any finite word $\underline{a}=\left(a_{0}, \ldots, a_{n}\right)$ with transitions in $\mathcal{B}$, the map $g_{a}$ of Subsection 3.4 defined an $I$-persistent affine-like iterate $\left(P_{\underline{a}}, Q_{\underline{a}}, n\right)$.

Notation. If $P$ is a vertical strip $\left\{\varphi_{-}(y) \leqslant x \leqslant \varphi_{+}(y)\right\}$ we denote by $\partial P$ the vertical part of the boundary, i.e. the two graphs $\left\{x=\varphi^{ \pm}(y)\right\}$. Similarly for horizontal strips.

If $(P, Q, n)$ is an $I$-persistent affine-like iterate and $I^{\prime}$ is a parameter interval contained in $I,(P, Q, n)$ also defines by restriction an $I^{\prime}$-persistent affine-like iterate. A slightly less trivial property is given by

Proposition 1. Let $(P, Q, n),\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ be I-persistent affine-like iterates. We have
a) if $n=n^{\prime}$, then either $P=P^{\prime}$ and $Q=Q^{\prime}$ for all $t \in I$ or $P \cap P^{\prime}=\emptyset, Q \cap Q^{\prime}=\emptyset$ for all $t \in I$.
b) if $n<n^{\prime}$, then either $P \supset P^{\prime}, \partial P \cap P^{\prime}=\emptyset$ for all $t \in I$ or $P \cap P^{\prime}=\emptyset$, for all $t \in I$.

Remark. Throughout the paper, except in Section 9 (where we break the symmetry assuming $d_{s}^{0} \geqslant d_{u}^{0}$ ), we will keep a time-symmetric setting. Thus every property stated for the domains $P^{\prime} s$ is also valid for the images $Q^{\prime}$ s. This apply for instance to part b) of the proposition.

Proof. By the definition of an $I$-persistent affine-like iterate, for all $t \in I, P$ is a connected component of $R \cap g_{t}^{-n}(R)$ and also of $\bigcap_{0 \leqslant m \leqslant n} g_{t}^{-m}(\widehat{R})$.
a) If $n=n^{\prime}$ and $P \cap P^{\prime} \neq \emptyset$ for some $t_{0} \in I$, we must have $P=P^{\prime}$ at $t_{0}$ and hence, $P \cap P^{\prime} \neq \emptyset$ for $t$ close to $t_{0}$. It follows that $P=P^{\prime}$ for all $t \in I$, and also $Q=Q^{\prime}$ for all $t \in I$.
b) Assume that $n<n^{\prime}$ and $P \cap P^{\prime} \neq \emptyset$ for some $t_{0} \in I$, then $P^{\prime} \subset P$ at $t_{0}$ (since $\bigcap_{0 \leqslant m \leqslant n^{\prime}} g_{t_{0}}^{-m}(\widehat{R})$ is contained in $\left.\bigcap_{0 \leqslant m \leqslant n} g_{t_{0}}^{-m}(\widehat{R})\right)$, hence $P^{\prime} \cap P \neq \emptyset$ for $t$ close to $t_{0}$ and $P^{\prime} \subset P$ for all $t \in I$.

Let $t \in I, z \in \partial P$; then, $g_{t}^{n}(z)$ belongs to the vertical boundary of $R$ and $g_{t}^{n+1}(z) \notin \widehat{R}$; therefore, $z \notin P^{\prime}$. This proves that $\partial P \cap P^{\prime}$ is empty for all $t \in I$.

### 5.2 The Classes $\mathcal{R}(I)$ : General Overview

It would be nice to work with the class of all $I$-persistent affine-like iterates, but with this approach one faces two problems:

- I-persistent affine-like iterates do not satisfy a uniform cone condition, and they do not have uniformly bounded distortion;
- even if we force such uniformity in the definition, a major problem is that we lack some control on the way in which long $I$-persistent affine-like iterates are constructed from shorter ones by simple or parabolic composition.

To overcome these problems, we will define, for every parameter interval (whose parent is regular, see Subsection 4.4 and the end of the present section on the definition of regularity) a subset $\mathcal{R}(I)$ of the set of all $I$-persistent affine-like iterates; all elements of $\mathcal{R}(I)$ with $n>1$ can be, from the very definition, obtained from shorter ones by simple or parabolic composition; the elements of $\mathcal{R}(I)$ will turn out to satisfy a uniform cone condition and have uniformly bounded distortion.

The main ingredient in the definition of $\mathcal{R}(I)$ is a transversality relation which is an appropriate strengthening of condition (PC2) in Subsection 3.5. Simple composition is allowed whenever it makes sense, but parabolic composition is only allowed when this transversality relation holds.

The definition of the transversality relation, given later in this section, is quite involved; this is because we want some combinatorial properties proved in Section 6 to be satisfied. Such properties make our later work much easier.

All the process of this definition is based on a double induction:

- an induction on the level of the parameter interval, starting with $I_{0}=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]$ at level 0 ;
- for a given parameter interval, an induction on the length $n$ of the $I$-persistent affine-like iterates which are considered.

In order for the definition of $\mathcal{R}(I)$ to be consistent, a number of properties (uniform cone condition, uniform bound on distortion and many others) must hold; unfortunately, the proof of any of these properties is inductive, using several other properties for shorter iterates. Our presentation, is therefore, organized through the following scheme: we will first define $\mathcal{R}(I)$ and list all the required properties conditioning the definition. Then we will proceed to the proofs of these properties, and of other related ones. The whole process will occupy the rest of the present section and Sections 6 and 7 .

### 5.3 Definition of the Special Class of Affine-Like Iterates $\mathcal{R}(I)$

Let $I$ be a parameter interval of some level.

We claim that there exists a (unique) class of $I$-persistent affine-like iterates which satisfy the properties (R1)-(R7) below.
(R1) For any word $\underline{a}=\left(a_{0}, \ldots, a_{n}\right)$ with transitions in $\mathcal{B}$, the element $\left(P_{\underline{a}}, Q_{\underline{a}}, n\right)$ (see example 3 above) belongs to $\mathcal{R}(I)$.

For the starting interval $I_{0}=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]$, it will turn out that one obtains in this way all elements of $\mathcal{R}\left(I_{0}\right)$.

Recall from (MP5), (MP6) in Subsection 3.4 that all $\left(P_{\underline{a}}, Q_{\underline{a}}, n\right)$ with $n>0$ satisfy for all $t \in I_{0}$ a uniform cone condition (AL2) with parameters $\lambda, u, v$ (satisfying $1<u v \leqslant \lambda^{2}$ ), and have distortion bounded by $D_{0}$. Let $u_{0}=u^{\frac{1}{2}}, v_{0}=v^{\frac{1}{2}}$.
(R2) All $(P, Q, n) \in \mathcal{R}(I)$ satisfy for all $t \in I$ the cone condition (AL2) with parameters $\lambda, u_{0}, v_{0}$ and have distortion bounded by $2 D_{0}$ for all $t \in I$.

Let $(P, Q, n),\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ be elements of $\mathcal{R}(I)$ such that $Q \subset R_{a}, P^{\prime} \subset R_{a}$ for some $a \in a$. As both iterates satisfy the cone condition (AL2) with parameters $\lambda, u_{0}$, $v_{0}$, we know from Subsection 3.3 that the simple composition defined by

$$
\begin{equation*}
P^{\prime \prime}=P \cap g^{-n}\left(P^{\prime}\right), \quad Q^{\prime \prime}=Q^{\prime} \cap g^{n^{\prime}}(Q), \quad n^{\prime \prime}=n+n^{\prime} \tag{5.1}
\end{equation*}
$$

is an (I-persistent) affine-like iterate.

The next condition states that it should also belong to $\mathcal{R}(I)$.
(R3) The class $\mathcal{R}(I)$ is stable under simple composition.

We now turn to parabolic composition.

We first define two special elements which belong to $\mathcal{R}(I)$ according to (R1): define $\left(P_{s}, Q_{s}, n_{s}\right)$ (resp. $\left.\left(P_{u}, Q_{u}, n_{u}\right)\right)$ to be the element $\left(P_{\underline{a}}, Q_{\underline{a}}, n\right)$ with maximal length $n$ such that $L_{s} \subset P_{s}$ for all $t \in I_{0}$ (resp. $L_{u} \subset Q_{u}$ for all $t \in I_{0}$ ). We have that $p_{s} \in P_{s}$ and $p_{u} \in Q_{u}$. See figure 7 .


Figure 7

We obviously have, for all $t \in I_{0}$

$$
\begin{align*}
& C^{-1} \varepsilon_{0} \leqslant\left|P_{s}\right| \leqslant C \varepsilon_{0}  \tag{5.2}\\
& C^{-1} \varepsilon_{0} \leqslant\left|Q_{u}\right| \leqslant C \varepsilon_{0}
\end{align*}
$$

The next condition guarantees that property (PC1) in Subsection 3.5 is satisfied.
(R4) Let $(A, B)$ be the implicit representation of an affine-like iterate $(P, Q, n) \in \mathcal{R}(I)$.
a) If $P \subset P_{s}$, then for all $t \in I$ we have

$$
\left|A_{y}\right| \leqslant C \varepsilon_{0}, \quad\left|A_{y y}\right| \leqslant C \varepsilon_{0}
$$

b) If $Q \subset Q_{u}$, then for all $t \in I$ we have

$$
\left|B_{x}\right| \leqslant C \varepsilon_{0}, \quad\left|B_{x x}\right| \leqslant C \varepsilon_{0}
$$

Here and in the sequel, the letter $C$ denotes various constants which depend only on our initial diffeomorphism $f$, but not on $\eta, \tau, \varepsilon_{0}$.

Let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements in $\mathcal{R}(I)$ with $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$. In these circunstances, we will define in Subsection 5.4 a transversality relation denoted by $Q_{0} \pitchfork_{I} P_{1}$ which may or may not hold. When it holds, it implies condition (PC2) of Subsection 3.5 for all $t \in I$ (see (R7) below).
(R5) If $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ as above satisfy $Q_{0} \pitchfork_{I} P_{1}$, then both $I$-persistent affine-like iterates obtained from the parabolic composition $g_{t}^{n_{1}} \circ G_{t} \circ g_{t}^{n_{0}}$ belong to $\mathcal{R}(I)$.

Writing $\left(P^{+}, Q^{+}, n\right)$ and $\left(P^{-}, Q^{-}, n\right)$ for these two iterates, we have $n=n_{0}+n_{1}+N_{0}$. The domains $P^{+}$and $P^{-}$are the two components of $g^{-n_{0}}\left(Q_{0} \cap G_{t}^{-1}\left(P_{1}\right)\right)$; the images $Q^{+}$and $Q^{-}$are the two components of $g^{n_{1}}\left(P_{1} \cap G_{t}\left(Q_{0}\right)\right)$. See figure 6, Subsection 3.5.

When $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ satisfy $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}, Q_{0} \pitchfork_{I} P_{1}$, we say that their parabolic composition is allowed in $\mathcal{R}(I)$.
(R6) Any $(P, Q, n) \in \mathcal{R}(I)$ with $n>1$ can be obtained from shorter elements by simple composition or (allowed) parabolic composition.

Typically, an element of $\mathcal{R}(I)$ can be obtained in many ways by composition of shorter ones. We say that an element of $\mathcal{R}(I)$ is prime if it cannot be obtained by simple composition of shorter ones. Prime elements play a key role in the description of the dynamics for regular parameters in Section 10.

It is pretty clear from conditions (R1), (R3), (R5), (R6) alone that there is at most one class $\mathcal{R}(I)$ satisfying these conditions. The existence of $\mathcal{R}(I)$, i.e. the proof of the consistency of conditions (R1)-(R6), is much more delicate. There is actually a seventh property (R7) formulated in the next subsection and related to the condition (PC2) for parabolic composition.

## Parent-Child Terminology and Notations for Compositions.

Let $(P, Q, n),(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}(I)$ with $P \subset \widetilde{P}, n>\widetilde{n}$. If there is no $(\widehat{P}, \widehat{Q}, \widehat{n}) \in \mathcal{R}(I)$ with $P \subset \widehat{P} \subset \widetilde{P}$ and $n>\widehat{n}>\widetilde{n}$, we say that $P$ is a child of $\widetilde{P}$ and that $\widetilde{P}$ is the parent of $P$; if moreover $n=\widetilde{n}+1$, we say that $P$ is a simple child; otherwise it is a non-simple child.

Let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$. If $Q_{0}, P_{1}$ are contained in a same rectangle $R_{a}$, the simple composition $(P, Q, n) \in \mathcal{R}(I)$ of these elements will be written as

$$
\begin{equation*}
(P, Q, n)=\left(P_{0}, Q_{0}, n_{0}\right) \quad * \quad\left(P_{1}, Q_{1}, n_{1}\right) \tag{5.3}
\end{equation*}
$$

If $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$ and $Q_{0} \pitchfork_{I} P_{1}$, any of the two elements ( $\widehat{P}, \widehat{Q}, \widehat{n}$ ) obtained by the corresponding allowed parabolic composition will be written as

$$
\begin{equation*}
(\widehat{P}, \widehat{Q}, \widehat{n}) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(P_{1}, Q_{1}, n_{1}\right) \tag{5.4}
\end{equation*}
$$

### 5.4 Definition of the Transversality Relation

Let $I$ be a parameter interval of some level, and let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}(I)$ which satisfy $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$.

From (R4) the condition (PC1) of Subsection 3.5 is satisfied provided $\varepsilon_{0}$ small enough. Denote by $\left(x_{0}, y_{0}\right)$ (resp. $\left.\left(x_{1}, y_{1}\right)\right)$ the coordinates in the rectangle containing $P_{0}$ (resp. $Q_{1}$ ). A function $\bar{C}\left(y_{0}, x_{1}\right)$ was defined in Subsection 3.5, together with

$$
\begin{equation*}
\delta\left(Q_{0}, P_{1}\right)=\min _{y_{0}} \min _{x_{1}}-\bar{C}\left(y_{0}, x_{1}\right) . \tag{5.5}
\end{equation*}
$$

In Subsection 3.5, we were asking for $\delta$ to be much larger than $\left|P_{0}\right|$ and $\left|Q_{1}\right|$. From the formulas of Appendix A, we have

$$
\begin{align*}
& C^{-1}\left|P_{1}\right| \leqslant\left|\bar{C}_{x}\right| \leqslant C\left|P_{1}\right|,  \tag{5.6}\\
& C^{-1}\left|Q_{0}\right| \leqslant\left|\bar{C}_{y}\right| \leqslant C\left|Q_{0}\right| . \tag{5.7}
\end{align*}
$$

We will introduce

$$
\begin{align*}
\delta_{L}\left(Q_{0}, P_{1}\right) & :=\max _{y_{0}} \min _{x_{1}}-\bar{C}\left(y_{0}, x_{1}\right),  \tag{5.8}\\
\delta_{R}\left(Q_{0}, P_{1}\right) & :=\min _{y_{0}} \max _{x_{1}}-\bar{C}\left(y_{0}, x_{1}\right),  \tag{5.9}\\
\delta_{L R}\left(Q_{0}, P_{1}\right) & :=\max _{y_{0}} \max _{x_{1}}-\bar{C}\left(y_{0}, x_{1}\right) . \tag{5.10}
\end{align*}
$$

All together, $\delta, \delta_{L}, \delta_{R}, \delta_{L R}$ are the values of $-\bar{C}$ at the four corners of the rectangle of definition of $\bar{C}$. We have from (5.6), (5.7) that

$$
\begin{align*}
& C^{-1}\left|Q_{0}\right| \leqslant \delta_{L}\left(Q_{0}, P_{1}\right)-\delta\left(Q_{0}, P_{1}\right) \leqslant C\left|Q_{0}\right|,  \tag{5.11}\\
& C^{-1}\left|P_{1}\right| \leqslant \delta_{R}\left(Q_{0}, P_{1}\right)-\delta\left(Q_{0}, P_{1}\right) \leqslant C\left|P_{1}\right|,  \tag{5.12}\\
& C^{-1}\left|Q_{0}\right| \leqslant \delta_{L R}\left(Q_{0}, P_{1}\right)-\delta_{R}\left(Q_{0}, P_{1}\right) \leqslant C\left|Q_{0}\right|,  \tag{5.13}\\
& C^{-1}\left|P_{1}\right| \leqslant \delta_{L R}\left(Q_{0}, P_{1}\right)-\delta_{L}\left(Q_{0}, P_{1}\right) \leqslant C\left|P_{1}\right| . \tag{5.14}
\end{align*}
$$

Preliminary Definition. We write $Q_{0} \bar{\Pi}_{I} P_{1}$ if the following holds
(T1) for all $t \in I$,

$$
\delta_{L R}\left(Q_{0}, P_{1}\right) \geqslant 2|I|,
$$

(T2) for some $t_{0} \in I$,

$$
\delta_{R}\left(Q_{0}, P_{1}\right) \geqslant 2\left|Q_{0}\right|^{1-\eta}
$$

(T3) for some $t_{1} \in I$,

$$
\delta_{L}\left(Q_{0}, P_{1}\right) \geqslant 2\left|P_{1}\right|^{1-\eta} .
$$

## Definition.

We say that $Q_{0}, P_{1}$ are $I$-transverse and write $Q_{0} \pitchfork_{I} P_{1}$ if there exist a parameter interval $\tilde{I} \supset I$, elements $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}(\widetilde{I})$ with $\widetilde{P}_{1} \supset P_{1}, \widetilde{Q}_{0} \supset Q_{0}$ such that $\widetilde{Q}_{0} \bar{\Pi}_{\widetilde{I}} \widetilde{P}_{1}$.

## Remark.

1. Taking $\widetilde{I}=I, \widetilde{P}_{0}=P_{0}, \widetilde{Q}_{1}=Q_{1}$, it is obvious that if $Q_{0} \overleftarrow{\Pi}_{I} P_{1}$, then $Q_{0} \pitchfork_{I} P_{1}$.
2. In view of our inductive procedure, all $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ which have to be considered have been constructed before $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$.
3. As mentioned before, the definition of the transversality relation is quite involved. Some justification for the choice of quantifiers in (T1), (T2), (T3) can be found in Appendix C.

At first sight, it appears that properties (T2), (T3) above are not quite sufficient to guarantee condition (PC2) of parabolic composition (Subsection 3.5), because they involve only one value of the parameter. The next property takes care of this problem.
(R7) If $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$ satisfy $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$ and $Q_{0} \pitchfork_{I} P_{1}$ holds, then, for all $t \in I$, we have

$$
\delta\left(Q_{0}, P_{1}\right) \geqslant C^{-1}\left(\left|P_{1}\right|^{1-\eta}+\left|Q_{0}\right|^{1-\eta}\right)
$$

### 5.5 The Class $\mathcal{R}\left(I_{0}\right)$

Recall that in Subsection 4.2 we had the normalization

$$
\begin{equation*}
\theta(0,0, t) \equiv t, \quad|t| \leqslant t_{0} \tag{4.4}
\end{equation*}
$$

As $\theta$ is monotonous in both variables, it follows from the definition of $C$ (formula (3.18) of Subsection 3.5) that for all $t \in\left(-t_{0}, t_{0}\right)$ :

$$
\begin{equation*}
-\bar{C}\left(y_{0}, x_{1}, t\right) \leqslant t \tag{5.15}
\end{equation*}
$$

Therefore, for the starting interval $I_{0}=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]$, condition (T1) above is never satisfied.

Parabolic composition is never allowed in $\mathcal{R}\left(I_{0}\right)$. Thus (see Subsection 5.3) the class $\mathcal{R}\left(I_{0}\right)$ consists only of the elements $\left(P_{\underline{a}}, Q_{\underline{a}}, n\right)$ given by (R1) which are associated with the horseshoe $K$.

Condition (R2) is satisfied because of the choice of $u_{0}, v_{0}, D_{0}$. Condition (R3) is obviously satisfied, as are conditions (R5) and (R6). Condition (R4) will be checked in Section 7.

### 5.6 Definition of the Regularity Property

We introduce first some terminology and some concepts related to the transversality relation.
5.6.1 Let $I$ be a parameter interval of some level, and let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}(I)$ such that $Q_{0} \subset Q_{u}$ and $P_{1} \subset P_{s}$. When $Q_{0}$ and $P_{1}$ are not $I$-transverse, two cases may happen:

- if $G_{t}\left(Q_{0}\right) \cap P_{1}=\emptyset$ for all $t \in I$, we say that $Q_{0}$ and $P_{1}$ are $I$-separated;
- otherwise, we say that $Q_{0}$ and $P_{1}$ are I-critically related.
5.6.2 Let $(P, Q, n) \in \mathcal{R}(I)$. An I-decomposition of $P$ is a finite family $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of elements of $\mathcal{R}(I)$ such that the $P_{\alpha}^{\prime}$ s are disjoint, contained in $P$ and satisfy

$$
\begin{equation*}
W^{s}(\Lambda, \widehat{R}) \cap P=\bigsqcup_{\alpha}\left(W^{s}(\Lambda, \widehat{R}) \cap P_{\alpha}\right), \tag{5.16}
\end{equation*}
$$

where $W^{s}(\Lambda, \widehat{R})$ was defined in Subsection 2.2. We say that $P$ is $I$-decomposable if it admits a non-trivial $I$-decomposition. Then, there is a coarsest non-trivial $I$-decomposition, namely by the children of $P$, which is called the canonical I-decomposition.

Remark. We will see in Section 8 that any $P$ has only finitely many children.
5.6.3 Let $(P, Q, n) \in \mathcal{R}(I)$. We say that $Q$ is $I$-transverse if either $Q \cap Q_{u}=\emptyset$ or $Q \subset Q_{u}$ and there exists an $I$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $P_{s}$ such that, for any $\alpha, Q$ and $P_{\alpha}$ are either $I$-transverse or $I$-separated.

We say that $Q$ is $I$-critical when it is not $I$-transverse. This is always the case if $Q \supset Q_{u}$.
We also define in a symmetric way an $I$-decomposition for $Q$, and $I$-transversality or $I$-criticality for $P$.
5.6.4 We say that $(P, Q, n) \in \mathcal{R}(I)$ is $I$-bicritical if both $P$ and $Q$ are $I$-critical. The corresponding iterate should be seen as describing some recurrence to the "critical locus".

Definition. Let $\beta>1$. We say that the parameter interval $I$ is $\beta$-regular (or just regular when the value of $\beta$ is fixed) if any $I$-bicritical element $(P, Q, n) \in \mathcal{R}(I)$ satisfies, for all $t \in I$ :

$$
\begin{equation*}
|P|<|I|^{\beta}, \quad|Q|<|I|^{\beta} . \tag{5.17}
\end{equation*}
$$

## Remark.

1. When $d_{s}^{0} \geqslant d_{u}^{0}$, in Section 9 we will take for $\beta$ a number satisfying

$$
\begin{equation*}
1<\beta<\beta_{\max }:=\frac{\left(1-d_{u}^{0}\right)\left(d_{s}^{0}+d_{u}^{0}\right)}{d_{s}^{0}\left(2 d_{s}^{0}+d_{u}^{0}-1\right)} . \tag{5.18}
\end{equation*}
$$

Condition (H4) in Section 1.2 is actually equivalent to $\beta_{\max }>1$. When $d_{u}^{0} \geqslant d_{s}^{0}$, we exchange $d_{s}^{0}$ and $d_{u}^{0}$ in the definition of $\beta_{\text {max }}$.
2. We will see in Section 6 that if $\widetilde{I}$ is a $\beta$-regular parameter interval, and $I \subset \widetilde{I}$ is a candidate at the next level, then I is at least $\bar{\beta}$-regular with

$$
\begin{equation*}
\bar{\beta}=\beta(1+\tau)^{-1} . \tag{5.19}
\end{equation*}
$$

This will allow us to apply to candidates, with only slightly worse constants, all results that have been proven for regular parameter intervals. But, obviously, we cannot let the regularity exponent to deteriorate too much (it must stay $>1$ ), which explains why candidates have to pass the $\beta$ regularity test.

## 6 Some Properties of the Classes $\mathcal{R}(I)$

We recall that all parameter intervals that we consider in the sequel are assumed to be regular or candidates (meaning in this case that their parent is regular).

### 6.1 Transversality is Hereditary

The following obvious but fundamental property was forced into the definition of the transversality relation.

Proposition 2. Let $\widetilde{I} \supset I$ be parameter intervals. Let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$ and $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}(\widetilde{I})$. Assume that $Q_{0} \subset \widetilde{Q}_{0} \subset Q_{u}$ and $P_{1} \subset \widetilde{P}_{1} \subset P_{s}$. If $\widetilde{Q}_{0}$ and $\widetilde{P}_{1}$ are $\widetilde{I}$-transverse, then $Q_{0}$ and $P_{1}$ are I-transverse.

Corollary 1. Let $\widetilde{I} \supset I$ be parameter intervals, and let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(\widetilde{I}) \cap \mathcal{R}(I)$ be such that $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$. If their parabolic composition is allowed in $\mathcal{R}(\widetilde{I})$, it is also allowed in $\mathcal{R}(I)$.

Proof. This is the case $\widetilde{Q}_{0}=Q_{0}, \widetilde{P}_{1}=P_{1}$ of the proposition.
Corollary 2. Let $\widetilde{I} \supset I$ be parameter intervals. Then $\mathcal{R}(\widetilde{I}) \subset \mathcal{R}(I)$.
Remark. This is a slight abus de langage of no consequence: properly speaking, we mean that the restriction to I of any $(P, Q, n) \in \mathcal{R}(\widetilde{I})$ belongs to $\mathcal{R}(I)$.

Proof. As simple composition does not depend on the parameter interval, Corollary 2 is an immediate consequence by induction on length of property (R6).

Corollary 3. Let $\widetilde{I} \supset I$ be parameter intervals, and let $(P, Q, n) \in \mathcal{R}(\widetilde{I})$. If $Q$ is $\widetilde{I}$-transverse, then it is also I-transverse.

Proof. If $Q \cap Q_{u}=\emptyset$, this is obvious. Assume therefore that $Q \subset Q_{u}$. Then there exists an $\widetilde{I}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)_{\alpha}$ of $P_{s}$ by elements of $\mathcal{R}(\widetilde{I})$ such that for all $\alpha, Q$ and $P_{\alpha}$ are either $\widetilde{I}$-transverse or $\widetilde{I}$-separated.

First observe that $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{R}(I)$ and therefore this is also an $I$-decomposition of $P_{s}$. By Corollary 1, if $Q$ and $P_{\alpha}$ are $\widetilde{I}$-transverse, they are also $I$-transverse. On the other hand, it is obvious from the definition that if $Q$ and $P_{\alpha}$ are $\widetilde{I}$-separated they are also $I$-separated. The result follows.

### 6.2 Criticality and Decomposability

Proposition 3. Let I be a parameter interval, and let $(P, Q, n) \in \mathcal{R}(I)$. If $Q$ is I-transverse, then $P$ is I-decomposable.

Proof. Let us first assume that $Q \cap Q_{u}=\emptyset$. Let $a \in a$ be such that $Q \subset R_{a}$. We have

$$
\begin{equation*}
R_{a} \cap W^{s}(\Lambda, \widehat{R}) \subset \bigcup_{\left(a, a^{\prime}\right) \in \mathcal{B}}\left(P_{a, a^{\prime}} \cap W^{s}(\Lambda, \widehat{R})\right) \cup L_{u} ; \tag{6.1}
\end{equation*}
$$

for each $a^{\prime} \in a$ such that $\left(a, a^{\prime}\right) \in \mathcal{B}$, we have the simple child of $P$ :

$$
\begin{equation*}
\left(P\left(a^{\prime}\right), Q\left(a^{\prime}\right), n+1\right)=(P, Q, n) *\left(P_{a a^{\prime}}, Q_{a a^{\prime}}, 1\right), \tag{6.2}
\end{equation*}
$$

and together they form by (6.1) an $I$-decomposition of $P$ (the canonical one).
Let us now assume that $Q \subset Q_{u}$. As $Q$ is $I$-transverse, there is an $I$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)_{\alpha}$ of $P_{s}$ such that, for each $\alpha, Q$ and $P_{\alpha}$ are not $I$-critically related. For each $\alpha$ such that $Q$ and $P_{\alpha}$ are $I$-transverse, let $\left(P_{\alpha}^{ \pm}, Q_{\alpha}^{ \pm}, n_{\alpha}+n+N_{0}\right)$ be the two elements produced by the allowed parabolic composition. Together with the simple children defined by (6.2), they form an $I$-decomposition of $P$.

Corollary 4. Let I be a $\beta$-regular parameter interval and let $(P, Q, n) \in \mathcal{R}(I)$. If $P$ is $I$-critical and $|P|>|I|^{\beta}$ or $|Q|>|I|^{\beta}$ for some $t \in I$, then $P$ is $I$-decomposable.

Proof. Indeed, by the very definition of regularity, $Q$ cannot be $I$-critical.

The decomposability of "fat" critical rectangles is crucial to our analysis. As mentioned before, Corollary 4 will apply to candidate intervals with $\bar{\beta}=\beta(1+\tau)^{-1}$ instead of $\beta$.

### 6.3 Concavity

The following result is a partial converse to Proposition 2. We call concavity the corresponding property of $\mathcal{R}(I)$.

Proposition 4. Let I be a parameter interval. Let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{0}^{\prime}, Q_{0}^{\prime}, n_{0}^{\prime}\right),\left(P_{1}, Q_{1}, n_{1}\right)\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)$ be elements of $\mathcal{R}(I)$ such that

$$
Q_{0} \subset Q_{0}^{\prime} \subset Q_{u}, \quad P_{1} \subset P_{1}^{\prime} \subset P_{s}
$$

If both $Q_{0} \pitchfork_{I} P_{1}^{\prime}$ and $Q_{0}^{\prime} \pitchfork_{I} P_{1}$ hold, then $Q_{0}^{\prime} \pitchfork_{I} P_{1}^{\prime}$ also holds.

Proof. From the definition (3.18) of the function $C$ and the monotonicity of $\theta$ with respect to each variable, it follows that we have, for each $t \in I$ :

$$
\begin{align*}
& \delta\left(Q_{0}, P_{1}\right) \geqslant \max \left(\delta\left(Q_{0}^{\prime}, P_{1}\right), \delta\left(Q_{0}, P_{1}^{\prime}\right)\right)  \tag{6.3}\\
& \delta_{L}\left(Q_{0}^{\prime}, P_{1}\right) \geqslant \delta_{L}\left(Q_{0}, P_{1}\right) \geqslant \delta_{L}\left(Q_{0}, P_{1}^{\prime}\right)  \tag{6.4}\\
& \delta_{R}\left(Q_{0}, P_{1}^{\prime}\right) \geqslant \delta_{R}\left(Q_{0}, P_{1}\right) \geqslant \delta_{R}\left(Q_{0}^{\prime}, P_{1}\right)  \tag{6.5}\\
& \min \left(\delta_{L R}\left(Q_{0}^{\prime}, P_{1}\right), \delta_{L R}\left(Q_{0}, P_{1}^{\prime}\right)\right) \geqslant \delta_{L R}\left(Q_{0}, P_{1}\right) \tag{6.6}
\end{align*}
$$

By definition of the transversality relation, there exist parameter intervals $\widetilde{I}_{1}, \widetilde{I}_{2}$ containing $I$, elements $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right)$, $\left(\widetilde{P}_{1}^{\prime}, \widetilde{Q}_{1}^{\prime}, \widetilde{n}_{1}^{\prime}\right)$ in $\mathcal{R}\left(\widetilde{I}_{1}\right)$, elements $\left(\widetilde{P}_{0}^{\prime}, \widetilde{Q}_{0}^{\prime}, \widetilde{n}_{0}^{\prime}\right)$, ( $\left.\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ in $\mathcal{R}\left(\widetilde{I}_{2}\right)$ such that $Q_{0} \subset \widetilde{Q}_{0}, Q_{1}^{\prime} \subset \widetilde{Q}_{1}^{\prime}, Q_{0}^{\prime} \subset \widetilde{Q}_{0}^{\prime}, P_{1} \subset \widetilde{P}_{1}$ and

$$
\begin{align*}
& \widetilde{Q}_{0} \bar{\Pi}_{\widetilde{I}_{1}} \widetilde{P}_{1}^{\prime}  \tag{6.7}\\
& \widetilde{Q}_{0}^{\prime} \bar{\Pi}_{\widetilde{I}_{2}} \widetilde{P}_{1} \tag{6.8}
\end{align*}
$$

If we have either $Q_{0}^{\prime} \subset \widetilde{Q}_{0}$ or $P_{1}^{\prime} \subset \widetilde{P}_{1}$, we can already conclude that $Q_{0}^{\prime} \pitchfork_{I} P_{1}^{\prime}$. Assume therefore that $\widetilde{Q}_{0} \subset Q_{0}^{\prime}, \widetilde{P}_{1} \subset P_{1}^{\prime}$. Assume also for instance that $\widetilde{I}_{1} \subset \widetilde{I}_{2}$. We have $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}\left(\widetilde{I}_{2}\right)$ and $\widetilde{P}_{1} \subset P_{1}^{\prime}$. By the coherence property (Proposition 6 in Subsection 6.5 ), the element $\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)$ belongs to $\mathcal{R}\left(\widetilde{I}_{2}\right)$. We will show that

$$
\begin{equation*}
\widetilde{Q}_{0}^{\prime} \bar{\Pi}_{\widetilde{I}_{2}} P_{1}^{\prime} \tag{6.9}
\end{equation*}
$$

which implies that $Q_{0}^{\prime}$ and $P_{1}^{\prime}$ are $I$-transverse. We check properties (T1)-(T3) of Subsection 5.5. First, by (6.6) and (6.8), we have, for all $t \in \widetilde{I}_{2}$

$$
\begin{equation*}
\delta_{L R}\left(\widetilde{Q}_{0}^{\prime}, P_{1}^{\prime}\right) \geqslant \delta_{L R}\left(\widetilde{Q}_{0}^{\prime}, \widetilde{P}_{1}\right) \geqslant 2\left|\widetilde{I}_{2}\right| \tag{6.10}
\end{equation*}
$$

Next, by (6.8) and (6.5), there exists $t_{0} \in \widetilde{I}_{2}$ such that

$$
\begin{equation*}
\delta_{R}\left(\widetilde{Q}_{0}^{\prime}, P_{1}^{\prime}\right) \geqslant \delta_{R}\left(\widetilde{Q}_{0}^{\prime}, \widetilde{P}_{1}\right) \geqslant 2\left|\widetilde{Q}_{0}^{\prime}\right|^{1-\eta} \tag{6.11}
\end{equation*}
$$

Finally, by (6.7), there exists $t_{1} \in \widetilde{I}_{1} \subset \widetilde{I}_{2}$ such that

$$
\begin{equation*}
\delta_{L}\left(\widetilde{Q}_{0}, \widetilde{P}_{1}^{\prime}\right) \geqslant 2\left|\widetilde{P}_{1}^{\prime}\right|^{1-\eta} . \tag{6.12}
\end{equation*}
$$

For this value $t_{1}$, we then have, by (6.4):

$$
\begin{equation*}
\delta_{L}\left(\widetilde{Q}_{0}^{\prime}, P_{1}^{\prime}\right) \geqslant \delta_{L}\left(\widetilde{Q}_{0}, \widetilde{P}_{1}^{\prime}\right) \geqslant 2\left|\widetilde{P}_{1}^{\prime}\right|^{1-\eta} \geqslant 2\left|P_{1}^{\prime}\right|^{1-\eta} . \tag{6.13}
\end{equation*}
$$

We have proved (6.9) and this concludes the proof of the proposition.

Remark. The concavity property is very helpful in the sequel. The proof of the proposition shows why the definition of the transversality relation had to be complicated.

Corollary 5. Let I be a parameter interval, $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right),\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)$ be elements of $\mathcal{R}(I)$ such that $Q_{0} \subset Q_{u}, P_{1} \subset P_{1}^{\prime} \subset P_{s}$. If $P_{1}^{\prime}$ is $I$-transverse and $Q_{0} \pitchfork_{I} P_{1}$ holds, then $Q_{0} \pitchfork_{I} P_{1}^{\prime}$ also holds.

Proof. There exists an $I$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $Q_{u}$ such that, for any $\alpha, Q_{\alpha}$ and $P_{1}^{\prime}$ are either $I$-transverse or $I$-separated. There exists $\alpha$ such that $Q_{\alpha}$ and $Q_{0}$ do intersect. As $Q_{0} \pitchfork_{I} P_{1}$ holds, $Q_{\alpha}$ and $P_{1}^{\prime}$ must be $I$-transverse. If $Q_{\alpha} \supset Q_{0}$, it follows from Proposition 2 that $Q_{0}$ and $P_{1}^{\prime}$ are $I$-transverse. If $Q_{\alpha} \subset Q_{0}$, the same conclusion follows from Proposition 4.

### 6.4 Children are Born From Their Parent

Let $I$ be a parameter interval, $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be an element of $\mathcal{R}(I)$ and $a \in a$ the letter such that $\widetilde{Q} \subset \mathcal{R}_{a}$. The simple children of $\widetilde{P}$ are given by formula (6.2) and parametrized by the elements $a^{\prime} \in a$ such that $\left(a, a^{\prime}\right) \in \mathcal{B}$. All children are simple unless $\widetilde{Q} \subset Q_{u}$.

Proposition 5. Assume $\widetilde{Q} \subset Q_{u}$ and let $(P, Q, n)$ be a non-simple child of $\widetilde{P}$. Then there exists $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$ such that $\widetilde{Q} \pitchfork_{I} P_{1}$ and

$$
(P, Q, n) \in(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \square\left(P_{1}, Q_{1}, n_{1}\right) .
$$

Moreover, the parent $\widetilde{P}_{1}$ of $P_{1}$ is I-critical.

Proof.

1. By (R6) in Subsection $5.3,(P, Q, n)$ is obtained from shorter elements by simple or parabolic composition. Let us first assume that we can write

$$
\begin{equation*}
(P, Q, n)=\left(P_{0}, Q_{0}, n_{0}\right) *\left(P_{1}, Q_{1}, n_{1}\right), \tag{6.14}
\end{equation*}
$$

with $n_{0}, n_{1}>0$. As $P$ is a non-simple child, we must have $n_{1}>1$. Let $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ be the element of $\mathcal{R}(I)$ such that $\widetilde{P}_{1}$ is the parent of $P_{1}$. If $P_{1}$ was a simple child, one would be able to write

$$
\begin{equation*}
\left(P_{1}, Q_{1}, n_{1}\right)=\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) *\left(P_{a a^{\prime}}, Q_{a a^{\prime}}, 1\right) \tag{6.15}
\end{equation*}
$$

for some $\left(a, a^{\prime}\right) \in \mathcal{B}$, and then

$$
\begin{equation*}
(P, Q, n)=\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) *\left(P_{a a^{\prime}}, Q_{a a^{\prime}}, 1\right) \tag{6.16}
\end{equation*}
$$

where

$$
\begin{equation*}
\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right)=\left(P_{0}, Q_{0}, n_{0}\right) \quad *\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) . \tag{6.17}
\end{equation*}
$$

Thus $P$ would be a simple child, in contradiction with the assumption of the proposition. Therefore, $P_{1}$ is a non-simple child; by induction on the length, we can write

$$
\begin{equation*}
\left(P_{1}, Q_{1}, n_{1}\right) \in\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \square\left(P_{2}, Q_{2}, n_{2}\right) \tag{6.18}
\end{equation*}
$$

for some $\left(P_{2}, Q_{2}, n_{2}\right) \in \mathcal{R}(I)$. We must have

$$
\begin{equation*}
\widetilde{Q}_{1} \pitchfork_{I} P_{2} . \tag{6.19}
\end{equation*}
$$

We define

$$
\begin{equation*}
\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right):=\left(P_{0}, Q_{0}, n_{0}\right) *\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) . \tag{6.20}
\end{equation*}
$$

We then have $\widetilde{Q}_{0} \subset \widetilde{Q}_{1}$ and hence, from (6.19) and Proposition 1

$$
\begin{equation*}
\widetilde{Q}_{0} \pitchfork_{I} P_{2} \tag{6.21}
\end{equation*}
$$

Thus, the parabolic composition of $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right)$ and $\left(P_{2}, Q_{2}, n_{2}\right)$ is allowed; we obviously have:

$$
\begin{equation*}
(P, Q, n) \in\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) \square\left(P_{2}, Q_{2}, n_{2}\right) . \tag{6.22}
\end{equation*}
$$

2. We have shown so far that it is always possible to write

$$
\begin{equation*}
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(P_{1}, Q_{1}, n_{1}\right) \tag{6.2.2}
\end{equation*}
$$

for some allowed parabolic composition in $\mathcal{R}(I)$. We take $n_{0}$ maximal in (6.23), and assume by contradiction that $n_{0}<\widetilde{n}$. Let $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ be the element of $\mathcal{R}(I)$ such that $\widetilde{P}_{1}$ is the parent of $P_{1}$; let $\left(P_{0}^{\prime}, Q_{0}^{\prime}, n_{0}^{\prime}\right)$ be the element of $\mathcal{R}(I)$ such that $P_{0}^{\prime}$ is the child of $P_{0}$ containing $P$. As $n_{0}<\widetilde{n}$, we have $n_{0}^{\prime}<n$. As $g_{t}^{n_{0}}(P) \subset L_{u}$ by (6.23), $P_{0}^{\prime}$ must be a non-simple child. Then, from the induction hypothesis, we can write

$$
\begin{equation*}
\left(P_{0}^{\prime}, Q_{0}^{\prime}, n_{0}^{\prime}\right) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) \tag{6.24}
\end{equation*}
$$

for some $\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) \in \mathcal{R}(I)$ with $P_{1}^{\prime} \supset P_{1}$. We have thus $P_{1}^{\prime} \supset \widetilde{P}_{1}$. As $Q_{0}$ is $I$-transverse to $P_{1}^{\prime}$, it is also $I$-transverse to $\widetilde{P}_{1}$ (Proposition 2), and we can define ( $\left.\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) \in \mathcal{R}(I)$ by

$$
\begin{equation*}
\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right), \tag{6.25}
\end{equation*}
$$

and $\widetilde{P}_{0} \supset P$. If $P_{1}$ was a simple child of $\widetilde{P}_{1}, P$ would be a simple child of $\widetilde{P}_{0}$. Therefore, by the induction hypothesis, we can write

$$
\begin{equation*}
\left(P_{1}, Q_{1}, n_{1}\right) \in\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \square\left(P_{2}, Q_{2}, n_{2}\right) . \tag{6.26}
\end{equation*}
$$

for some ( $P_{2}, Q_{2}, n_{2}$ ) with $\widetilde{Q}_{1} \pitchfork_{I} P_{2}$. But then $\widetilde{Q}_{0} \pitchfork_{I} P_{2}$ (Proposition 2) and we have

$$
\begin{equation*}
(P, Q, n) \in\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) \square\left(P_{2}, Q_{2}, n_{2}\right) \tag{6.27}
\end{equation*}
$$

which contradicts the maximality of $n_{0}$. We have proven the first part of the proposition.
3. Assume by contradiction that the parent $\widetilde{P}_{1}$ of $P_{1}$ is $I$-transverse. As $Q_{0} \pitchfork_{I} P_{1}$, we have $\widetilde{P}_{1} \subset P_{s}$. There exists, therefore, an $I$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $Q_{u}$ such that $Q_{\alpha}$ and $\widetilde{P}_{1}$ are never critically related. By definition of an $I$-decomposition, there exists $\alpha$ such that $g^{\tilde{n}}(P) \cap Q_{\alpha} \neq \emptyset$; then $Q_{\alpha}$ and $\widetilde{P}_{1}$ are not $I$-separated and they must be $I$-transverse.

As $\widetilde{Q} \cap Q_{\alpha} \neq \emptyset$, we have either $\widetilde{Q} \subset Q_{\alpha}$, and $\widetilde{Q} \pitchfork_{I} \widetilde{P}_{1}$ by Proposition 2, or $\widetilde{Q} \supset Q_{\alpha}$, and $\widetilde{Q} \pitchfork_{I} \widetilde{P}_{1}$ again by concavity (Proposition 4), because both $\widetilde{Q}_{\alpha} \pitchfork_{I} \widetilde{P}_{1}$ and $\widetilde{Q} \pitchfork_{I} P_{1}$ hold. Thus, we can define ( $\widehat{P}, \widehat{Q}, \widehat{n}$ ) by

$$
\begin{equation*}
(\widehat{P}, \widehat{Q}, \widehat{n}) \in(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \square\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right), \tag{6.28}
\end{equation*}
$$

and $P \subset \widehat{P}$. We then have $P \subset \widehat{P} \subset \widetilde{P}, \widehat{P} \neq P, \widehat{P} \neq \widetilde{P}$; thus, $P$ would not be a child of $\widetilde{P}$.

### 6.5 Coherence and Parametric Concavity

The coherence property, asserted in the next proposition, means that larger rectangles are constructed before thinner ones.

Proposition 6. Let $I \subset \widetilde{I}$ be parameter intervals, and let $(P, Q, n) \in \mathcal{R}(I),(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}(\widetilde{I})$. If $\widetilde{P} \subset P$, then $(P, Q, n) \in \mathcal{R}(\widetilde{I})$.

The next property, called parametric concavity, is another partial converse to Proposition 2; it is formally very similar to Proposition 4.

Proposition 7. Let $I \subset I^{\prime}$ be parameter intervals and let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{0}^{\prime}, Q_{0}^{\prime}, n_{0}^{\prime}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}\left(I^{\prime}\right)$ such that $Q_{0} \subset Q_{0}^{\prime} \subset Q_{u}$ and $P_{1} \subset P_{s}$. If both $Q_{0} \pitchfork_{I^{\prime}} P_{1}$ and $Q_{0}^{\prime} \pitchfork_{I} P_{1}$ hold, then $Q_{0}^{\prime} \pitchfork_{I^{\prime}} P_{1}$ also holds.

Obviously there is a similar statement exchanging $P^{\prime} \mathrm{s}$ and $Q^{\prime}$ s. In the proof of Proposition 6 , we will use the following result, which is of independent interest.

Proposition 8. Let $I$ be a parameter interval, and let $(P, Q, n),\left(P^{\prime}, Q^{\prime}, n^{\prime}\right),\left(P^{\prime \prime}, Q^{\prime \prime}, n^{\prime \prime}\right)$ be elements $\mathcal{R}(I)$ with $P \subset P^{\prime} \subset P^{\prime \prime}, P^{\prime} \neq P^{\prime \prime}$. We have:
a) If $(P, Q, n)=\left(P^{\prime \prime}, Q^{\prime \prime}, n^{\prime \prime}\right) *\left(P_{1}, Q_{1}, n_{1}\right)$ for some $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$, then

$$
\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)=\left(P^{\prime \prime}, Q^{\prime \prime}, n^{\prime \prime}\right) *\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)
$$

for some $\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) \in \mathcal{R}(I)$.
b) If $(P, Q, n) \in\left(P^{\prime \prime}, Q^{\prime \prime}, n^{\prime \prime}\right) \square\left(P_{1}, Q_{1}, n_{1}\right)$ for some $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$, then

$$
\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in\left(P^{\prime \prime}, Q^{\prime \prime}, n^{\prime \prime}\right) \square\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)
$$

for some $\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) \in \mathcal{R}(I)$.

Proof of Proposition 8. We only consider case b), case a) being similar but easier. It is sufficient to consider the case where $P^{\prime}$ is the parent of $P$. Let $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ be the element of $\mathcal{R}(I)$ such that $\widetilde{P}_{1}$ is the parent of $P_{1}$. We claim that $Q^{\prime \prime}$ and $\widetilde{P}_{1}$ are $I$-transverse.

Indeed, let $(\bar{P}, \bar{Q}, \bar{n}) \in \mathcal{R}(I)$ such that $\bar{P}$ is the child of $P^{\prime \prime}$ containing $P^{\prime}$. As $g^{n^{\prime \prime}}(P) \subset L_{u}, \bar{P}$ is a non-simple child; by Proposition 5, there exists ( $\bar{P}_{1}, \bar{Q}_{1}, \bar{n}_{1}$ ) such that

$$
\begin{equation*}
(\bar{P}, \bar{Q}, \bar{n}) \in\left(P^{\prime \prime}, Q^{\prime \prime}, n^{\prime \prime}\right) \square\left(\bar{P}_{1}, \bar{Q}_{1}, \bar{n}_{1}\right) \tag{6.29}
\end{equation*}
$$

We then have $Q^{\prime \prime} \pitchfork_{I} \bar{P}_{1}$ and $\bar{P}_{1} \supset \widetilde{P}_{1}$; the claim then follows from Proposition 2.
We can therefore define $(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}(I)$ by

$$
\begin{equation*}
(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in\left(P^{\prime \prime}, Q^{\prime \prime}, n^{\prime \prime}\right) \square\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \tag{6.30}
\end{equation*}
$$

and $P \subset \widetilde{P}$. Let us show that $\widetilde{P}=P^{\prime}$. Otherwise, we have $P^{\prime} \subset \widetilde{P}$, and $P^{\prime} \neq \widetilde{P}$. Let $(\widehat{P}, \widehat{Q}, \widehat{n})$ be the element of $\mathcal{R}(I)$ such that $\widehat{P}$ is the child of $\widetilde{P}$ containing $P^{\prime}$.

If $P_{1}$ was a simple child of $\widetilde{P}_{1}$, we would have $\widetilde{n}_{1}=n_{1}-1$ and $\widetilde{n}=n-1$, forcing $\widetilde{P}=P^{\prime}$. Hence, $P_{1}$ is a non-simple child; this implies that $g^{\widetilde{n}}(P) \subset L_{u}$ and that $\widehat{P}$ is also a non-simple child. By Proposition 5, there exist $\left(P_{2}, Q_{2}, n_{2}\right),\left(\widehat{P}_{2}, \widehat{Q}_{2}, \widehat{n}_{2}\right)$ in $\mathcal{R}(I)$ such that $P_{2} \subset \widehat{P}_{2}$ and

$$
\begin{align*}
\left(P_{1}, Q_{1}, n_{1}\right) & \in\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \square\left(P_{2}, Q_{2}, n_{2}\right),  \tag{6.31}\\
(\widehat{P}, \widehat{Q}, \widehat{n}) & \in(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \square\left(\widehat{P}_{2}, \widehat{Q}_{2}, \widehat{n}_{2}\right) . \tag{6.32}
\end{align*}
$$

Then both $\widetilde{Q}_{1} \pitchfork_{I} P_{2}$ and $\widetilde{Q} \pitchfork_{I} \widehat{P}_{2}$ hold. By concavity (Proposition 4), we also have $\widetilde{Q}_{1} \pitchfork_{I} \widehat{P}_{2}$. We then define ( $\widehat{P}_{1}, \widehat{Q}_{1}, \widehat{n}_{1}$ ) by:

$$
\begin{equation*}
\left(\widehat{P}_{1}, \widehat{Q}_{1}, \widehat{n}_{1}\right) \in\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \square\left(\widehat{P}_{2}, \widehat{Q}_{2}, \widehat{n}_{2}\right) \tag{6.33}
\end{equation*}
$$

and $P_{1} \subset \widehat{P}_{1}$. Then $P_{1} \subset \widehat{P}_{1} \subset \widetilde{P}_{1}$ and $\widehat{P}_{1}$ is distinct from $P_{1}$ and $\widetilde{P}_{1}$, which contradicts that $\widetilde{P}_{1}$ is the parent of $P_{1}$.

Proof of Proposition 6. It is sufficient to consider the case where $P$ is the parent of $\widetilde{P}$ in $\mathcal{R}(I)$. Let $\left(P_{0}, Q_{0}, n_{0}\right)$ be the element of $\mathcal{R}(\widetilde{I})$ such that $P_{0}$ is the parent of $\widetilde{P}$ in $\mathcal{R}(\widetilde{I})$. We want to show that $P_{0}=P$. This is clear if $\tilde{P}$ is a simple child of $P_{0}$. We assume therefore that $\tilde{P}$ is a non-simple child of $P_{0}$. We know that $P \subset P_{0}$. By Proposition 5 , there exists $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}(\widetilde{I})$ such that

$$
\begin{equation*}
(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \tag{6.34}
\end{equation*}
$$

with $Q_{0} \Pi_{\widetilde{I}} \widetilde{P}_{1}$. By Proposition 8 , there exists $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$ such that

$$
\begin{equation*}
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(P_{1}, Q_{1}, n_{1}\right) \tag{6.35}
\end{equation*}
$$

with $Q_{0} \pitchfork_{I} P_{1}$. By induction on the length, as $\widetilde{P}_{1} \subset P_{1}$ we must have $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(\widetilde{I})$. By parameter concavity (Proposition 7 ), as both $Q_{0} \pitchfork_{\widetilde{I}} \widetilde{P}_{1}$ and $Q_{0} \pitchfork_{I} P_{1}$ hold, $Q_{0} \pitchfork_{\tilde{I}} P_{1}$ must also hold, which implies $(P, Q, n) \in \mathcal{R}(\widetilde{I})$.

Proof of Proposition 7. By definition of the transversality relation, there exist parameter intervals $\widetilde{I} \supset I, \widetilde{I}^{\prime} \supset I^{\prime}$ and elements $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}\left(\widetilde{I}^{\prime}\right),\left(\widetilde{P}_{0}^{\prime}, \widetilde{Q}_{0}^{\prime}, \widetilde{n}_{0}^{\prime}\right),\left(\widetilde{P}_{1}^{\prime}, \widetilde{Q}_{1}^{\prime}, \widetilde{n}_{1}^{\prime}\right) \in \mathcal{R}(\widetilde{I})$ such that $Q_{0} \subset \widetilde{Q}_{0}, Q_{0}^{\prime} \subset \widetilde{Q}_{0}^{\prime}, P_{1} \subset \widetilde{P}_{1}, P_{1} \subset \widetilde{P}_{1}^{\prime}$ and

$$
\begin{gather*}
\widetilde{Q}_{0} \Pi_{\tilde{I}^{\prime}} \widetilde{P}_{1},  \tag{6.36}\\
\widetilde{Q}_{0}^{\prime} \Pi_{\tilde{I}} \widetilde{P}_{1}^{\prime} \tag{6.37}
\end{gather*}
$$

both hold. If either $Q_{0}^{\prime} \subset \widetilde{Q}_{0}$ or $I^{\prime} \subset \widetilde{I}$, we conclude immediately that $Q_{0}^{\prime} \pitchfork_{I^{\prime}} P_{1}$ holds. Assume therefore that $\widetilde{Q}_{0} \subset Q_{0}^{\prime}$ and $\widetilde{I} \subset I^{\prime}$. Let $P_{1}^{*}$ be the largest of $\widetilde{P}_{1}, \widetilde{P}_{1}^{\prime}$.

If $\widetilde{P}_{1} \subset \widetilde{P}_{1}^{\prime},\left(\widetilde{P}_{1}^{\prime}, \widetilde{Q}_{1}^{\prime}, \widetilde{n}_{1}^{\prime}\right) \in \mathcal{R}\left(\widetilde{I}^{\prime}\right)$ by coherence (Proposition 6$)$, hence ( $P_{1}^{*}, Q_{1}^{*}, n_{1}^{*}$ ) always belong to $\mathcal{R}\left(\widetilde{I^{\prime}}\right)$. We will show that

$$
\begin{equation*}
\widetilde{Q}_{0}^{\prime} \bar{\Pi}_{\widetilde{I}^{\prime}} P_{1}^{*} \tag{6.38}
\end{equation*}
$$

holds, which implies $Q_{0}^{\prime} \pitchfork_{I^{\prime}} P_{1}$.
We check properties (T1)-(T3) of Subsection 5.4. For all $t \in \widetilde{I}^{\prime}$, we have by (6.36) and (6.6)

$$
\begin{equation*}
\delta_{L R}\left(\widetilde{Q}_{0}^{\prime}, P_{1}^{*}\right) \geqslant \delta_{L R}\left(\widetilde{Q}_{0}, \widetilde{P}_{1}\right) \geqslant 2\left|\widetilde{I}^{\prime}\right| \tag{6.39}
\end{equation*}
$$

By (6.37), there exists $t_{0} \in \widetilde{I} \subset I^{\prime} \subset \widetilde{I}^{\prime}$ such that

$$
\begin{equation*}
\delta_{R}\left(\widetilde{Q}_{0}^{\prime}, \widetilde{P}_{1}^{\prime}\right) \geqslant 2\left|\widetilde{Q}_{0}^{\prime}\right|^{1-\eta} . \tag{6.40}
\end{equation*}
$$

Then, by (6.5), for the same $t_{0}$, we have

$$
\begin{equation*}
\delta_{R}\left(\widetilde{Q}_{0}^{\prime}, P_{1}^{*}\right) \geqslant \delta_{R}\left(\widetilde{Q}_{0}^{\prime}, \widetilde{P}_{1}^{\prime}\right) \geqslant 2\left|\widetilde{Q}_{0}^{\prime}\right|^{1-\eta} \tag{6.41}
\end{equation*}
$$

When $P_{1}^{*}=\widetilde{P}_{1}^{\prime}$, it follows directly from (6.37) that we have

$$
\begin{equation*}
\delta_{L}\left(\widetilde{Q}_{0}^{\prime}, P_{1}^{*}\right) \geqslant 2\left|P_{1}^{*}\right|^{1-\eta} \tag{6.42}
\end{equation*}
$$

for some $t_{1} \in \widetilde{I} \subset \widetilde{I^{\prime}}$.
When $P_{1}^{*}=\widetilde{P}_{1}$, we use (6.36) and (6.4) to find $t_{1} \in \widetilde{I}^{\prime}$ such that

$$
\begin{equation*}
\delta_{L}\left(\widetilde{Q}_{0}^{\prime}, P_{1}^{*}\right) \geqslant \delta_{L}\left(\widetilde{Q}_{0}, \widetilde{P}_{1}\right) \geqslant 2\left|P_{1}^{*}\right|^{1-\eta} \tag{6.43}
\end{equation*}
$$

We have thus proved (6.38).

## Remark.

1. The coherence property means in particular that the parent-child relation does not depend on the parameter interval (once both parent and child are defined).
2. We have presented together Propositions 6, 7 and 8 because the proofs are interconnected. But, actually, coherence (Proposition 6) was already used in the proof of Proposition 4. Logically speaking, all properties in Sections 5-7 should be proved together (as was already mentioned in Section 5).

### 6.6 Further Criteria for Transversality

In this subsection, we give other sufficient conditions for transversality that can be seen as partial converses to Proposition 2.

Proposition 9. Let $I$ be a parameter interval and let $(P, Q, n),\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}(I)$ such that $Q \subset Q_{u}$ and $P_{0} \subset P_{1} \subset P_{s}$. Assume that $Q \pitchfork_{I} P_{0}$ holds and that $2\left|P_{1}\right|^{1-\eta} \leqslant|I|$ for some $t_{1} \in I$. Then $Q$ and $P_{1}$ are also $I$-transverse.

Proof. By definition of the transversality relation, there exist $\widetilde{I} \supset I,(\widetilde{P}, \widetilde{Q}, \widetilde{n}),\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) \in \mathcal{R}(\widetilde{I})$ such that $Q \subset \widetilde{Q}, P_{0} \subset \widetilde{P}_{0}$ and $\widetilde{Q} \bar{\Pi}_{\widetilde{I}} \widetilde{P}_{0}$.

If $P_{1} \subset \widetilde{P}_{0}$ this already implies that $Q \pitchfork_{I} P_{1}$. Let us assume that $\widetilde{P}_{0} \subset P_{1}$. We will show that $\widetilde{Q} \overline{\widetilde{T}}_{\tilde{I}} P_{1}$ holds. By coherence (Proposition 4), we have ( $\left.P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(\widetilde{I})$. Let us check (T1)-(T3).

By (T1) for $\widetilde{Q}, \widetilde{P}_{0}$ and (6.6), we have, for all $t \in \widetilde{I}$ :

$$
\begin{equation*}
\delta_{L R}\left(\widetilde{Q}, P_{1}\right) \geqslant \delta_{L R}\left(\widetilde{Q}, \widetilde{P}_{0}\right) \geqslant 2|\widetilde{I}| . \tag{6.44}
\end{equation*}
$$

By (T2) for $\widetilde{Q}, \widetilde{P}_{0}$ and (6.5), there exists $t_{0} \in \widetilde{I}$ such that

$$
\begin{equation*}
\delta_{R}\left(\widetilde{Q}, P_{1}\right) \geqslant \delta_{R}\left(\widetilde{Q}, \widetilde{P_{0}}\right) \geqslant 2|\widetilde{Q}|^{1-\eta} . \tag{6.45}
\end{equation*}
$$

Finally, we have, for all $t \in I$, by (5.14)

$$
\begin{align*}
\delta_{L}\left(\widetilde{Q}, P_{1}\right) & \geqslant \delta_{L R}\left(\widetilde{Q}, P_{1}\right)-C\left|P_{1}\right|  \tag{6.46}\\
& \geqslant 2|I|-C\left|P_{1}\right| .
\end{align*}
$$

But, for $t=t_{1}$, we have, if $\varepsilon_{0}$ is small enough

$$
\begin{equation*}
2|I|-C\left|P_{1}\right| \geqslant 4\left|P_{1}\right|^{1-\eta}-C\left|P_{1}\right| \geqslant 2\left|P_{1}\right|^{1-\eta} \tag{6.47}
\end{equation*}
$$

Proposition 10. Let $I$ be a parameter interval and let $(P, Q, n),\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}(I)$ such that $Q \subset Q_{u}, P_{0} \subset P_{1} \subset P_{s}$. Assume that $Q \pitchfork_{I} P_{0}$ holds and that $\left|P_{1}\right| \leqslant \frac{1}{2}|Q|$ for all $t \in I$. Then $Q$ and $P_{1}$ are also I-transverse.

Proof. The argument is the same as in Proposition 9, with a slight difference to check (T3) for $\widetilde{Q}$, $P_{1}$. We use (5.13), (5.14) to obtain, for the value $t_{0}$ of the parameter given by (T2) for $\widetilde{Q}, \widetilde{P}_{0}$ :

$$
\begin{align*}
\delta_{L}\left(\widetilde{Q}, P_{1}\right) & \geqslant \delta_{R}\left(\widetilde{Q}, P_{1}\right)-C\left|P_{1}\right|  \tag{6.48}\\
& \geqslant 2|Q|^{1-\eta}-C\left|P_{1}\right| \\
& \geqslant 2\left(2\left|P_{1}\right|\right)^{1-\eta}-C\left|P_{1}\right| \\
& \geqslant 2\left|P_{1}\right|^{1-\eta}
\end{align*}
$$

if $\varepsilon_{0}$ is small enough.
Proposition 11. Let $I \subset I^{\prime}$ be parameter intervals and let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}\left(I^{\prime}\right)$ such that $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$. Assume that $Q_{0} \pitchfork_{I} P_{1}$ holds and that we have $2\left|I^{\prime}\right|<\left|P_{1}\right|^{1-\eta}$, for all $t \in I$. Then $Q_{0}$ and $P_{1}$ are also $I^{\prime}$-transverse.

Proof. By definition of the transversality relation, there exists $\widetilde{I} \supset I,\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in$ $\mathcal{R}(\widetilde{I})$ such that $\widetilde{Q}_{0} \supset Q_{0}, \widetilde{P}_{1} \supset P_{1}$ and $\widetilde{Q}_{0} \bar{\hbar}_{\widetilde{I}} \widetilde{P}_{1}$ holds. If $\widetilde{I} \supset I^{\prime}$, this already implies that $Q_{0} \pitchfork_{I^{\prime}} P_{1}$ holds. Assume thus that $\widetilde{I} \subset I^{\prime}$. We show that $\widetilde{Q}_{0} \bar{\Pi}_{I^{\prime}} \widetilde{P}_{1}$.

Conditions (T2), (T3) in $I^{\prime}$ follow from the same conditions in $\widetilde{I}$. For the value $t_{1}$ given by (T3), we have

$$
\begin{equation*}
\delta_{L R}\left(\widetilde{Q}_{0}, \widetilde{P}_{1}\right) \geqslant \delta_{L}\left(\widetilde{Q}_{0}, \widetilde{P}_{1}\right) \geqslant 2\left|P_{1}\right|^{1-\eta} \geqslant 4\left|I^{\prime}\right| \tag{6.49}
\end{equation*}
$$

By Corollary 11 in Subsection 7.6, this implies

$$
\begin{equation*}
\delta_{L R}\left(\widetilde{Q}_{0}, \widetilde{P}_{1}\right) \geqslant 2\left|I^{\prime}\right|, \quad \forall t \in I^{\prime} \tag{6.50}
\end{equation*}
$$

which is (T1).

### 6.7 A Structure Theorem for New Rectangles

6.7.1 Associativity of Parabolic Composition. Let $I$ be a parameter interval, and let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right),\left(P_{2}, Q_{2}, n_{2}\right)$ be elements in $\mathcal{R}(I)$ such that $Q_{0} \subset Q_{u}, Q_{1} \subset Q_{u}, P_{1} \subset P_{s}$, $P_{2} \subset P_{s}$. We assume that both $Q_{0} \pitchfork_{I} P_{1}$ and $Q_{1} \pitchfork_{I} P_{2}$ hold.

Parabolic composition of $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ produces two elements $\left(P_{01}^{+}, Q_{01}^{+}, n_{01}^{+}\right),\left(P_{01}^{-}, Q_{01}^{-}, n_{01}^{-}\right)$.

As $Q_{01}^{+}$and $Q_{01}^{-}$are contained in $Q_{1}$, it follows from Prop. 2 that both $Q_{01}^{+} \pitchfork_{I} P_{2}$ and $Q_{01}^{-} \pitchfork_{I} P_{2}$ hold.
In the same way, parabolic composition of $\left(P_{1}, Q_{1}, n_{1}\right),\left(P_{2}, Q_{2}, n_{2}\right)$ produces two elements ( $P_{12}^{+}$, $\left.Q_{12}^{+}, n_{12}^{+}\right),\left(P_{12}^{-}, Q_{12}^{-}, n_{12}^{-}\right)$such that both $Q_{0} \pitchfork_{I} P_{12}^{+}$and $Q_{0} \pitchfork_{I} P_{12}^{-}$hold.

It is clear that the four elements of $\mathcal{R}(I)$ obtained by parabolic composition of $\left(P_{01}^{+}, Q_{01}^{+}, n_{01}^{+}\right)$or $\left(P_{01}^{-}, Q_{01}^{-}, n_{01}^{-}\right)$with $\left(P_{2}, Q_{2}, n_{2}\right)$ are the same as the four elements obtained by the parabolic composition of $\left(P_{0}, Q_{0}, n_{0}\right)$ with $\left(P_{12}^{+}, Q_{12}^{+}, n_{12}^{+}\right)$or $\left(P_{12}^{-}, Q_{12}^{-}, n_{12}^{-}\right)$. Their domains are the components of $P_{0} \cap\left(G_{t} \circ g_{t}^{n_{0}}\right)^{-1} P_{1} \cap\left(G_{t} \circ g_{t}^{n_{1}} \circ G_{t} \circ g_{t}^{n_{0}}\right)^{-1} P_{2}$. If $(P, Q, n)$ is any of these four elements, we will write

$$
\begin{equation*}
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(P_{1}, Q_{1}, n_{1}\right) \square\left(P_{2}, Q_{2}, n_{2}\right) \tag{6.51}
\end{equation*}
$$

The same considerations extend immediately, by induction on $k$, to the case of elements $\left(P_{0}, Q_{0}, n_{0}\right)$, $\left(P_{k}, Q_{k}, n_{k}\right)$ such that $P_{i} \subset P_{s}$ for $0<i<k, Q_{i} \subset Q_{u}$ for $0 \leq i<k$, and $Q_{i} \pitchfork_{I} P_{i+1}$ holds for $0 \leq i<k$. Then the successive parabolic compositions of $\left(P_{0}, Q_{0}, n_{0}\right), \cdots,\left(P_{k}, Q_{k}, n_{k}\right)$ produce $2^{k}$ elements and we will write for any such element $(P, Q, n)$ :

$$
\begin{equation*}
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{k}, Q_{k}, n_{k}\right) \tag{6.52}
\end{equation*}
$$

6.7.2 Statement of the Structure Theorem. We have seen in Subsection 5.5 that parabolic composition is never allowed in the class $\mathcal{R}\left(I_{0}\right)$ associated to the starting interval $I_{0}=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]$. This class consists exactly of the affine-like iterates associated to the Markov partition of the initial horseshoe $K_{g_{t}}$.

On the other hand, for elements $(P, Q, n)$ belonging to some class $\mathcal{R}(I)$ but which are not (restrictions of) an element of $\mathcal{R}\left(I_{0}\right)$, parabolic composition must occur. The following theorem gives some rather precise information on this process.

Theorem 1. Let $I$, be a parameter interval of level $>0, \widetilde{I}$ be the parent interval, and let $(P, Q, n)$ be an element of $\mathcal{R}(I)$ which is not (the restriction of) an element of $\mathcal{R}(\widetilde{I})$. Then there exists $k>0$, elements $\left(P_{0}, Q_{0}, n_{0}\right), \cdots,\left(P_{k}, Q_{k}, n_{k}\right)$ of $\mathcal{R}(\widetilde{I})$ such that $Q_{i} \subset Q_{u}$ for $0 \leq i<k, P_{i} \subset P_{s}$ for $0<i \leq k, Q_{i} \pitchfork_{I} P_{i+1}$ holds for $0 \leq i<k$, $Q_{i} \pitchfork_{\widetilde{I}} P_{i+1}$ does not hold for $0 \leq i<k$ and

$$
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{k}, Q_{k}, n_{k}\right)
$$

Moreover, these elements are uniquely determined by these conditions, $P_{i}$ is $\widetilde{I}$-critical for $0<i \leq k$ and $Q_{i}$ is $\widetilde{I}$-critical for $0 \leq i<k$.
6.7.3 We will first introduce a concept, relative to an element $(P, Q, n)$ as in the theorem above, that leads to the determination of the $\left(P_{i}, Q_{i}, n_{i}\right)$.

Let $m, p$ be integers such that $0 \leq m \leq p \leq n$. We say that $[m, p]$ is an $\widetilde{I}$-interval if there exists $(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}(\widetilde{I})$ such that

$$
g_{t}^{m}(P) \subset \widetilde{P} \text { for all } t \in I \text { and } \widetilde{n}=p-m
$$

Lemma 1. The union of two $\widetilde{I}$-intervals with non empty intersection is an $\widetilde{I}$-interval.

Proof. Let $[m, p],\left[m^{\prime}, p^{\prime}\right]$ be these $\widetilde{I}$-intervals, and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n}),\left(\widetilde{P}^{\prime}, \widetilde{Q}^{\prime}, \widetilde{n}^{\prime}\right)$ be the corresponding elements of $\mathcal{R}(\widetilde{I})$. Without loss of generality, we may assume that $m<m^{\prime} \leq p<p^{\prime}$. Replacing if necessary $\widetilde{P}$ by a larger rectangle, we also assume that the element $(\widehat{P}, \widehat{Q}, \widehat{n})$ of $\mathcal{R}(\widetilde{I})$ such that $\widehat{P}$ is the parent of $\widetilde{P}$ satisfies $m+\widehat{n}<m^{\prime}$. There are now two cases:
a) $p=m^{\prime}$.

Let $R_{a}$ be the rectangle containing $\widetilde{Q}$. Then $R_{a} \supset \widetilde{Q}=g^{\widetilde{n}}(\widetilde{P}) \supset g^{p}(P)=g^{m^{\prime}}(P)$; thus $\widetilde{P}^{\prime}$ is also contained in $R_{a}$ and the simple composition

$$
\begin{equation*}
\left(\widetilde{P}^{\prime \prime}, \widetilde{Q}^{\prime \prime}, \widetilde{n}^{\prime \prime}\right):=(\widetilde{P}, \widetilde{Q}, \widetilde{n}) *\left(\widetilde{P}^{\prime}, \widetilde{Q}^{\prime}, \widetilde{n}^{\prime}\right) \tag{6.53}
\end{equation*}
$$

is defined. We have $m+\widetilde{n}^{\prime \prime}=p^{\prime}$ and $g_{t}^{m}(P) \subset P^{\prime \prime}$.
b) $p>m^{\prime}$.

Then, $\widetilde{P}$ is not a simple child of $\widehat{P}$, because otherwise we would have $\widehat{n}=\widetilde{n}-1 \geq m^{\prime}-m$. By Proposition 5, there exists ( $\left.\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right)$ in $\mathcal{R}(\widetilde{I})$ such that

$$
\begin{equation*}
(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in(\widehat{P}, \widehat{Q}, \widehat{n}) \square\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) . \tag{6.54}
\end{equation*}
$$

The element $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right)$ of $\mathcal{R}(\widetilde{I})$ is associated to the $\widetilde{I}$-interval $[\widehat{m}, p]$, where $\widehat{m}=m+\widehat{n}+N_{0}$. We have $m+\widehat{n}<m^{\prime}$ and $g_{t}^{m+\widehat{n}}(P) \subset L_{u}$, hence also $m+\widehat{n}+N_{0}=\widehat{m} \leq m^{\prime}$.

To conclude the proof, we argue by induction on the total length $p^{\prime}-m$ of the interval considered. The case $p^{\prime}-m=0$ is trivial. In the other case, we have the $\widetilde{I}$-intervals $[\widehat{m}, p]$ and $\left[m^{\prime}, p^{\prime}\right]$ with $m<\widehat{m} \leq m^{\prime}$ and hence by induction $\left[\widehat{m}, p^{\prime}\right]$ is an $\widetilde{I}$-interval. Let ( $\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}$ ) be the corresponding element of $\mathcal{R}(\widetilde{I})$; we have $g_{t}^{\widehat{m}}(P) \subset \widetilde{P}_{1} \subset \widetilde{P}_{0}$. From (6.54), $\widehat{Q} \pitchfork_{\widetilde{I}} \widetilde{P}_{0}$ holds, hence also does $\widehat{Q} \pitchfork_{\tilde{I}} \widetilde{P}_{1}$ by Proposition 2. Then, the parabolic composition of $(\widehat{P}, \widehat{Q}, \widehat{n})$ and $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ is allowed and defines an element of $\mathcal{R}(\widetilde{I})$ which guarantees that $\left[m, p^{\prime}\right]$ is an $\widetilde{I}$-interval.
6.7.4 We will now show that the $\left(P_{i}, Q_{i}, n_{i}\right)$ in the theorem are uniquely determined by their properties. Indeed, define $m_{0}=0, p_{0}=n_{0}$ and for $i>0$ :

$$
\begin{equation*}
m_{i}=p_{i-1}+N_{0}, \quad p_{i}=m_{i}+n_{i} . \tag{6.55}
\end{equation*}
$$

Lemma 2. The maximal $\widetilde{-}$-intervals are exactly the $\left[m_{i}, p_{i}\right], 0 \leq i \leq k$, with associated elements $\left(P_{i}, Q_{i}, n_{i}\right)$.

Proof. First, the $\left[m_{i}, p_{i}\right]$ are indeed $\widetilde{I}$-intervals with associated elements $\left(P_{i}, Q_{i}, n_{i}\right)$. To complete the proof, it is sufficient to show that no $\widetilde{I}$-interval $[m, p]$ can intersect a gap $\left(p_{i}, m_{i+1}\right)$. Assume by contradiction that there exists such a $[m, p]$ with associated element $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ and minimal $\widetilde{n}=p-m$. As $g^{\ell}(P) \subset g^{\ell-p_{i}}\left(L_{u}\right)$ does not intersect $R$ for $p_{i}<\ell<m_{i+1}$, we must have $m \leqslant p_{i}$ and $m_{i+1} \leqslant p$. By property (R6) of $\mathcal{R}(\widetilde{I})$ (Subsection 5.3) and the minimality of $\widetilde{n}$, there exists $\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ in $\mathcal{R}(\widetilde{I})$ such that

$$
\begin{equation*}
(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) \square\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \tag{6.56}
\end{equation*}
$$

with $\widetilde{n}_{0}=p_{i}-m \leqslant n_{i}, \widetilde{n}_{1}=p-m_{i+1} \leqslant n_{i+1}$. But then, from $\widetilde{Q}_{0} \supset Q_{i}, \widetilde{P}_{1} \supset P_{i+1}$ and $\widetilde{Q}_{0} 巾_{\tilde{I}} \widetilde{P}_{1}$, we deduce from Proposition 2 that $Q_{i} \pitchfork_{\tilde{I}} P_{i+1}$ holds, a contradiction.
6.7.5 Lemma 2 allow us to define $k$ as being the number of maximal $\widetilde{I}$-intervals minus one, and to define the $\left(P_{i}, Q_{i}, n_{i}\right) \in \mathcal{R}(\widetilde{I})$ as the elements of $\mathcal{R}(\widetilde{I})$ associated to the successive maximal $\widetilde{I}$-intervals. Observe that the maximal $\widetilde{I}$-intervals $\left[m_{i}, p_{i}\right],(0 \leqslant i \leqslant k)$ must indeed satisfy $m_{0}=0$, $m_{i+1}=p_{i}+N_{0}$ for $0 \leqslant i<k$ : every $\ell \in[0, n]$ not contained in an $\widetilde{I}$-interval is such that $g_{t}^{\ell-N}(P) \subset L_{u}$ for some $0<N<N_{0}$ and then no $\widetilde{I}$-interval intersects with $\left(\ell-N, \ell-N+N_{0}\right)$, while $\left\{\left[\ell-N, \ell-N+N_{0}\right]\right\}$ are $\widetilde{I}$-intervals. We observe also that $Q_{i} \pitchfork_{\tilde{I}} P_{i+1}$ does not hold because otherwise $\left[m_{i}, p_{i+1}\right]$ would be an $\widetilde{I}$-interval.
6.7.6 Let $0 \leqslant i<k$. Let us assume by induction over $i$ that $P_{j}$ is $\widetilde{I}$-critical for $0<j \leqslant i, Q_{j}$ is $\widetilde{I}$-critical for $0 \leqslant j<i, Q_{j} \pitchfork_{I} P_{j+1}$ holds for $0 \leqslant j<i$ and that we have an element of $\mathcal{R}(I)$ :

$$
\begin{equation*}
\left(P^{(i)}, Q^{(i)}, p_{i}\right) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{i}, Q_{i}, n_{i}\right) \tag{6.57}
\end{equation*}
$$

such that $P \subset P^{(i)}$. The assumption is vacuously true for $i=0$. We will prove it at step $i+1$. For $i=k$, it gives the properties stated in the theorem for the ( $P_{i}, Q_{i}, n_{i}$ ).
6.7.7 We first prove that $Q_{i}$ is $\widetilde{I}$-critical. Assume by contradiction that $Q_{i}$ is $\widetilde{I}$-transverse. Then $P_{i}$ is $\widetilde{I}$-decomposable. Let $\left(\widehat{P}_{i}, \widehat{Q}_{i}, \widehat{n}_{i}\right)$ be an element of $\mathcal{R}(\widetilde{I})$ such that $\widehat{P}_{i}$ is a child of $P_{i}$ intersecting $g^{m_{i}}(P \cap \Lambda)$. On the other hand, let $\left(\widehat{P}^{(i)}, \widehat{Q}^{(i)}, \widehat{n}^{(i)}\right)$ be the element of $\mathcal{R}(I)$ such that $\widehat{P}^{(i)}$ is the child of $P^{(i)}$ containing $P$. We apply Proposition 5 twice. We find ( $\left.\widetilde{P}_{i+1}, \widetilde{Q}_{i+1}, \widetilde{n}_{i+1}\right)$ in $\mathcal{R}(I)$, $\left(\widetilde{P}_{i+1}^{\prime}, \widetilde{Q}_{i+1}^{\prime}, \widetilde{n}_{i+1}^{\prime}\right)$ in $\mathcal{R}(\widetilde{I})$ such that

$$
\begin{gather*}
\left(\widehat{P}_{i}, \widehat{Q}_{i}, \widehat{n}_{i}\right) \in\left(P_{i}, Q_{i}, n_{i}\right) \square\left(\widetilde{P}_{i+1}^{\prime}, \widetilde{Q}_{i+1}^{\prime}, \widetilde{n}_{i+1}^{\prime}\right),  \tag{6.58}\\
\left(\widehat{P}^{(i)}, \widehat{Q}^{(i)}, \widehat{n}^{(i)}\right) \in\left(P^{(i)}, Q^{(i)}, p_{i}\right) \square\left(\widetilde{P}_{i+1}, \widetilde{Q}_{i+1}, \widetilde{n}_{i+1}\right) . \tag{6.59}
\end{gather*}
$$

If we had $\widetilde{n}_{i+1}^{\prime}>\widetilde{n}_{i+1}$, from $Q_{i} \Pi_{\tilde{I}} \widetilde{I}_{i+1}^{\prime}$ and $Q^{(i)} \pitchfork_{I} \widetilde{P}_{i+1}$, we would deduce first by Proposition 2 that $Q_{i} \pitchfork_{I} \widetilde{P}_{i+1}^{\prime}$, then by concavity (Proposition 4) that $Q_{i} \pitchfork_{I} \widetilde{P}_{i+1}$; parabolic composition would yield an
element $\left(\bar{P}_{i}, \bar{Q}_{i}, \bar{n}_{i}\right) \in \mathcal{R}(I)$ with $P_{i} \supsetneqq \bar{P}_{i} \supsetneqq \hat{P}_{i}$, in contradiction with coherence (Proposition 6) and the definition of $\widehat{P}_{i}$.

Therefore, we must have $\widetilde{n}_{i+1}^{\prime} \leqslant \widetilde{n}_{i+1} ;$ as $\widehat{n}^{(i)} \leqslant n$, we have also $m_{i}+\widehat{n}_{i} \leqslant n$, and $\left[m_{i}, m_{i}+\widehat{n}_{i}\right]$ is an $\widehat{I}$-interval strictly larger than $\left[m_{i}, p_{i}\right]$; this contradiction shows that $Q_{i}$ is $\widetilde{I}$-critical.
6.7.8 The proof that $P_{i+1}$ is $\widetilde{I}$-critical is rather similar. We assume by contradiction that it is $\widetilde{I}$ transverse. Then $Q_{i+1}$ is $\widetilde{I}$-decomposable and we can find $\left(P_{i+1}^{*}, Q_{i+1}^{*}, n_{i+1}^{*}\right) \in \mathcal{R}(\widetilde{I})$ such that $Q_{i+1}^{*}$ is a child of $Q_{i+1}$ intersecting $g^{p_{i+1}}(P \cap \Lambda)$. By Proposition 5, there exists $\left(P_{*}^{(i)}, Q_{*}^{(i)}, n_{*}^{(i)}\right) \in \mathcal{R}(\widetilde{I})$ such that

$$
\begin{equation*}
\left(P_{i+1}^{*}, Q_{i+1}^{*}, n_{i+1}^{*}\right) \in\left(P_{*}^{(i)}, Q_{*}^{(i)}, n_{*}^{(i)}\right) \square\left(P_{i+1}, Q_{i+1}, n_{i+1}\right) . \tag{6.60}
\end{equation*}
$$

If we had $n_{*}^{(i)}>p_{i}$, we would derive a contradiction as follows: we should have $Q^{(i)} \pitchfork_{I} P_{i+1}$ either (if $\widetilde{n}_{i+1} \leqslant n_{i+1}$ ) from $Q^{(i)} \pitchfork_{I} \widetilde{P}_{i+1}$ by Proposition 2 or (if $\widetilde{n}_{i+1} \geqslant n_{i+1}$ ) from $Q^{(i)} \pitchfork_{I} \widetilde{P}_{i+1}$ and $Q_{*}^{(i)} \pitchfork_{I} P_{i+1}$ by Proposition 4; then, parabolic composition of $\left(P^{(i)}, Q^{(i)}, p_{i}\right)$ and $\left(P_{i+1}\right.$, $\left.Q_{i+1}, n_{i+1}\right)$ produces an element $\left(P^{(i+1)}, Q^{(i+1)}, p_{i+1}\right)$ in $\mathcal{R}(I)$ with $Q_{i+1} \supsetneqq Q^{(i+1)} \supsetneqq Q_{i+1}^{*}$, which is not compatible with coherence (Proposition 6).

Thus, we must have $n_{*}^{(i)} \leqslant p_{i}$; then we also have $n_{i+1}^{*} \leqslant p_{i+1}$ and $\left[p_{i+1}-n_{i+1}^{*}, p_{i+1}\right]$ is an $\widetilde{I}$-interval strictly larger than $\left[m_{i+1}, p_{i+1}\right]$. This contradiction shows that $P_{i+1}$ is $\widetilde{I}$-critical.
6.7.9 We now prove that $Q^{(i)}$ and $P_{i+1}$ are $I$-transverse. If $\widetilde{n}_{i+1} \leq n_{i+1}$, we have $Q^{(i)} \pitchfork_{I} \widetilde{P}_{i+1}$ by (6.59) and thus also $Q^{(i)} \pitchfork_{I} P_{i+1}$ by Proposition 2. Let us assume that $\widetilde{n}_{i+1}>n_{i+1}$. In this case, we claim that $Q_{i+1}$ is $\widetilde{I}$-critical. Indeed, if it was $\widetilde{I}$-transverse, $P_{i+1}$ would be $\widetilde{I}$-decomposable and we would find an element ( $\widehat{P}_{i+1}, \widehat{Q}_{i+1}, \widehat{n}_{i+1}$ ) of $\mathcal{R}(\widetilde{I})$ such that $\widehat{P}_{i+1}$ is a child of $P_{i+1}$ intersecting $g^{m_{i+1}}(P \cap \Lambda)$. By coherence (Proposition 6), we should have $\widehat{P}_{i+1} \supset \widetilde{P}_{i+1}$ and $\left[m_{i+1}, m_{i+1}+\widehat{n}_{i+1}\right]$ would be an $\widetilde{I}$-interval larger than $\left[m_{i+1}, p_{i+1}\right]$, a contradiction.

As $\left(P_{i+1}, Q_{i+1}, n_{i+1}\right)$ is $\widetilde{I}$-bicritical, and the parent interval $\widetilde{I}$ is always assumed to be $\beta$-regular, we have, for all $t \in \widetilde{I}$

$$
\begin{equation*}
\left|P_{i+1}\right|<|\widetilde{I}|^{\beta} \tag{6.61}
\end{equation*}
$$

and thus also (with $\varepsilon_{0}$ small enough)

$$
\begin{equation*}
2\left|P_{i+1}\right|^{1-\eta}<|I| . \tag{6.62}
\end{equation*}
$$

It now follows from Proposition 9 and $Q^{(i)} \pitchfork_{I} \widetilde{P}_{i+1}$ that $Q^{(i)}$ and $P_{i+1}$ are $I$-transverse.
6.7.10 When $i=0, Q^{(0)}=Q_{0}$ and we have already shown that $Q_{i}$ and $P_{i+1}$ are $I$-transverse.

When $i>0,\left(P_{i}, Q_{i}, n_{i}\right)$ is $\widetilde{I}$-bicritical and, therefore, we have for all $t \in \widetilde{I}$ :

$$
\begin{equation*}
\left|Q_{i}\right|<|\widetilde{I}|^{\beta}, \tag{6.63}
\end{equation*}
$$

and thus also

$$
\begin{equation*}
2\left|Q_{i}\right|^{1-\eta}<|I| . \tag{6.64}
\end{equation*}
$$

It follows from Proposition 9 and $Q^{(i)} \pitchfork_{I} P_{i+1}$ that $Q_{i}$ and $P_{i+1}$ are $I$-transverse.
To conclude the induction step of 6.7.6, we simply observe that the parabolic composition of $\left(P^{(i)}, Q^{(i)}, n^{(i)}\right)$ and $\left(P_{i+1}, Q_{i+1}, n_{i+1}\right)$ is allowed in $\mathcal{R}(I)$; it produces an element $\left(P^{(i+1)}, Q^{(i+1)}, p_{i+1}\right) \in$ $\mathcal{R}(I)$ such that $P^{(i+1)}$ intersects $P$ and therefore contains $P$.

The proof of the theorem is now complete.
6.7.11 In the next two Corollaries, the setting and notations are those of Theorem 1 .

Corollary 6. For all $t \in I$, we have

$$
|P| \leqslant C^{k}\left|P_{0}\right|\left|P_{1}\right| \cdots\left|P_{k}\right||I|^{-\frac{k}{2}}
$$

Proof. We have $P=P^{(k)}$, with $P^{(i)}$ defined in (6.57); we prove that, for all $t \in I$

$$
\begin{equation*}
\left|P^{(i)}\right| \leqslant C^{i}\left|P_{0}\right| \cdots\left|P_{i}\right||I|^{-\frac{i}{2}} . \tag{6.65}
\end{equation*}
$$

As $P^{(0)}=P_{0}$, this is true for $i=0$. The induction step is a consequence of the key estimate (3.27) for parabolic composition if we know that, for all $t \in I$ :

$$
\begin{equation*}
\delta\left(Q^{(i)}, P_{i+1}\right) \geqslant|I| . \tag{6.66}
\end{equation*}
$$

As $Q_{i}$ and $P_{i+1}$ are $I$-transverse, there exists $I^{*} \supset I,\left(P_{i}^{*}, Q_{i}^{*}, n_{i}^{*}\right),\left(P_{i+1}^{*}, Q_{i+1}^{*}, n_{i+1}^{*}\right) \in \mathcal{R}\left(I^{*}\right)$ such that $Q_{i} \subset Q_{i}^{*}, P_{i+1} \subset P_{i+1}^{*}$ and $Q_{i}^{*} \bar{\Phi}_{I^{*}} P_{i+1}^{*}$ holds. From (T1) in Subsection 5.4, we have, for all $t \in I^{*}$

$$
\begin{equation*}
\delta_{L R}\left(Q_{i}^{*}, P_{i+1}^{*}\right) \geqslant 2\left|I^{*}\right| . \tag{6.67}
\end{equation*}
$$

From (6.3), (5.11)-(5.14) and (R7), we get, for all $t \in I$ :

$$
\begin{align*}
\delta\left(Q^{(i)}, P_{i+1}\right) & \geqslant \delta\left(Q_{i}^{*}, P_{i+1}^{*}\right)  \tag{6.68}\\
& \geqslant \delta_{L R}\left(Q_{i}^{*}, P_{i+1}^{*}\right)-c\left(\left|P_{i+1}^{*}\right|+\left|Q_{i}^{*}\right|\right) \\
& \geqslant \frac{1}{2} \delta_{L R}\left(Q_{i}^{*}, P_{i+1}^{*}\right) \\
& \geqslant\left|I^{*}\right| \geqslant|I|,
\end{align*}
$$

which concludes the proof.
Corollary 7. For all $0<i<k$, and $t \in \widetilde{I}$, we have

$$
\left|P_{i}\right|<|\widetilde{I}|^{\beta} .
$$

For all $t \in I$, we also have

$$
\left|P_{k}\right|<c|\widetilde{I}|^{(1-\eta)^{-1}} .
$$

Proof. The first assertion is an immediate consequence of the regularity of the parameter interval $\widetilde{I}$, because $\left(P_{i}, Q_{i}, n_{i}\right)$ is $\widetilde{I}$-bicritical for $0<i<k$.

For the second assertion, we first observe that in the proof of Corollary 6 , we must have $I^{*}=I$ because $Q_{k-1} \pitchfork_{\tilde{I}} P_{k}$ does not hold; for the same reason, there exists $t^{*} \in \widetilde{I}$ such that

$$
\begin{equation*}
\delta_{L R}\left(Q_{k-1}^{*}, P_{k}^{*}\right)<2|\widetilde{I}| . \tag{6.69}
\end{equation*}
$$

But then, from (6.6) and Corollary 11 in Subsection 7.6, we have, for all $t \in \widetilde{I}$ :

$$
\begin{align*}
\delta\left(Q_{k-1}, P_{k}\right) & <\delta_{L R}\left(Q_{k-1}, P_{k}\right)  \tag{6.70}\\
& <\delta_{L R}\left(Q_{k-1}^{*}, P_{k}^{*}\right) \\
& <c|\widetilde{I}| .
\end{align*}
$$

From (R7), we have, for all $t \in I$

$$
\begin{equation*}
\delta\left(Q_{k-1}, P_{k}\right)>c^{-1}\left|P_{k}\right|^{1-\eta} \tag{6.71}
\end{equation*}
$$

and the second assertion of the Corollary follows.

### 6.7.12

Corollary 8. Any $(P, Q, n)$ in $\mathcal{R}(I)$ but not in $\mathcal{R}(\widetilde{I})$ satisfies, for all $t \in I$ :

$$
|P| \leqslant|\widetilde{I}|^{\frac{1}{2}}, \quad|Q| \leqslant|\widetilde{I}|^{\frac{1}{2}}
$$

Proof. From (6.70), (6.71) and (3.27), we actually have, for all $t \in I$

$$
\begin{equation*}
|P| \leqslant\left|P_{0}\right||\widetilde{I}|^{\frac{1}{2}}, \tag{6.72}
\end{equation*}
$$

and we must have $\left|P_{0}\right| \ll 1$ as $Q_{0} \subset Q_{u}$.

### 6.7.13

Corollary 9. Any candidate interval I is $\bar{\beta}$-regular, with $\bar{\beta}=\beta(1+\tau)^{-1}$.

Proof. Let $(P, Q, n)$ be an $I$-bicritical element of $\mathcal{R}(I)$. If ( $P, Q, n$ ) belongs to $\mathcal{R}(\widetilde{I})$, then it is $\widetilde{I}$-bicritical and we have, for all $t \in \widetilde{I}$ :

$$
\begin{equation*}
\max (|P|,|Q|)<|\widetilde{I}|^{\beta}=|I|^{\bar{\beta}} . \tag{6.73}
\end{equation*}
$$

Assume now that $(P, Q, n)$ does not belong to $\mathcal{R}(\widetilde{I})$. We apply Theorem 1. The element $\left(P_{0}, Q_{0}, n_{0}\right) \in$ $\mathcal{R}(\widetilde{I})$ is $\widetilde{I}$-bicritical: $Q_{0}$ is $\widetilde{I}$-critical by Theorem 1 , and $P_{0}$ is $\widetilde{I}$-critical because $P$ is $I$-critical. Applying Corollaries 6 and 7 gives, for all $t \in I$

$$
\begin{align*}
|P| & \leqslant C\left|P_{0}\right|\left|P_{k}\right||I|^{-\frac{1}{2}}  \tag{6.74}\\
& \leqslant C|\widetilde{I}|^{\beta+1-\frac{1}{2}(1+\tau)},
\end{align*}
$$

with $\beta+1-\frac{1}{2}(1+\tau)>\beta+\frac{1}{3}$, so $I$-bicritical elements are much thinner in this case.
Remark. The same phenomenon will be important again in Section 9: "fat" bicritical elements were created much earlier.
6.7.14 The last result in this section is a complement to Proposition 5 in Subsection 6.4.

Let $I$ be a parameter interval, and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n}),(P, Q, n)$ be elements of $\mathcal{R}(I)$ such that $P$ is a non-simple child of $\widetilde{P}$. From Proposition 5, we know that there exists $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$ such that $\widetilde{Q} \pitchfork_{I} P_{1}$ holds and

$$
\begin{equation*}
(P, Q, n) \in(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \square\left(P_{1}, Q_{1}, n_{1}\right) . \tag{6.75}
\end{equation*}
$$

Proposition 12. Assume moreover that $\widetilde{P}$ is $I$-critical. Then, for all $t \in I$ we have

$$
\delta\left(\widetilde{Q}, P_{1}\right) \geqslant|\widetilde{P}|^{\frac{1}{\bar{\beta}}}
$$

Proof. Consider first the case where $\widetilde{Q}$ is $I$-critical. Then $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ is $I$-bicritical. From Corollary 9 , we know that $I$ is $\bar{\beta}$-regular and therefore we have, for all $t \in I$ :

$$
\begin{equation*}
|\widetilde{P}|<|I|^{\bar{\beta}} \tag{6.76}
\end{equation*}
$$

On the other hand, as $\widetilde{Q} \pitchfork_{I} P_{1}$ holds, we can find $\left(\widetilde{P}^{*}, \widetilde{Q}^{*}, \widetilde{n}^{*}\right),\left(P_{1}^{*}, Q_{1}^{*}, n_{1}^{*}\right) \in \mathcal{R}(I)$ with $\widetilde{Q} \subset \widetilde{Q}^{*}$, $P_{1} \subset P_{1}^{*}$ such that, for all $t \in I$

$$
\begin{equation*}
\delta_{L R}\left(\widetilde{Q}^{*}, P_{1}^{*}\right) \geqslant 2|I| \tag{6.77}
\end{equation*}
$$

From (R7) (cf. Subsection 5.4) and (5.11)-(5.14), it now follows that

$$
\begin{align*}
\delta\left(\widetilde{Q}, P_{1}\right) & \geqslant \delta\left(\widetilde{Q}^{*}, P_{1}^{*}\right)  \tag{6.78}\\
& \geqslant \frac{1}{2} \delta_{L R}\left(\widetilde{Q}^{*}, P_{1}^{*}\right) \\
& \geqslant|I| \\
& \geqslant|\widetilde{P}|^{\frac{1}{3}} .
\end{align*}
$$

From now on we assume that $\widetilde{Q}$ is $I$-transverse. Let $I^{*} \supset I$ be the largest parameter interval such that $(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}\left(I^{*}\right)$ and $\widetilde{Q}$ is $I^{*}$-transverse. As the transversality relation never holds for the
starting interval $\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]=I_{0}, I^{*}$ is not equal to $I_{0}$. Let $\widetilde{I}{ }^{*}$ be the parent of $I^{*}$. We claim that for all $t \in I^{*}$, we have:

$$
\begin{equation*}
|\widetilde{P}| \leqslant\left|\widetilde{I^{*}}\right|^{\beta}=\left|I^{*}\right|^{\bar{\beta}} . \tag{6.79}
\end{equation*}
$$

Indeed, from the definition of $I^{*}$, we have that either $(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \notin \mathcal{R}\left(\widetilde{I}^{*}\right)$, or $(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}\left(\widetilde{I}^{*}\right)$ and $\widetilde{Q}$ is $\widetilde{I}^{*}$-critical. In the second case, $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ is $\widetilde{I}^{*}$-bicritical and $\widetilde{I}^{*}$ is $\beta$-regular, which gives (6.79). In the first case, we obtain from (6.74) in the proof of Corollary 9 (where only the $I$-criticality of $P$ was used) that, for all $t \in I^{*}$

$$
\begin{equation*}
|\widetilde{P}| \leqslant C\left|\widetilde{I}^{*}\right|^{\beta+\frac{1}{3}}<\left|I^{*}\right|^{\bar{\beta}} \tag{6.80}
\end{equation*}
$$

The claim is proved.
As $\widetilde{Q}$ is $I^{*}$-transverse, there exists an $\widetilde{I}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $P_{s}$ in $\mathcal{R}\left(I^{*}\right)$ such that, for any $\alpha, \widetilde{Q}$ and $P_{\alpha}$ are either $I^{*}$-separated or $I^{*}$-transverse. There exists $\alpha_{0}$ such that $P_{1}$ and $P_{\alpha_{0}}$ intersect; $\widetilde{Q}$ and $P_{\alpha_{0}}$ must be $I^{*}$-transverse. This implies, as for (6.77) above, that we have, for all $t \in I^{*}:$

$$
\begin{equation*}
\delta\left(\widetilde{Q}, P_{\alpha_{0}}\right) \geqslant \frac{3}{2}\left|I^{*}\right| . \tag{6.81}
\end{equation*}
$$

If $P_{1} \subset P_{\alpha_{0}}$, we have

$$
\begin{equation*}
\delta\left(\widetilde{Q}, P_{1}\right) \geqslant \delta\left(\widetilde{Q}, P_{\alpha_{0}}\right) \tag{6.82}
\end{equation*}
$$

If $P_{\alpha_{0}} \subset P_{1}$, we have from (5.11)-(5.14) and (R7), for all $t \in I$ :

$$
\begin{align*}
\delta\left(\widetilde{Q}, P_{1}\right) & \geqslant \frac{2}{3} \delta_{R}\left(\widetilde{Q}, P_{1}\right)  \tag{6.83}\\
& \geqslant \frac{2}{3} \delta_{R}\left(\widetilde{Q}, P_{\alpha_{0}}\right) \\
& \geqslant \frac{2}{3} \delta\left(\widetilde{Q}, P_{\alpha_{0}}\right) .
\end{align*}
$$

In all cases, combining this with (6.81) and (6.79) gives the required estimate.

## 7 Estimates for the Classes $\mathcal{R}(I)$

### 7.1 A Stretched Exponential Estimate for Widths

The next proposition is a substitute for the uniform exponential estimates for widths that are characteristic of the uniformly hyperbolic dynamics. We denote by $\gamma$ the constant

$$
\begin{equation*}
\gamma:=\frac{\log \frac{3}{2}}{\log 2} \in(0,1) . \tag{7.1}
\end{equation*}
$$

Proposition 13. Let $I$ be a parameter interval and let $(P, Q, n)$ be an element of $\mathcal{R}(I)$. For all $t \in I$, we have

$$
|P| \leqslant C \exp \left(-n^{\gamma}\right)
$$

with the stronger estimate

$$
|P| \leqslant C \exp \left(-2 n^{\gamma}\right)
$$

when the parent of $P$ is I-critical.

Proof. If $I$ is the starting interval $I_{0}=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]$, a stronger exponential bound actually holds. The proof is by induction on $n$, and the estimates are therefore valid when $n=O\left(\log \varepsilon_{0}^{-1}\right)$ (in which case $\left.(P, Q, n) \in \mathcal{R}\left(I_{0}\right)\right)$. Let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be the element of $\mathcal{R}(I)$ such that $\widetilde{P}$ is the parent of $P$. In the case where $P$ is a simple child of $\widetilde{P}$, the bound for $P$ easily follows from the bound for $\widetilde{P}$. Let us therefore assume that $P$ is a non-simple child of $\widetilde{P}$, in which case there exists by Proposition 5 an element $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$ such that $\tilde{Q} \pitchfork_{I} P_{1}$ holds, the parent of $P_{1}$ is $I$-critical and

$$
\begin{equation*}
(P, Q, n) \in(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \square\left(P_{1}, Q_{1}, n_{1}\right) . \tag{7.2}
\end{equation*}
$$

By the induction hypothesis, we have

$$
\begin{align*}
|\widetilde{P}| & \leqslant C \exp \left(-\widetilde{n}^{\gamma}\right)  \tag{7.3}\\
\left|P_{1}\right| & \leqslant C \exp \left(-2 n_{1}^{\gamma}\right) . \tag{7.4}
\end{align*}
$$

From (R7) and (3.27), we have, for all $t \in I$ :

$$
\begin{equation*}
|P| \leqslant C|\widetilde{P}|\left|P_{1}\right|^{\frac{1+\eta}{2}} \ll|\widetilde{P}|\left|P_{1}\right|^{\frac{1}{2}} \tag{7.5}
\end{equation*}
$$

and this gives

$$
\begin{equation*}
|P| \leqslant \exp \left(-\widetilde{n}^{\gamma}-n_{1}^{\gamma}\right) . \tag{7.6}
\end{equation*}
$$

This proves the required estimate in the general case, because, when $\widetilde{n}$ and $n_{1}$ are large, we have

$$
\begin{equation*}
\tilde{n}^{\gamma}+n_{1}^{\gamma} \geqslant\left(\widetilde{n}+n_{1}+N_{0}\right)^{\gamma}=n^{\gamma} . \tag{7.7}
\end{equation*}
$$

Assume now that $\widetilde{P}$ is $I$-critical. Instead of (7.3), we have

$$
\begin{equation*}
|\widetilde{P}| \leqslant C \exp \left(-2 \widetilde{n}^{\gamma}\right) \tag{7.8}
\end{equation*}
$$

and then from (7.5) we obtain

$$
\begin{equation*}
|P| \leqslant \exp \left(-2 \widetilde{n}^{\gamma}-n_{1}^{\gamma}\right) \tag{7.9}
\end{equation*}
$$

This will be useful when $\widetilde{n} \geqslant n_{1}$. When $\widetilde{n} \leqslant n_{1}$, we prefer to rely on Proposition 12 which gives

$$
\begin{equation*}
\delta\left(\widetilde{Q}, P_{1}\right) \geqslant|\widetilde{P}|^{1 / \bar{\beta}} \gg|\widetilde{P}| . \tag{7.10}
\end{equation*}
$$

When we use this in (3.27) and combine with (7.4) and (7.8), we get

$$
\begin{equation*}
|P| \leqslant \exp \left(-\widetilde{n}^{\gamma}-2 n_{1}^{\gamma}\right) \tag{7.11}
\end{equation*}
$$

To get the required estimate from (7.9), (7.11), we have only to observe that the function $u \mapsto$ $u^{\gamma}+2(1-u)^{\gamma}$ is concave on $\left[0, \frac{1}{2}\right]$ and equal to 2 for $u=0$ and $u=\frac{1}{2}$.

### 7.2 Uniform Cone Condition

In this subsection, we will check that all elements $(P, Q, n) \in \mathcal{R}(I)$ satisfy the cone condition (AL2) of Subsection 3.2 for the parameters $\lambda, u_{0}, v_{0}$ of Subsection 5.3: we have $u_{0}=u^{1 / 2}, v_{0}=v^{1 / 2}$, and all $(P, Q, n) \in \mathcal{R}\left(\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]\right)$ satisfy (AL2) with parameters $\lambda, u, v$.

Let $(A, B)$ be the implicit representation of the affine-like iterate $(P, Q, n)$; we have to prove that
(AL2)

$$
\begin{aligned}
& \lambda\left|A_{x}\right|+u_{0}\left|A_{y}\right| \leqslant 1 \\
& \lambda\left|B_{y}\right|+v_{0}\left|B_{x}\right| \leqslant 1
\end{aligned}
$$

Let $u_{1}=u^{\frac{3}{4}}, v_{1}=v^{\frac{3}{4}}$. We will prove that, for all $t \in I$, we have

$$
\begin{equation*}
\left|A_{y}\right|<u_{1}^{-1}, \quad\left|B_{x}\right|<v_{1}^{-1} \tag{7.12}
\end{equation*}
$$

This is sufficient to obtain (AL2): we already know that if $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$ them (AL2) is satisfied; on the other hand, if $(P, Q, n) \notin \mathcal{R}\left(I_{0}\right)$, then, for all $t \in I$, we have from Corollary 8

$$
\begin{equation*}
\left|A_{x}\right|<C \varepsilon_{0}^{\frac{1}{2}}, \quad\left|B_{y}\right|<C \varepsilon_{0}^{\frac{1}{2}} \tag{7.13}
\end{equation*}
$$

with $\varepsilon_{0}$ small enough, (7.13) and (7.12) give (AL2).

Let us now proceed with the proof of (7.12). When $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$, we have the stronger estimate:

$$
\begin{equation*}
\left|A_{y}\right|<u^{-1}, \quad\left|B_{x}\right|<v^{-1} \tag{7.14}
\end{equation*}
$$

Let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be the element of $\mathcal{R}(I)$ such that $\widetilde{P}$ is the parent of $P$. Denote by $(\widetilde{A}, \widetilde{B})$ the implicit representation of the corresponding affine-like iterate.

If $P$ is a simple child, we use formula (3.11) of Subsection 3.3 to obtain

$$
\begin{equation*}
\left|A_{y}-\widetilde{A}_{y}\right| \leqslant C|\widetilde{P}||\widetilde{Q}| \tag{7.15}
\end{equation*}
$$

If $P$ is a non-simple child, it is obtained (Proposition 5) by the parabolic composition of $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ with some $\left(P_{1}, Q_{1}, n_{1}\right)$; we use formula (3.30) of Subsection 3.5 to obtain

$$
\begin{equation*}
\left|A_{y}-\widetilde{A}_{y}\right| \leqslant C|\widetilde{P}||\widetilde{Q}|\left(\delta\left(\widetilde{Q}, P_{1}\right)\right)^{-\frac{1}{2}} \tag{7.16}
\end{equation*}
$$

From (R7), $\delta\left(\widetilde{Q}, P_{1}\right)$ is much larger than $|\widetilde{Q}|$. In all cases, we have

$$
\begin{equation*}
\left|A_{y}-\widetilde{A}_{y}\right| \leqslant C|\widetilde{P}| \leqslant C \exp \left(-\widetilde{n}^{\gamma}\right) \tag{7.17}
\end{equation*}
$$

where we have used Proposition 13 in the last inequality. We only need (7.17) when $\widetilde{n}$ is large (because we already have (7.14) otherwise), and the series $\sum \exp \left(-n^{\gamma}\right)$ is convergent. Therefore (7.12) is a consequence of (7.14) and (7.17).

The proof of (AL2), i.e., the first part of condition (R2) in Subsection 5.3, is now complete.

### 7.3 Bounded Distortion

We now check the second half of property (R2) in Subsection 5.3. We have to prove that, for all $(P, Q, n) \in \mathcal{R}(I)$, we have the following estimate on distortion:

$$
\begin{equation*}
D\left(g_{t}^{n} / P\right) \leqslant 2 D_{0} \tag{7.18}
\end{equation*}
$$

Here, the constant $D_{0}$ corresponds to the stronger estimate we obtain from (MP6) when $(P, Q, n) \in$ $\mathcal{R}\left(I_{0}\right)$ :

$$
\begin{equation*}
D\left(g_{t}^{n} / P\right) \leqslant D_{0} \tag{7.19}
\end{equation*}
$$

For $m>0$, define

$$
\begin{equation*}
D(m)=\sup _{\substack{(P, Q, n) \in \mathcal{R}(I) \\ n \leqslant m}} \sup _{t \in I} D\left(g_{t}^{n} / P\right) \tag{7.20}
\end{equation*}
$$

To obtain, by induction on $m$, a bound for the non-decreasing sequence $D(m)$, we combine (7.19), (which gives $D(m) \leqslant D_{0}$ for $m=O\left(\log \varepsilon_{0}^{-1}\right)$ ), Proposition 13 and the bounds (3.13) in Subsection 3.3 (for simple composition) and (3.29) in Subsection 3.5 (for parabolic composition). We set

$$
\begin{equation*}
D_{s}(m)=\max _{\substack{n>0, n^{\prime}>0 \\ n+n^{\prime} \leqslant m}} D_{s}\left(n, n^{\prime}\right) \tag{7.21}
\end{equation*}
$$

with

$$
\begin{equation*}
D_{s}\left(n, n^{\prime}\right)=D(n)+c \exp \left(-n^{\gamma}\right)\left(D(n)+D\left(n^{\prime}\right)\right) . \tag{7.22}
\end{equation*}
$$

We also set

$$
\begin{equation*}
D_{p}(m)=\max _{\substack{n \gg 0, n^{\prime} \gg 0 \\ n+n^{\prime}+N_{0} \leqslant m}} D_{p}\left(n, n^{\prime}\right) \tag{7.23}
\end{equation*}
$$

with

$$
\begin{equation*}
D_{p}\left(n, n^{\prime}\right)=D(n)+c \exp \left(-\eta n^{\gamma}\right) \tag{7.24}
\end{equation*}
$$

Then, we have, as long as $D(m)$ is not too large (cf. the condition for (3.29) to hold)

$$
\begin{equation*}
D(m) \leqslant \max \left(D_{s}(m), D_{p}(m)\right) \tag{7.25}
\end{equation*}
$$

The reason for formula (7.24) to hold is that the term $c\left|P_{1}\right| \delta^{-1}$ of (3.29) is smaller than $\left|P_{1}\right|^{\eta}$ by (R7); then one uses Proposition 13.

It is now clear that (7.18) follows from (7.19), (7.21)-(7.25).

### 7.4 Estimates for the Special Rectangles $P_{s}$ and $Q_{u}$

In the next subsection, we will check the estimates contained in condition (R4) of Subsection 5.3 concerning the class $\mathcal{R}(I)$.

These estimates, which are related to parabolic composition, are valid for an element $(P, Q, n)$ of $\mathcal{R}(I)$ which satisfies $Q \subset Q_{u}$ (or $P \subset P_{s}$ ).

In the present section, we will be concerned with the affine-like iterates which are directly associated with the elements $\left(P_{s}, Q_{s}, n_{s}\right)$ and $\left(P_{u}, Q_{u}, n_{u}\right)$.

We will make the computations for $\left(P_{s}, Q_{s}, n_{s}\right)$ the other case is obviously symmetric. We will assume that the periodic point $p_{s}$ is fixed: the general case is completely similar, but the notations are more awkward.

In this subsection, we just write $(x, y)$ for the coordinates in the rectangle $R_{a_{s}}$ containing $p_{s}$; we denote by $(A, B)$ the implicit representation of the affine-like iterate

$$
\begin{equation*}
G_{t}:\left(R_{a_{s}}\right) \cap g_{t}^{-1}\left(R_{a_{s}}\right) \longrightarrow g_{t}\left(R_{a_{s}}\right) \cap R_{a_{s}} \tag{7.26}
\end{equation*}
$$

For $n \geqslant 0$, we denote by $\left(A^{(n)}, B^{(n)}\right)$ the implicit representation of the $n^{\text {th }}$ iterate of this restriction.

As the equation of $W_{\text {loc }}^{s}\left(p_{s}\right)$ is $\{x=0\}$ (cf. (MP3) in Subsection 2.2), we have

$$
\begin{equation*}
A(y, 0, t) \equiv 0 \tag{7.27}
\end{equation*}
$$

from which we deduce

$$
\begin{align*}
\left|A_{y}(y, x, t)\right| & \leqslant c|x|  \tag{7.28}\\
\left|A_{t}(y, x, t)\right| & \leqslant c|x| \\
\left|A_{y y}(y, x, t)\right| & \leqslant c|x| \\
\left|A_{y t}(y, x, t)\right| & \leqslant c|x|
\end{align*}
$$

Denote by $\mu=\mu(t)$ the unstable eigenvalue of $D_{g_{t}}$ at $p_{s}$. For all $t, x, y, n$, we have

$$
\begin{equation*}
c^{-1} \mu^{-n} \leqslant\left|A_{x}^{(n)}(x, y, t)\right| \leqslant c \mu^{-n} . \tag{7.29}
\end{equation*}
$$

Let $\left(x_{i}, y_{i}\right)_{0 \leqslant i \leqslant n}$ be an orbit of $g_{t}$ in $R_{a_{s}}$. For all $0 \leqslant \ell \leqslant m \leqslant n$, we have:

$$
\begin{equation*}
c^{-1} \mu^{m-\ell}\left|x_{\ell}\right| \leqslant\left|x_{m}\right| \leqslant c \mu^{m-\ell}\left|x_{\ell}\right| \tag{7.30}
\end{equation*}
$$

Proposition 14. The following estimates hold:

$$
\begin{align*}
& \left|A_{y}^{(n)}\left(y_{0}, x_{n}, t\right)\right| \leqslant c\left|x_{0}\right| \leqslant c \mu^{-n}\left|x_{n}\right|,  \tag{7.31}\\
& \left|A_{t}^{(n)}\left(y_{0}, x_{n}, t\right)\right| \leqslant c n\left|x_{0}\right| \leqslant c n \mu^{-n}\left|x_{n}\right|,  \tag{7.32}\\
& \left|A_{y y}^{(n)}\left(y_{0}, x_{n}, t\right)\right| \leqslant c\left|x_{0}\right| \leqslant c \mu^{-n}\left|x_{n}\right| \text {. } \tag{7.33}
\end{align*}
$$

Proof of 7.31: From formula (3.11) in Subsection 3.3, we have:

$$
\begin{equation*}
A_{y}^{(n)}\left(y_{0}, x_{n}, t\right)=A_{y}^{(n-1)}\left(y_{0}, x_{n-1}, t\right)+A_{y} A_{x}^{(n-1)} B_{y}^{(n-1)} \Delta^{-1} \tag{7.34}
\end{equation*}
$$

with $\left.B_{y}^{( } n-1\right) \Delta$ exponentially small with $n$ and, using (7.28)-(7.30):

$$
\begin{equation*}
\left|A_{y}\left(y_{n-1}, x_{n}, t\right) A_{x}^{(n-1)}\left(y_{0}, x_{n-1}, t\right)\right| \leqslant c\left|x_{0}\right| . \tag{7.35}
\end{equation*}
$$

The inequality (7.31) is now clear.

Proof of 7.32: We use here formulas (A6), (A10) of Appendix A which give

$$
\begin{gather*}
\left|A_{t}^{(n)}\left(y_{0}, x_{n}, t\right)-A_{t}^{(n-1)}\left(y_{0}, x_{n-1}, t\right)\right|  \tag{7.36}\\
\leqslant C \mu^{-n}\left(\left|A_{t}\left(y_{n-1}, x_{n}, t\right)\right|+\left|B_{t}^{(n-1)}\left(y_{0}, x_{n-1}, t\right)\right|\left|A_{y}\left(y_{n-1}, x, t\right)\right|\right), \\
\left|B_{t}^{(n)}\left(y_{0}, x_{n}, t\right)-B_{t}^{(n-1)}\left(y_{0}, x_{n-1}, t\right)\right| \leqslant C\left|B_{y}^{(n-1)}\right|_{C^{0}}\left(1+\left|A_{t}^{(n-1)}\right|_{C^{0}}\right) . \tag{7.37}
\end{gather*}
$$

As $\left|B_{y}^{(n)}\right|_{C_{0}}$ is exponentially small, we deduce from (7.37) that

$$
\begin{equation*}
\left|B_{t}^{(n)}\left(y_{0}, x_{n}, t\right)\right| \leqslant C, \tag{7.38}
\end{equation*}
$$

and then, from (7.36), (7.28) that (7.32) holds.

Proof of 7.33: We use formulas (A6), (A18), (A20) of Appendix A to obtain

$$
\begin{equation*}
A_{y y}^{(n)}=A_{y y}^{(n-1)}+2 A_{x y}^{(n-1)} X_{y}+A_{x x}^{(n-1)} X_{y}^{2}+A_{x}^{(n-1)} X_{y y} \tag{7.39}
\end{equation*}
$$

with

$$
\begin{align*}
X_{y} & =A_{y} B_{y}^{(n-1)} \Delta^{-1}  \tag{7.40}\\
\Delta & =1-A_{y} B_{x}^{(n-1)}  \tag{7.41}\\
X_{y y} & =B_{y}^{(n-1)} \Delta^{-1}\left(A_{y y} B_{y}^{(n-1)} \Delta^{-1}+A_{y} \partial_{y} \log \left|B_{y}^{(n-1)}\right|+\right.  \tag{7.42}\\
& \left.+A_{y} X_{y} \partial_{x} \log \left|B_{y}^{(n-1)}\right|-A_{y}^{(n-1)} \Delta_{y} \Delta^{-1}\right) \\
-\Delta_{y} & =A_{y y} B_{y}^{(n-1)} B_{x}^{(n-1)} \Delta^{-1}+A_{y} B_{x y}^{(n-1)}+A_{y} B_{x x}^{(n-1)} X_{y} \tag{7.43}
\end{align*}
$$

In these formulas, $A^{(n-1)}, B^{(n-1)}$ and their derivatives are taken at $\left(y_{0}, x_{n-1}, t\right), A, B$ and their derivatives are taken at $\left(y_{n-1}, x_{n}, t\right)$. The terms $B_{x}^{(n-1)}, B_{x x}^{(n-1)}, \partial_{x} \log \left|B_{y}^{(n-1)}\right|, \partial_{y} \log \left|B_{y}^{(n-1)}\right|, \Delta^{-1}$ are bounded by the uniform cone condition and the uniform distortion; the terms $B_{y}^{(n-1)}, B_{x y}^{(n-1)}$ are exponentially small. Also, from (7.28) we have:

$$
\begin{align*}
\left|A_{y}\left(y_{n-1}, x_{n}, t\right)\right| & \leqslant C\left|x_{n}\right|  \tag{7.44}\\
\left|A_{y y}\left(y_{n-1}, x_{n}, t\right)\right| & \leqslant C\left|x_{n}\right|
\end{align*}
$$

We conclude that we can write

$$
\begin{equation*}
A_{y y}^{(n)}\left(y_{0}, x_{n}, t\right)=A_{y y}^{(n)}\left(y_{0}, x_{n-1}, t\right)+\mu^{-n} x_{n} r_{n} \tag{7.45}
\end{equation*}
$$

with $r_{n}$ exponentially small; this leads to (7.34).
Corollary 10. For the special rectangle $\left(P_{s}, Q_{s}, n_{s}\right)$, we have:

$$
\begin{aligned}
\left|A_{y}^{\left(n_{s}\right)}\right|_{C^{0}} & \leqslant C \varepsilon_{0} \\
\left|A_{y y}^{\left(n_{s}\right)}\right|_{C^{0}} & \leqslant C \varepsilon_{0} \\
\left|A_{t}^{\left(n_{s}\right)}\right|_{C^{0}} & \leqslant C \varepsilon_{0} \log \varepsilon_{0}^{-1}
\end{aligned}
$$

Proof. We have only to observe that $\mu^{-n_{s}}$ is of order $\varepsilon_{0}$.

Obviously, the same estimates hold for the other special element $\left(P_{u}, Q_{u}, n_{u}\right)$.

### 7.5 Proof of the Property (R4) of Section 5.3

We have to show that, for an element $(P, Q, n)$ in $\mathcal{R}(I)$, with associated implicit representation $(A, B)$, we have

$$
\begin{equation*}
\left|A_{y}\right| \leqslant C \varepsilon_{0}, \quad\left|A_{y y}\right| \leqslant C \varepsilon_{0} \tag{7.46}
\end{equation*}
$$

whenever $P \subset P_{s}$ and

$$
\begin{equation*}
\left|B_{x}\right| \leqslant C \varepsilon_{0}, \quad\left|B_{x x}\right| \leqslant C \varepsilon_{0} \tag{7.47}
\end{equation*}
$$

whenever $Q \subset Q_{u}$. We will deal only with (7.46), the other case being symmetric.

We have already proved (7.46) when $P=P_{s}$, cf. the first two inequalities of Corollary 10. We will prove (7.46) by induction on the length $n$. Let us denote by $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ the element of $\mathcal{R}(I)$ such that $\widetilde{P}$ is the parent of $P$, by $(\widetilde{A}, \widetilde{B})$ the implicit representation of this affine-like iterate. We assume now that $P \varsubsetneqq P_{s}$, i.e., $\widetilde{P} \subset P_{s}$. The are two possibilities.

First, we consider the easier case when $P$ is a simple child of $\widetilde{P}$. Let $\left(P^{*}, Q^{*}, 1\right)$, with associated implicit representation $\left(A^{*}, B^{*}\right)$, be the element such that

$$
\begin{equation*}
(P, Q, n)=(\widetilde{P}, \widetilde{Q}, \widetilde{n}) *\left(P^{*}, Q^{*}, 1\right) \tag{7.48}
\end{equation*}
$$

From formula (3.11) in Subsection 3.3, we have:

$$
\begin{equation*}
A_{y}=\widetilde{A}_{y}+\widetilde{A}_{x} \widetilde{B}_{y} A_{y}^{*} \Delta^{-1} \tag{7.49}
\end{equation*}
$$

with $\left|\widetilde{A}_{x}\right| \leq C \varepsilon_{0}, A_{y}^{*}, \Delta^{-1}$ bounded and $\widetilde{B}_{y}$ satisfying the stretched exponential estimate of Proposition 13: this is fully in line with the first inequality in (7.46).

Next, we have, as in (7.39) above:

$$
\begin{equation*}
A_{y y}=\widetilde{A}_{y y}+2 \widetilde{A}_{x y} X_{y}+\widetilde{A}_{x x} X_{y}^{2}+\widetilde{A}_{x} X_{y y} \tag{7.50}
\end{equation*}
$$

with

$$
\begin{align*}
X_{y} & =A_{y}^{*} \widetilde{B}_{y} \Delta^{-1}  \tag{7.51}\\
\Delta & =1-A_{y}^{*} \widetilde{B}_{x}  \tag{7.52}\\
X_{y y} & =\Delta^{-1} \widetilde{B}_{y}\left(A_{y y}^{*} \widetilde{B}_{y} \Delta^{-1}+A_{y}^{*} \partial_{y} \log \left|\widetilde{B}_{y}\right|+A_{y}^{*} X_{y} \partial_{x} \log \left|\widetilde{B}_{y}\right|-\widetilde{A}_{y} \Delta_{y} \Delta^{-1}\right)  \tag{7.53}\\
-\Delta_{y} & =A^{*}{ }_{y y} \widetilde{B}_{y} \widetilde{B}_{x} \Delta^{-1}+A_{y}^{*} \widetilde{B}_{x y}+A_{y}^{*} \widetilde{B}_{x x} X_{y} \tag{7.54}
\end{align*}
$$

Here, we have:

$$
\begin{align*}
\left|\widetilde{A}_{x y}\right| & =\left|\widetilde{A}_{x}\right|\left|\partial_{y} \log \right| \widetilde{A}_{x}| | \leqslant C \varepsilon_{0}  \tag{7.55}\\
\left|\widetilde{A}_{x x}\right| & =\left|\widetilde{A}_{x}\right|\left|\partial_{x} \log \right| \widetilde{A}_{x}| | \leqslant C \varepsilon_{0},  \tag{7.56}\\
\left|\widetilde{A}_{x}\right| & \leqslant C \varepsilon_{0} . \tag{7.57}
\end{align*}
$$

Moreover, $\widetilde{B}_{y}$ satisfies the stretched exponential estimate of Proposition 13; the uniform cone condition and uniform distortion imply that the same stretched exponential estimate holds for $X_{y}$ and $X_{y y}$. Therefore, the estimate for $A_{y y}-\widetilde{A}_{y y}$ in (7.50) is fully in line with the second inequality in (7.46). This concludes the case where $P$ is a simple child.

We now consider the more difficult case where $P$ is a non-simple child of $\widetilde{P}$. By Proposition 5 , we can find an element $\left(P^{*}, Q^{*}, n^{*}\right) \in \mathcal{R}(I)$, with associated implicit representation $\left(A^{*}, B^{*}\right)$, such that:

$$
\begin{equation*}
(P, Q, n) \in(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \square\left(P^{*}, Q^{*}, n^{*}\right) \tag{7.58}
\end{equation*}
$$

From the formulas in Appendix A, we have:

$$
\begin{align*}
A_{y} & =\widetilde{A}_{y}+\widetilde{A}_{x}\left(X_{y}+X_{w} W_{y}\right)  \tag{7.59}\\
X_{y} & =X_{u, y} \widetilde{B}_{y} \Delta_{0}^{-1}  \tag{7.60}\\
X_{w} & =X_{u, w} \Delta_{0}^{-1}  \tag{7.61}\\
\Delta_{0} & =1-X_{u, y} \widetilde{B}_{x}  \tag{7.62}\\
W_{y} & =-C_{y} C_{w}^{-1}  \tag{7.63}\\
-C_{y} & =\theta_{y} \widetilde{B}_{y} \Delta_{0}^{-1} \tag{7.64}
\end{align*}
$$

Here $\theta, X_{u}, W$ and $C$ are associated with the fold map $G$ as in Subsections 2.3, 3.5.
As $\left|C_{w w}-2\right|$ is small (cf. 3.22)), the quantity $C_{w}$ is related to $\delta\left(\widetilde{Q}, P^{*}\right)$ by

$$
\begin{equation*}
C^{-1} \delta\left(\widetilde{Q}, P^{*}\right)^{\frac{1}{2}} \leqslant\left|C_{w}\right| \leqslant C \delta\left(\widetilde{Q}, P^{*}\right)^{\frac{1}{2}} \tag{7.65}
\end{equation*}
$$

From (R7), $\delta\left(\widetilde{Q}, P^{*}\right)$ is always much bigger than $\left|\widetilde{B}_{y}\right|$; therefore, from (7.63), (7.64) we have:

$$
\begin{equation*}
\left|W_{y}\right| \leqslant C\left|\widetilde{B}_{y}\right|^{\frac{1}{2}} \tag{7.66}
\end{equation*}
$$

We then obtain from Proposition 13, that the term $X_{y}+X_{w} W_{y}$ in (7.59) satisfies a stretched exponential estimate; in view of (7.57), this concludes the proof of the first inequality in (7.46).

For the second inequality, we write, following the formulas in Appendix A:

$$
\begin{align*}
A_{y y} & =\widetilde{A}_{y y}+2 \widetilde{A}_{x y}\left(X_{y}+X_{w} W_{y}\right)+\widetilde{A}_{x x}\left(X_{y}+X_{w} W_{y}\right)^{2}+  \tag{7.67}\\
& +\widetilde{A}_{x}\left(X_{y y}+2 X_{w y} W_{y}+X_{w w} W_{y}^{2}+X_{w} W_{y y}\right) \\
W_{y y} & =-C_{w}^{-1}\left(C_{w w} W_{y}^{2}+2 C_{w y} W_{y}+C_{y y}\right)  \tag{7.68}\\
-C_{w y} & =\theta_{y y} \widetilde{B}_{x} \widetilde{B}_{y} X_{w} \Delta_{0}^{-1}+\theta_{x y} \widetilde{B}_{y} A_{y}^{*} Y_{s, w} \Delta_{0}^{-1} \Delta_{1}^{-1}+\theta_{y} \bar{Y}_{w y}  \tag{7.69}\\
-C_{y y} & =\theta_{y y} \widetilde{B}_{y}^{2} \Delta_{0}^{-2}+\theta_{y} \bar{Y}_{y y}  \tag{7.70}\\
\bar{Y}_{w y} & =\widetilde{B}_{x y} X_{w}+\widetilde{B}_{x x} X_{w} X_{y}+\widetilde{B}_{x} X_{w y}  \tag{7.71}\\
\bar{Y}_{y y} & =\widetilde{B}_{y y}+2 \widetilde{B}_{x y} X_{y}+\widetilde{B}_{x x} X_{y}^{2}+\widetilde{B}_{x} X_{w y}  \tag{7.72}\\
X_{w w} & =\Delta_{0}^{-1}\left(X_{u, w w}+2 \widetilde{B}_{x} X_{w} X_{u, w y}+\widetilde{B}_{x}^{2} X_{u, y y} X_{w}^{2}+\widetilde{B}_{x x} X_{u, y} X_{w}^{2}\right)  \tag{7.73}\\
X_{w y} & =\Delta_{0}^{-1}\left[\left(X_{u, w y}+\widetilde{B}_{x} X_{u, y y} X_{w}\right)\left(\widetilde{B}_{y}+\widetilde{B}_{x} X_{y}\right)\right.  \tag{7.74}\\
& \left.+X_{u, y} X_{w}\left(\widetilde{B}_{x y}+\widetilde{B}_{x x} X_{y}\right)\right] \\
X_{y y} & =\Delta_{0}^{-1}\left[X_{u, y y}\left(\widetilde{B}_{y}+\widetilde{B}_{x} X_{y}\right)^{2}+X_{u, y}\left(\widetilde{B}_{y y}+2 \widetilde{B}_{x y} X_{y}+\widetilde{B}_{x x} X_{y}^{2}\right)\right] \tag{7.75}
\end{align*}
$$

where $\Delta_{0}^{-1}, \Delta_{1}^{-1}$ are bounded. In formula (7.67), we have that
$-\left|\widetilde{A}_{x y}\right|,\left|\widetilde{A}_{x x}\right|,\left|\widetilde{A}_{x}\right|$ are bounded by $C \varepsilon_{0}$, cf. (7.55)-(7.57) above;
$-X_{w}, X_{w w}$ are bounded, cf. (7.61) and (7.73);

- $X_{y}, W_{y}$ satisfy a stretched exponential estimate, as we have already seen earlier;
- $X_{w y}, X_{y y}$ also satisfy a stretched exponential estimate, cf. (7.74) and (7.75).

The remaining term in (7.67) is $\widetilde{A}_{x} X_{w} W_{y y}$, and we have to estimate $W_{y y}$ from (7.68). From (7.75), we have, using bounded distortion:

$$
\begin{equation*}
\left|X_{y y}\right| \leqslant C\left|\widetilde{B}_{y}\right| \tag{7.76}
\end{equation*}
$$

from which we deduce, by (7.72), that

$$
\begin{equation*}
\left|\bar{Y}_{y y}\right| \leqslant C\left|\widetilde{B}_{y}\right| . \tag{7.77}
\end{equation*}
$$

In the same way, we obtain

$$
\begin{align*}
&\left|X_{w y}\right| \leqslant C\left|\widetilde{B}_{y}\right|,  \tag{7.78}\\
&\left|\bar{Y}_{w y}\right| \leqslant C\left|\widetilde{B}_{y}\right| . \tag{7.79}
\end{align*}
$$

Plugging this into (7.69), (7.70) yields

$$
\begin{align*}
\left|C_{w y}\right| & \leqslant C\left|\widetilde{B}_{y}\right|,  \tag{7.80}\\
\left|C_{y y}\right| & \leqslant C\left|\widetilde{B}_{y}\right| . \tag{7.81}
\end{align*}
$$

We can now conclude, using (7.65), (7.66), that

$$
\begin{align*}
\left|W_{y y}\right| & \leqslant C\left|\widetilde{B}_{y}\right| \delta\left(\widetilde{Q}, P^{*}\right)^{-\frac{1}{2}}  \tag{7.82}\\
& \leqslant C\left|\widetilde{B}_{y}\right|^{\frac{1}{2}} .
\end{align*}
$$

All terms in (7.67) are now under control, $A_{y y}-\widetilde{A}_{y y}$ being bounded by $C \varepsilon_{0}$ times a stretched exponentially small term: this concludes the proof of the second inequality in (7.46), and so of condition (R4).

### 7.6 Relative Speeds of Critical Rectangles

Let $I$ be a parameter interval, and let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}(I)$ such that $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$.

The displacements $\delta\left(Q_{0}, P_{1}\right), \delta_{L}\left(Q_{0}, P_{1}\right), \delta_{R}\left(Q_{0}, P_{1}\right), \delta_{L R}\left(Q_{0}, P_{1}\right)$ were introduced in formulas (5.5) and (5.8)-(5.10) of Subsection 5.4 (see also (3.23) in Subsection 3.5) and are the values at the four corners of the rectangle of definition of the function $\bar{C}\left(y_{0}, x_{1}\right)$ introduces in Subsection 3.5 as

$$
\begin{equation*}
\bar{C}\left(y_{0}, x_{1}\right)=\min _{w} C\left(w, y_{0}, x_{1}\right) . \tag{7.83}
\end{equation*}
$$

All these quantities also depend on the parameter $t$, and we want in this section to estimate the variation with the parameter of the displacements, which amounts to estimate the partial derivative $C_{t}$.

Let $\left(A_{0}, B_{0}\right),\left(A_{1}, B_{1}\right)$ be the implicit representations for $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ respectively. As will be seen below, an estimate for $C_{t}$ depends very much on estimates for the partial derivatives $A_{1, t}, B_{0, t}$. Good estimates for these two quantities are not available for all $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$. We will only consider elements satisfying conditions $(* s),(* u)$ below; fortunately, these conditions will always be satisfied whenever we are interested in the variation of the displacements.

Let $(P, Q, n)$ be an element of $\mathcal{R}(I)$, such that $P \subset P_{s}$. Let $I^{*}$ be the largest parameter interval such that $(P, Q, n)$ is (the restriction of) an element of $\mathcal{R}\left(I^{*}\right)$. If $I^{*}$ is not the starting interval $I_{0}=\left[\varepsilon_{0}, 2 \varepsilon_{0}\right]$, let $\widetilde{I}^{*}$ be the parent interval and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be the longest element of $\mathcal{R}\left(\widetilde{I^{*}}\right)$ such that $P \subset \widetilde{P} ;(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ is the element which is denoted by $\left(P_{0}, Q_{0}, n_{0}\right)$ in the structure theorem of Subsection 6.7, as $[0, \widetilde{n}]$ is the maximal initial $\widetilde{I}^{*}$-interval.

We say that $(P, Q, n)$ satisfies condition $(* s)$ if either $I^{*}=I_{0}$ or $I^{*} \neq I_{0}$ and $\widetilde{P}$ is $\widetilde{I}^{*}$-critical.
We define in a symmetric way a condition $(* u)$ (when $\left.Q \subset Q_{u}\right)$.
Proposition 15. Let $(P, Q, n)$ be an element of $\mathcal{R}(I)$ with $P \subset P_{s}$. Let $(A, B)$ be the implicit representation of $(P, Q, n)$. If $(P, Q, n)$ satisfies condition $(* s)$, we have

$$
\left|A_{t}\right| \leqslant \varepsilon_{0}^{\frac{1}{2}}
$$

If $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$, we have the stronger estimate:

$$
\left|A_{t}\right| \leqslant C \varepsilon_{0} \log \varepsilon_{0}^{-1}
$$

Proof. We first show for $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$ and $P \subset P_{s}$, that:

$$
\begin{equation*}
\left|A_{t}\right| \leqslant C \varepsilon_{0} \log \varepsilon_{0}^{-1},\left|B_{t}\right| \leqslant C . \tag{7.84}
\end{equation*}
$$

When $P=P_{s}$, the first inequality is part of Corollary 10 and the second is (7.38). If $P \varsubsetneqq P_{s}$, we set

$$
\begin{equation*}
(P, Q, n)=(\widetilde{P}, \widetilde{Q}, \widetilde{n}) *\left(P^{*}, Q^{*}, 1\right) . \tag{7.85}
\end{equation*}
$$

Writing $(\widetilde{A}, \widetilde{B})$ and $\left(A^{*}, B^{*}\right)$ for the implicit representations of $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ and ( $P^{*}, Q^{*}, 1$ ), respectively, we have, from the formulas in Appendix A:

$$
\begin{align*}
& A_{t}=\widetilde{A}_{t}+\widetilde{A}_{x}\left(A_{t}^{*}+A_{y}^{*} \widetilde{B}_{t}\right) \Delta^{-1}  \tag{7.86}\\
& B_{t}=B_{t}^{*}+B_{y}^{*}\left(\widetilde{B}_{t}+\widetilde{B}_{x} A_{t}^{*}\right) \Delta^{-1} \tag{7.87}
\end{align*}
$$

with, as usual, $\Delta=1-\widetilde{B}_{x} A_{y}^{*}$.
This gives, assuming (7.84) for ( $\widetilde{A}, \widetilde{B}$ ),

$$
\begin{align*}
\left|A_{t}-\widetilde{A}_{t}\right| & \leqslant C|P|,  \tag{7.88}\\
\left|B_{t}-\widetilde{B}_{t} B_{y} \widetilde{B}_{y}^{-1}\right| & \leqslant C, \tag{7.89}
\end{align*}
$$

where we have used $B_{y}=\widetilde{B}_{y} B_{y}^{*} \Delta^{-1}$. Clearly, (7.88) and (7.89) imply, by iteration, (7.84).
We now turn to the case where $I^{*} \neq I_{0}$. In this case, we introduce the integer $k \geqslant 1$ and the elements $\left(P_{i}, Q_{i}, n_{i}\right), 0 \leqslant i \leqslant k$, of $\mathcal{R}\left(\widetilde{I}^{*}\right)$ given by the structure theorem of Subsection 6.7. We also denote by $\left(P^{(i)}, Q^{(i)}, n^{(i)}\right)$ the element of $\mathcal{R}\left(\widetilde{I}^{*}\right)$ such that $P \subset P^{(i)}$ and

$$
\begin{equation*}
\left(P^{(i)}, Q^{(i)}, n^{(i)}\right) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{i}, Q_{i}, n_{i}\right) . \tag{7.90}
\end{equation*}
$$

Our proof will be by induction, on the level of the parameter interval and on the integer $k$ (for a fixed parameter interval).

We first observe that if $P$ is $I$-critical, then $P_{0}$ is $\widetilde{I}^{*}$-critical and $P$ satisfies ( $* s$ ). Therefore, if $P$ satisfies ( $* s$ ), all $P^{(i)}$ also satisfy ( $* s$ ); moreover, $Q^{(i)}$ satisfy ( $* u$ ) and $P_{i+1}$ satisfy ( $* s$ ) by Theorem 1.

Fix $0 \leqslant i<k$ and let $\left(A^{(i+1)}, B^{(i+1)}\right)$, $\left(A^{(i)}, B^{(i)}\right)$ and $\left(A^{*}, B^{*}\right)$ be the implicit representations of $\left(P^{(i+1)}, Q^{(i+1)}, n^{(i+1)}\right),\left(P^{(i)}, Q^{(i)}, n^{(i)}\right)$ and $\left(P_{i+1}, Q_{i+1}, n_{i+1}\right)$ respectively. From the formulas in

Appendix A, we have

$$
\begin{align*}
A_{t}^{(i+1)} & =A_{t}^{(i)}+A_{x}^{(i)}\left(X_{t}+X_{w} W_{t}\right),  \tag{7.91}\\
X_{t} & =\left(X_{u, t}+X_{u, y} B_{t}^{(i)}\right) \Delta_{0}^{-1},  \tag{7.92}\\
X_{w} & =X_{u, w} \Delta_{0}^{-1},  \tag{7.93}\\
W_{t} & =-C_{t} C_{w}^{-1} . \tag{7.94}
\end{align*}
$$

In these formulas, $X_{u, t}, X_{u, y}, X_{u, w}, \Delta_{0}^{-1}$ are uniformly bounded and $B_{t}^{(i)}$ is bounded by $\varepsilon_{0}^{1 / 2}$ from the induction hypothesis. The term $C_{w}^{-1}$ is estimated by (7.65) above and $C_{t}$ is uniformly bounded by the induction hypothesis and Corollary 11 below. We conclude that

$$
\begin{align*}
\left|A_{t}^{(i+1)}-A_{t}^{(i)}\right| & \leqslant C\left|P^{(i)}\right| \delta\left(Q^{(i)}, P_{i+1}\right)^{-\frac{1}{2}}  \tag{7.95}\\
& \leqslant C\left|P^{(i)}\right|\left|I^{*}\right|^{-\frac{1}{2}}
\end{align*}
$$

as $Q^{(i)}$ and $P_{i+1}$ are $I^{*}$-transverse. We have for $0 \leqslant j \leqslant i$

$$
\begin{equation*}
\left|P_{j}\right| \leqslant\left|\widetilde{I}^{*}\right|^{\beta} \tag{7.96}
\end{equation*}
$$

because ( $P_{j}, Q_{j}, n_{j}$ ) is $\widetilde{I}^{*}$-bicritical (we use here, for $j=0$, the hypothesis that $P_{0}$ is $\widetilde{I}^{*}$-critical).
Then, by Corollary 6 , we obtain

$$
\begin{equation*}
\left|P^{(i)}\right| \leqslant C^{i}\left|\widetilde{I}^{*}\right|^{(i+1) \beta-\frac{i}{2}(1+\tau)} \tag{7.97}
\end{equation*}
$$

which implies

$$
\begin{equation*}
\left|A_{t}^{(i+1)}-A_{t}^{(i)}\right| \leqslant\left|\widetilde{I^{*}}\right|^{\frac{\beta}{2}}(1+i) . \tag{7.98}
\end{equation*}
$$

Summing (7.98) over $i$ and the levels of parameter intervals leads to the estimate of the proposition.

Corollary 11. Let $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ be elements of $\mathcal{R}(I)$ with $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$. Assume that $Q_{0}$ satisfies ( $* u$ ) and that $P_{1}$ satisfies $(* s)$. Then, the function $C$ introduced in Subsection 3.5 satisfy

$$
\left|C_{t}+1\right| \leqslant C \varepsilon_{0}^{\frac{1}{2}}
$$

Proof. From formula (A35) in Appendix A, using the notations there, we have

$$
\begin{align*}
-C_{t} & =\theta_{x} \bar{X}_{t}+\theta_{y} \bar{Y}_{t}+\theta_{t},  \tag{7.99}\\
\bar{X}_{t} & =\left(A_{1, t}+A_{1, y} Y_{s, t}\right) \Delta_{1}^{-1},  \tag{7.100}\\
\bar{Y}_{t} & =\left(B_{0, t}+B_{0, x} X_{u, t}\right) \Delta_{0}^{-1}, \tag{7.101}
\end{align*}
$$

with $\Delta_{0}^{-1}, \Delta_{1}^{-1}$ uniformly bounded. The value of $\theta_{t}$ is taken at $(\bar{X}, \bar{Y}, t)$, with

$$
\begin{equation*}
|\bar{X}| \leqslant C \varepsilon_{0}, \quad|\bar{Y}| \leqslant C \varepsilon_{0} . \tag{7.102}
\end{equation*}
$$

On the other hand, we have, in Subsection 4.2, normalized the parameter in order to have

$$
\begin{equation*}
\theta_{t}(0,0, t) \equiv 1 . \tag{7.103}
\end{equation*}
$$

We, therefore, have

$$
\begin{equation*}
\left|\theta_{t}(\bar{X}, \bar{Y}, t)-1\right| \leqslant C \varepsilon_{0} . \tag{7.104}
\end{equation*}
$$

In (7.100) and (7.101), we have $\left|A_{1, y}\right|<C \varepsilon_{0},\left|B_{0, x}\right|<C \varepsilon_{0}$, by (R4) and $\left|A_{1, t}\right|<\varepsilon_{0}^{\frac{1}{2}},\left|B_{0, t}\right|<\varepsilon_{0}^{\frac{1}{2}}$ by Proposition 15. The Corollary follows, as $\theta_{x}, \theta_{y}, Y_{s, t}, X_{u, t}$ are uniformly bounded.

### 7.7 Variation of Width of Critical Rectangles

Our main purpose now is to prove property (R7) of Subsection 5.4:
(R7) If $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$ satisfy $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$ and $Q_{0} \pitchfork_{I} P_{1}$ holds, then, for all $t \in I$, we have

$$
\delta\left(Q_{0}, P_{1}\right) \geqslant C^{-1}\left(\left|P_{1}\right|^{1-\eta}+\left|Q_{0}\right|^{1-\eta}\right) .
$$

A priori, the transversality condition gives some control through (T2), (T3) in Subsection 5.4 only for some values of the parameter. However, from (T1) in Subsection 5.4 and Corollary 11 above, we know that the order of magnitude of $\delta\left(Q_{0}, P_{1}\right)$ is the same through out $I$. Therefore, to obtain (R7), we do need to control how the widths $\left|P_{1}\right|$ and $\left|Q_{0}\right|$ vary through $I$. Good estimates will be obtained under the same conditions $(* s)$ or $(* u)$ used to obtain Proposition 15. Again, the estimates are even better for an element $(P, Q, n)$ in $\mathcal{R}\left(I_{0}\right)$, involving only simple composition.

Proposition 16. Let $(P, Q, n)$ be an element of $\mathcal{R}\left(I_{0}\right)$, and let $(A, B)$ be the implicit representation of $(P, Q, n)$. We have

$$
\begin{array}{ll}
\left|\partial_{t} \log \right| A_{x}| | \leqslant C n, & \left|A_{y t}\right| \leqslant C \\
\left|\partial_{t} \log \right| B_{y}| | \leqslant C n, & \left|B_{x t}\right| \leqslant C .
\end{array}
$$

Proof. We first observe that we have, by the same proof as for the second inequality in (7.84):

$$
\begin{equation*}
\left|A_{t}\right| \leqslant C, \quad\left|B_{t}\right| \leqslant C \tag{7.105}
\end{equation*}
$$

We write

$$
\begin{equation*}
(P, Q, n)=(\widetilde{P}, \widetilde{Q}, \widetilde{n}) *\left(P^{*}, Q^{*}, 1\right), \tag{7.106}
\end{equation*}
$$

and denote by $(\widetilde{A}, \widetilde{B}),\left(A^{*}, B^{*}\right)$ the implicit representations of $(\widetilde{P}, \widetilde{Q}, \widetilde{n}),\left(P^{*}, Q^{*}, 1\right)$, respectively.

By the formulas of Appendix A, we have

$$
\begin{align*}
\partial_{t} \log \left|A_{x}\right| & =\partial_{t} \log \left|\widetilde{A}_{x}\right|+\partial_{t} \log \left|A_{x}^{*}\right|+X_{t} \partial_{x} \log \left|\widetilde{A}_{x}\right|  \tag{7.107}\\
& +Y_{t} \partial_{y} \log \left|A_{x}^{*}\right|-\Delta_{t} \Delta^{-1} \\
X_{t} & =\left(A_{t}^{*}+A_{y}^{*} \widetilde{B}_{t}\right) \Delta^{-1}  \tag{7.108}\\
Y_{t} & =\left(\widetilde{B}_{t}+A_{t}^{*} \widetilde{B}_{x}\right) \Delta^{-1}  \tag{7.109}\\
\Delta & =1-A_{y}^{*} \widetilde{B}_{x}  \tag{7.110}\\
-\Delta_{t} & =\widetilde{B}_{x t} A_{y}^{*}+\widetilde{B}_{x x} A_{y}^{*} X_{t}+\widetilde{B}_{x} A_{y t}^{*}+\widetilde{B}_{x} A_{y y}^{*} Y_{t} \tag{7.111}
\end{align*}
$$

As $\widetilde{B}_{t}, \widetilde{B}_{x}, A_{t}^{*}, A_{y}^{*}, \Delta^{-1}$ are uniformly bounded, the same is true for $X_{t}, Y_{t}$. Using bounded distortion and the cone condition then leads to

$$
\begin{equation*}
\left|\partial_{t} \log \right| A_{x}\left|-\partial_{t} \log \right| \widetilde{A}_{x}| | \leqslant C\left(1+\left|\widetilde{B}_{x t}\right|\right) . \tag{7.112}
\end{equation*}
$$

Still from Appendix A, we have

$$
\begin{align*}
A_{y t}-\widetilde{A}_{y t} & =X_{t} \widetilde{A}_{x y}+X_{y} \widetilde{A}_{x t}+X_{t} X_{y} \widetilde{A}_{x x}+\widetilde{A}_{x} X_{y t}  \tag{7.113}\\
X_{y} & =A_{y}^{*} \widetilde{B}_{y} \Delta^{-1}  \tag{7.114}\\
X_{y t} & =\widetilde{B}_{y} \Delta^{-1}\left(A_{y y}^{*} Y_{t}+A_{y}^{*} \partial_{t} \log \left|\widetilde{B}_{y}\right|+A_{y t}^{*}\right.  \tag{7.115}\\
& \left.+A_{y}^{*} X_{t} \partial_{x} \log \left|\widetilde{B}_{y}\right|-A_{y}^{*} \Delta_{t} \Delta^{-1}\right)
\end{align*}
$$

As $X_{t}, X_{y}, Y_{t}$ are bounded and also using bounded distortion, we have

$$
\begin{align*}
\left|X_{t} \widetilde{A}_{x y}\right| & \leqslant C\left|\widetilde{A}_{x}\right|  \tag{7.116}\\
\left|X_{y} \widetilde{A}_{x t}\right| & \leqslant C\left|\widetilde{A}_{x}\right|\left|\partial_{t} \log \right| \widetilde{A}_{x}| |  \tag{7.117}\\
\left|X_{t} X_{y} \widetilde{A}_{x x}\right| & \leqslant C\left|\widetilde{A}_{x}\right|  \tag{7.118}\\
\left|X_{y t}\right| & \leqslant C\left|\widetilde{B}_{y}\right|\left(1+\left|\partial_{t} \log \widetilde{B}_{y}\right|+\left|\widetilde{B}_{x t}\right|\right) \tag{7.119}
\end{align*}
$$

We, therefore, obtain

$$
\begin{equation*}
\left|A_{y t}-\widetilde{A}_{y t}\right| \leqslant C\left|\widetilde{A}_{x}\right|\left(1+\left|\partial_{t} \log \right| A_{x}| |\right)+C\left|\widetilde{A}_{x}\right|\left|\widetilde{B}_{y}\right|\left(\left|\partial_{t} \log \right| \widetilde{B}_{y}| |+\left|\widetilde{B}_{x t}\right|\right) \tag{7.120}
\end{equation*}
$$

We have symmetric estimates for $B_{x t}$ and $\partial_{t} \log \left|B_{y}\right|$, writing now

$$
\begin{equation*}
(P, Q, n)=\left(\widehat{P}^{*}, \widehat{Q}^{*}, 1\right) *(\widehat{P}, \widehat{Q}, n-1) \tag{7.121}
\end{equation*}
$$

As $\left|\widetilde{A}_{x}\right|,\left|\widetilde{B}_{y}\right|$ are exponentially small, the estimates $(7.112)$, (7.120) and the other two estimates for $B_{x t}, \partial_{t} \log \left|B_{y}\right|$ lead by summation to the bounds of the proposition.

For the last estimate of this section, the setting is the same as that of Proposition 15.

Proposition 17. Let $(P, Q, n)$ be an element of $\mathcal{R}(I)$ with $P \subset P_{s}$. Let $(A, B)$ be the implicit representation of $(P, Q, n)$. If $(P, Q, n)$ satisfies condition $(* s)$, we have

$$
\begin{align*}
\left|\partial_{t} \log \right| A_{x}| | & \leqslant C_{0} \frac{\log |P|}{|I| \log |I|}  \tag{7.122}\\
\left|A_{y t}\right| & \leqslant C_{0} \tag{7.123}
\end{align*}
$$

Proof. The method is the same as in Proposition 15, but, as now we have to deal with second instead of first order partial derivatives, calculations are more complicated.

When $(P, Q, n) \in \mathcal{R}\left(I_{0}\right), n$ and $|\log | P \|$ are of the same order; as $|I||\log | I|\mid$ is always smaller than $\varepsilon_{0} \log \varepsilon_{0}^{-1}$, the estimates in Proposition 16 are stronger in the present case than the ones in (7.122), (7.123).

We will now assume that $(P, Q, n) \notin \mathcal{R}\left(I_{0}\right)$; the proof of (7.122), (7.123) is by induction on the level of the parameter interval. Let $I^{*} \neq I_{0}$ be the largest parameter interval such that $(P, Q, n) \in \mathcal{R}\left(I^{*}\right)$; as $|I||\log | I\left|\left|\leqslant\left|I^{*}\right|\right| \log \right| I^{*}| |,(7.122)$ for $\mathcal{R}(I)$ follows from (7.122) for $\mathcal{R}\left(I^{*}\right)$. We can therefore assume that $I^{*}=I$ and denote by $\widetilde{I}$ the parent interval.

We apply the structure theorem of Subsection 6.7 and use the same notations as in the proof of Proposition 15: we have an integer $k$, elements $\left(P_{i}, Q_{i}, n_{i}\right)$ in $\mathcal{R}(\widetilde{I})$ for $0 \leqslant i \leqslant k$ and partial compositions of $\left(P^{(i)}, Q^{(i)}, n^{(i)}\right)$ in $\mathcal{R}(I)$ for $0 \leqslant i \leqslant k$. We denote by $(A, B),(\widetilde{A}, \widetilde{B}),\left(A^{*}, B^{*}\right)$ the implicit representations of $\left(P^{(i+1)}, Q^{(i+1)}, n^{(i+1)}\right),\left(P^{(i)}, Q^{(i)}, n^{(i)}\right),\left(P_{i+1}, Q_{i+1}, n_{i+1}\right)$, respectively, for some fixed integer $0 \leqslant i<k$. We know that both $\left(P^{(i)}, Q^{(i)}, n^{(i)}\right),\left(P_{i+1}, Q_{i+1}, n_{i+1}\right)$ satisfy $(* s)$, and $\left(P^{(i)}, Q^{(i)}, n^{(i)}\right)$ also satisfies $(* u)$. We have from formulas (A47), (A49) in Appendix A:

$$
\begin{align*}
\partial_{t} \log \left|A_{x}\right|- & \partial_{t} \log \left|\widetilde{A}_{x}\right|
\end{aligned} \begin{aligned}
& =\partial_{x} \log \left|\widetilde{A}_{x}\right|\left(X_{t}+X_{w} W_{t}\right)+\partial_{t} \log \left|X_{w}\right|+  \tag{7.124}\\
& +W_{t} \partial_{w} \log \left|X_{w}\right|+\partial_{t} \log \left|W_{x}\right| \\
A_{y t}-\widetilde{A}_{y t} & =\widetilde{A}_{x y}\left(X_{t}+X_{w} W_{t}\right)+\widetilde{A}_{x t}\left(X_{y}+X_{w} W_{y}\right)+ \\
& +\widetilde{A}_{x x}\left(X_{t}+X_{w} W_{t}\right)\left(X_{y}+X_{w} W_{y}\right)+ \\
& +\widetilde{A}_{x}\left(X_{y t}+X_{w y} W_{t}+X_{w t} W_{y}+X_{w w} W_{y} W_{t}+X_{w} W_{y t}\right)
\end{align*}
$$

We need to estimate all terms in the right-hand sides of these formulas.

Terms involving $\widetilde{A}$. By bounded distortion, $\left|\partial_{x} \log \right| \widetilde{A}_{x}| |$ is bounded, $\widetilde{A}_{x}, \widetilde{A}_{x y}, \widetilde{A}_{x x}$ are bounded by $C\left|\widetilde{A}_{x}\right|$ and $\widetilde{A}_{x t}$ is part of the induction

$$
\begin{equation*}
\left|\widetilde{A}_{x t}\right|=\left|\widetilde{A}_{x}\right|\left|\partial_{t} \log \right| \widetilde{A}_{x}| | \tag{7.126}
\end{equation*}
$$

Terms involving the first order partial derivatives of $X$. We have dealt earlier with $X_{t}$, $X_{w}, X_{y}$ (cf. (7.92), (7.93), (7.60) and these formulas easily give

$$
\begin{align*}
&\left|X_{t}\right| \leqslant C,  \tag{7.127}\\
& C^{-1} \leqslant\left|X_{w}\right| \leqslant C,  \tag{7.128}\\
&\left|X_{y}\right| \leqslant C\left|\widetilde{B}_{y}\right| . \tag{7.129}
\end{align*}
$$

Terms involving the first order partial derivatives of $W$. We have dealt earlier with $W_{y}=$ $-C_{y} C_{w}^{-1}$ and $W_{t}=-C_{t} C_{w}^{-1}$; the term $C_{w}^{-1}$ is estimated by (7.65), the term $C_{y}$ from formula (7.64) and the term $C_{t}$ from Corollary 11. One obtains

$$
\begin{gather*}
C^{-1} \delta\left(Q^{(i)}, P_{i+1}\right)^{-\frac{1}{2}}\left|\widetilde{B}_{y}\right| \leqslant\left|W_{y}\right| \leqslant C \delta\left(Q^{(i)}, P_{i+1}\right)^{-\frac{1}{2}}\left|\widetilde{B}_{y}\right|,  \tag{7.130}\\
C^{-1} \delta\left(Q^{(i)}, P_{i+1}\right)^{-\frac{1}{2}} \leqslant\left|W_{t}\right| \leqslant C \delta\left(Q^{(i)}, P_{i+1}\right)^{-\frac{1}{2}} . \tag{7.131}
\end{gather*}
$$

Terms involving the second order partial derivatives of $X$. The formulas (A63) of Appendix A express the second order partial derivatives of $X$ in terms of partial derivatives of first and second order of $X_{u}$ (which are bounded), partial derivatives of second order of $\widetilde{B}$ (which are controlled by the distortion or by the induction hypothesis), partial derivatives of first order of $X$ itself (see above), the bounded quantity $\Delta_{0}^{-1}$, and partial derivatives of first order of $\bar{Y}$ (defined in (A31)). These partial derivatives given by (A33) are easy to estimate: we have $\bar{Y}_{w}=\widetilde{B}_{x} X_{w}$, $\bar{Y}_{y}=\widetilde{B}_{y} \Delta_{0}^{-1}, \bar{Y}_{t}=\left(\widetilde{B}_{t}+\widetilde{B}_{x} X_{u, t}\right) \Delta_{0}^{-1}$, hence

$$
\begin{align*}
\left|\bar{Y}_{w}\right| & \leqslant C  \tag{7.132}\\
\left|\bar{Y}_{y}\right| & \leqslant C\left|\widetilde{B}_{y}\right|,  \tag{7.133}\\
\left|\bar{Y}_{t}\right| & \leqslant C \tag{7.134}
\end{align*}
$$

where we have used Proposition 15 to bound $\widetilde{B}_{t}$. Plugging these estimates in the formulas (A63) gives:

$$
\begin{align*}
\left|X_{w w}\right| & \leqslant C  \tag{7.135}\\
\left|X_{w y}\right| & \leqslant C\left|\widetilde{B}_{y}\right|  \tag{7.136}\\
\left|X_{y y}\right| & \leqslant C\left|\widetilde{B}_{y}\right|  \tag{7.137}\\
\left|X_{w t}\right| & \leqslant C\left(1+\left|\widetilde{B}_{x t}\right|\right)  \tag{7.138}\\
\left|X_{y t}\right| & \leqslant C\left|\widetilde{B}_{y}\right|\left(1+\left|\widetilde{B}_{x t}\right|+\partial_{t} \log \left|\widetilde{B}_{y}\right|\right) \tag{7.139}
\end{align*}
$$

Terms involving the second order partial derivatives of $W$. The formulas (A55) of Appendix A express the second order partial derivatives of $W$ in terms of $C_{w}^{-1}$ (controlled by (7.65)),
the partial derivatives of first order of $W$ (see above) and the partial derivatives of second order of $C$. These partial derivatives of second order of $C$ are in turn expressed in formulas (A56)-(A60) in terms of partial derivatives of $\theta$ (which are bounded) and partial derivatives of first and second order of $\bar{X}$ and $\bar{Y}$. The partial derivatives of first order of $\bar{Y}$ have been estimated above, those of $\bar{X}$ satisfy in the same way the inequalities

$$
\begin{align*}
&\left|\bar{X}_{w}\right| \leqslant C  \tag{7.140}\\
&\left|\bar{X}_{x}\right| \leqslant C\left|\widetilde{A}_{x}^{*}\right|  \tag{7.141}\\
&\left|\bar{X}_{t}\right| \leqslant C \tag{7.142}
\end{align*}
$$

The partial derivatives of second order of $\bar{X}$ and $\bar{Y}$ are expressed in formulas (A61), (A62). The formulas (A61) contain partial derivatives of first and second order of $Y$, which are firstly estimated in the same way, through formulas (A29), (A64), as those of $X$ :

$$
\begin{align*}
\left|Y_{t}\right| & \leqslant C  \tag{7.143}\\
C^{-1} & \leqslant\left|Y_{w}\right| \leqslant C  \tag{7.144}\\
\left|Y_{x}\right| & \leqslant C\left|A_{x}^{*}\right|  \tag{7.145}\\
\left|Y_{w w}\right| & \leqslant C  \tag{7.146}\\
\left|Y_{w x}\right| & \leqslant C\left|A_{x}^{*}\right|  \tag{7.147}\\
\left|Y_{x x}\right| & \leqslant C\left|A_{x}^{*}\right|  \tag{7.148}\\
\left|Y_{w t}\right| & \leqslant C\left(1+\left|A_{y t}^{*}\right|\right)  \tag{7.149}\\
\left|Y_{x t}\right| & \leqslant C\left|A_{x}^{*}\right|\left(1+\left|A_{y t}^{*}\right|+\left|\partial_{t} \log \right| A_{x}^{*}| |\right) \tag{7.150}
\end{align*}
$$

We can then estimate the second order partial derivatives of $\bar{X}$ and $\bar{Y}$ :

$$
\begin{align*}
\left|\bar{X}_{w w}\right| & \leqslant C  \tag{7.151}\\
\left|\bar{X}_{w x}\right| & \leqslant C\left|A_{x}^{*}\right|  \tag{7.152}\\
\left|\bar{X}_{w t}\right| & \leqslant C\left(1+\left|A_{y t}^{*}\right|\right.  \tag{7.153}\\
\left|\bar{X}_{x x}\right| & \leqslant C\left|A_{x}^{*}\right|  \tag{7.154}\\
\left|\bar{X}_{x t}\right| & \leqslant C\left|A_{x}^{*}\right|\left(1+\left|A_{y t}^{*}\right|+\left|\partial_{t} \log \right| A_{x}^{*}| |\right) \tag{7.155}
\end{align*}
$$

$$
\begin{align*}
\left|\bar{Y}_{w w}\right| & \leqslant C  \tag{7.156}\\
\left|\bar{Y}_{w y}\right| & \leqslant C\left|\widetilde{B}_{y}\right|  \tag{7.157}\\
\left|\bar{Y}_{w t}\right| & \leqslant C\left(1+\left|\widetilde{B}_{x t}^{*}\right|\right)  \tag{7.158}\\
\left|\bar{Y}_{y y}\right| & \leqslant C\left|\widetilde{B}_{y}\right|  \tag{7.159}\\
\left|\bar{Y}_{y t}\right| & \leqslant C\left|\widetilde{B}_{y}\right|\left(1+\left|\widetilde{B}_{x t}\right|+\left|\partial_{t} \log \right| \widetilde{B}_{y}| |\right) \tag{7.160}
\end{align*}
$$

The next step is to estimate the partial derivatives of second order of $C$ (besides $C_{w w}$ which is
already known to the close to 2 )

$$
\begin{align*}
\left|C_{w x}\right| & \leqslant C\left|A_{x}^{*}\right|  \tag{7.161}\\
\left|C_{w y}\right| & \leqslant C\left|\widetilde{B}_{y}\right| \\
\left|C_{w t}\right| & \leqslant C\left(1+\left|\widetilde{B}_{x t}^{*}\right|+\left|A_{y t}^{*}\right|\right) \\
\left|C_{x x}\right| & \leqslant C\left|A_{x}^{*}\right| \\
\left|C_{x y}\right| & \leqslant C\left|A_{x}^{*}\right|\left|\widetilde{B}_{y}\right| \\
\left|C_{y y}\right| & \leqslant C\left|\widetilde{B}_{y}\right| \\
\left|C_{x t}\right| & \leqslant C\left|A_{x}^{*}\right|\left(1+\left|A_{y t}^{*}\right|+\left|\partial_{t} \log \right| A_{x}^{*}| |\right) \\
\left|C_{y t}\right| & \leqslant C\left|\widetilde{B}_{y}\right|\left(1+\left|\widetilde{B}_{x t}\right|+\left|\partial_{t} \log \right| \widetilde{B}_{y}| |\right)
\end{align*}
$$

Finally, we are able to estimate the partial derivatives of second order of $W$ :

$$
\begin{align*}
& \left|W_{x t}\right| \leqslant C \delta^{-\frac{1}{2}}\left|A_{x}^{*}\right|\left(\delta^{-1}+\delta^{-\frac{1}{2}}\left(\left|\widetilde{B}_{x t}\right|+\left|A_{y t}^{*}\right|\right)+\left|\partial_{t} \log \right| A_{x}^{*}| |\right)  \tag{7.169}\\
& \left|W_{y t}\right| \leqslant C \delta^{-\frac{1}{2}}\left|\widetilde{B}_{y}\right|\left(\delta^{-1}+\delta^{-\frac{1}{2}}\left(\left|\widetilde{B}_{x t}\right|+\left|A_{y t}^{*}\right|\right)+\left|\partial_{t} \log \right| \widetilde{B}_{y}| |\right)
\end{align*}
$$

We have written $\delta$ for $\delta\left(Q^{(i)}, P_{i+1}\right)$.

We are now ready to come back to formulas (7.124), (7.125) above. We get:

$$
\begin{gather*}
\left|\partial_{t} \log \right| A_{x}\left|-\partial_{t} \log \right| \widetilde{A}_{x}| | \leqslant C\left(\delta^{-1}+\delta^{-\frac{1}{2}}\left(\left|\widetilde{B}_{x t}\right|+\left|A_{y t}^{*}\right|\right)+\left|\partial_{t} \log \right| A_{x}^{*}| |\right)  \tag{7.171}\\
\left|A_{y t}-\widetilde{A}_{y t}\right| \leqslant C \delta^{-\frac{1}{2}}\left|\widetilde{A}_{x}\right|\left(1+\left|\widetilde{B}_{y}\right| K\right)
\end{gather*}
$$

with

$$
\begin{equation*}
\left.K=\delta^{-1}+\delta^{-\frac{1}{2}}\left(\left|\widetilde{B}_{x t}\right|+\left|A_{y t}^{*}\right|\right)+\left|\partial_{t} \log \right| \widetilde{A}_{x}\left|+\left|\partial_{t} \log \right| \widetilde{B}_{y}\right| \right\rvert\, \tag{7.173}
\end{equation*}
$$

By the induction hypothesis, we have

$$
\begin{align*}
\left|\widetilde{B}_{x t}\right| & \leqslant C_{0}  \tag{7.174}\\
\left|A_{y t}^{*}\right| & \leqslant C_{0}  \tag{7.175}\\
\left|\partial_{t} \log \right| A_{x}^{*}| | & \leqslant C_{0} \frac{\log \left|A_{x}^{*}\right|}{|\widetilde{I}| \log |\widetilde{I}|},  \tag{7.176}\\
\left|\partial_{t} \log \right| \widetilde{A}_{x}| | & \leqslant C_{0} \frac{\log \left|\widetilde{A}_{x}\right|}{|I| \log |I|},  \tag{7.177}\\
\left|\partial_{t} \log \right| \widetilde{B}_{y}| | & \leqslant C_{0} \frac{\log \left|\widetilde{B}_{y}\right|}{|I| \log |I|} \tag{7.178}
\end{align*}
$$

Here $C_{0}$ is large but independent of $\varepsilon_{0}$. This means that the term $\delta^{-\frac{1}{2}}\left(\left|\widetilde{B}_{x t}\right|+\left|A_{y t}^{*}\right|\right)$ in (7.171) and (7.173) is dominated by $\delta^{-1}$. As $|I|=|\widetilde{I}|^{1+\tau}$, in order to prove (7.122) by induction, we need to have, in view of (7.171):

$$
\begin{equation*}
C|I||\log | I\left|\left|\delta^{-1}+C C_{0}\right| \widetilde{I}\right|^{\tau}|\log | A_{x}^{*}| |+C_{0}|\log | \widetilde{A}_{x}| | \leqslant C_{0}|\log | A_{x}| | \tag{7.179}
\end{equation*}
$$

We have here $\delta \geqslant 2|I|$ from the definition of the transversality relation and, by (3.27):

$$
\begin{equation*}
|\log | A_{x}| | \geqslant|\log | \widetilde{A}_{x}| |+|\log | A_{x}^{*}| |-\frac{1}{2}|\log | I| | \tag{7.180}
\end{equation*}
$$

Therefore, (7.179) will hold as far as

$$
\begin{equation*}
\left(\frac{C_{0}}{2}+C\right)|\log | I\left|\left|\leqslant C_{0}\left(1-C|\widetilde{I}|^{\tau}\right)\right| \log \right| A_{x}^{*}| | \tag{7.181}
\end{equation*}
$$

From (R7), we know that $\left|A_{x}^{*}\right|$ is much smaller than $\delta$. On the other hand, as $Q_{i}$ and $P_{i+1}$ are not $\widetilde{I}$-transverse, $\delta$ cannot be much larger than $\widetilde{I}$. Therefore, we must have

$$
\begin{equation*}
|\log | A_{x}^{*}| | \geqslant \log |\widetilde{I}|=(1+\tau) \log |I| \tag{7.182}
\end{equation*}
$$

from which (7.181) follows if we take $C_{0} \geqslant 3 C$. This completes the proof of the induction step for (7.122).

To do the same for (7.123), we estimate the right-hand side of (7.172). From the proof of Proposition 15 , formula (7.97), we have

$$
\begin{equation*}
\left|\widetilde{A}_{x}\right| \leqslant C^{i}|\widetilde{I}|^{(i+1) \beta-\frac{i}{2}(1+\tau)} \tag{7.183}
\end{equation*}
$$

From (R7), we have

$$
\begin{equation*}
\left|\widetilde{B}_{y}\right| \leqslant C \delta^{(1-\eta)^{-1}} \tag{7.184}
\end{equation*}
$$

The displacement $\delta=\delta\left(Q^{(i)}, P_{i+1}\right)$ satisfies

$$
\begin{equation*}
2|I| \leqslant \delta \leqslant C|\widetilde{I}| \tag{7.185}
\end{equation*}
$$

This gives (as $\beta>1$ )

$$
\begin{align*}
\delta^{-\frac{1}{2}}\left|\widetilde{A}_{x}\right| & \leqslant|\widetilde{I}|^{\beta / 2(i+1)},  \tag{7.186}\\
\delta^{-\frac{3}{2}}\left|\widetilde{A}_{x}\right|\left|\widetilde{B}_{y}\right| & \leqslant|\widetilde{I}|^{\beta / 2(i+1)},  \tag{7.187}\\
\delta^{-\frac{1}{2}}\left|\widetilde{A}_{x}\right|\left|\widetilde{B}_{y}\right|\left|\partial_{t} \log \right| \widetilde{A}_{x}| | & \leqslant|\widetilde{I}|^{\beta / 2(i+1)},  \tag{7.188}\\
\delta^{-\frac{1}{2}}\left|\widetilde{A}_{x}\right|\left|\widetilde{B}_{y}\right|\left|\partial_{t} \log \right| \widetilde{B}_{y}| | & \leqslant|\widetilde{I}|^{\beta / 2(i+1)} \tag{7.189}
\end{align*}
$$

This leads to:

$$
\begin{equation*}
\left|A_{y t}-\widetilde{A}_{y t}\right| \leqslant C|\widetilde{I}|^{\beta / 2(i+1)} \tag{7.190}
\end{equation*}
$$

We can now sum over $i$ and then over the different levels of parameter intervals to obtain (7.123). The proof of Proposition 17 is complete.

Let us see that property (R7) of Subsection 5.4 is a consequence of Proposition 17. Consider $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$ to be elements of $\mathcal{R}(I)$ such that $Q_{0} \subset Q_{u}, P_{1} \subset P_{s}$ and $Q_{0} \pitchfork_{I} P_{1}$ holds.

We can assume that there are no $\widehat{I} \supset I,\left(\widehat{P}_{0}, \widehat{Q}_{0}, \widehat{n}_{0}\right),\left(\widehat{P}_{1}, \widehat{Q}_{1}, \widehat{n}_{1}\right) \in \mathcal{R}(\widehat{I})$ with $Q_{0} \subset \widehat{Q}_{0}, P_{1} \subset \widehat{P}_{1}$, $\left(Q_{0}, P_{1}, I\right) \neq\left(\widehat{Q}_{0}, \widehat{P}_{1}, \widehat{I}\right)$ and $\widehat{Q}_{0} \hbar_{\hat{I}} \widehat{P}_{1}$ : otherwise, as $\left|Q_{0}\right| \leqslant\left|\widehat{Q}_{0}\right|,\left|P_{1}\right| \leqslant\left|\widehat{P}_{1}\right|$ and $\delta\left(Q_{0}, P_{1}\right) \geqslant$ $\delta\left(\widehat{Q}_{0}, \widehat{P}_{1}\right)$, property (R7) for ( $\left.Q_{0}, P_{1}, I\right)$ would be inherited from $\left(\widehat{Q}_{0}, \widehat{P}_{1}, \widehat{I}\right)$.

In view of this maximality property, we claim that $Q_{0}$ must satisfy condition ( $* u$ ), and $P_{1}$ must satisfy condition $(* s)$.

Let us first finish the proof of (R7) assuming the claim to be true. Indeed, we have $\delta\left(Q_{0}, P_{1}\right) \geqslant 2|I|$ for all $t \in I$ by (T1) of Subsection 5.4. If $|\log | Q_{0}| |$ is much larger than $|\log | I|\mid$ for all $t \in I$, we obviously have $\delta\left(Q_{0}, P_{1}\right) \geqslant\left|Q_{0}\right|^{2}$; but if $|\log | Q_{0}| | \leqslant C|\log | I| |$ for some $t \in I$, we obtain from Proposition 17 that

$$
\begin{equation*}
\max _{I}\left|Q_{0}\right| \leqslant C \min _{I}\left|Q_{0}\right| . \tag{7.191}
\end{equation*}
$$

We also know from Proposition 15 that

$$
\begin{equation*}
\max _{I} \delta\left(Q_{0}, P_{1}\right) \leqslant C \min _{I} \delta\left(Q_{0}, P_{1}\right) . \tag{7.192}
\end{equation*}
$$

It then follows from (T2) in Subsection 5.4 that

$$
\begin{equation*}
\delta\left(Q_{0}, P_{1}\right) \geqslant C^{-1}\left|Q_{0}\right|^{1-\eta} . \tag{7.193}
\end{equation*}
$$

for all $t \in I$. We argue with $P_{1}$ in a symmetric way.
Finally, we prove the claim. Let us show, for instance, that $P_{1}$ satisfies condition ( $* s$ ). Let $I^{*}$ be the largest parameter interval such that $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}\left(I^{*}\right)$. If $I^{*}$ is the starting interval $I_{0}, P_{1}$ satisfies condition $(* s)$. Assume therefore that $I^{*} \neq I_{0}$; let $\widetilde{I}^{*}$ be the parent interval, $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ the element of $\mathcal{R}\left(\widetilde{I}^{*}\right)$ such that $\widetilde{P}_{1}$ is the thinnest rectangle containing $P_{1}$. We have to show that $\widetilde{P}_{1}$ is $\widetilde{I}^{*}$-critical. Assume by contradiction that $\widetilde{P}_{1}$ is $\widetilde{I}^{*}$-transverse. Then, there exists an $\widetilde{I}^{*}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $Q_{u}$ such that, for every $\alpha, Q_{\alpha}$ and $\widetilde{P}_{1}$ are $\widetilde{I}^{*}$-separated or $\widetilde{I}^{*}$ transverse. Let $\alpha_{0}$ be such that $Q_{\alpha_{0}}$ and $Q_{0}$ intersect. Then, $Q_{\alpha_{0}}$ and $\widetilde{P}_{1}$ must be $\widetilde{I}^{*}$-transverse. But then $\widetilde{P}_{1}$ and $Q_{0}$ must be $I$-transverse, either from Proposition 2 if $Q_{0} \subset Q_{\alpha_{0}}$, or from concavity (Proposition 4) if $Q_{\alpha_{0}} \subset Q_{0}$; this contradicts the maximality of ( $Q_{0}, P_{1}, I$ ) and proves the claim.

The proof of property (R7) is complete.
The existence and properties of the classes $\mathcal{R}(I)$ are now fully justified. What we do not know at this moment is whether there exists any regular parameter interval at all! This will be the subject of Section 9. Before, we develop in the next section some results that will turn out to be essential in Sections 9 and 10.

## 8 Number of Children and Dimension Estimates

### 8.1 Estimates on the Number of Children

We start with some preliminary results.
Proposition 18. Let $I \subset \widetilde{I}$ be parameter intervals, and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be an element of $\mathcal{R}(\widetilde{I})$. We assume that $\widetilde{Q}$ is $\widetilde{I}$-transverse. Then, any element $(P, Q, n)$ in $\mathcal{R}(I)$ such that $P$ is a child of $\widetilde{P}$ is already an element of $\mathcal{R}(\widetilde{I})$.

Proof. We can assume that $P$ is a non-simple child. Then $(P, Q, n)$ is obtained by parabolic composition in $\mathcal{R}(I)$ of $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ with some $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(I)$. As $\widetilde{Q}$ is $\widetilde{I}$-transverse, there exists an $\widetilde{I}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $P_{s}$ such that each $P_{\alpha}$ is $\widetilde{I}$-separated or $\widetilde{I}$-transverse with $\widetilde{Q}$. Let $\alpha_{0}$ be such that $P_{\alpha_{0}}$ and $P_{1}$ intersect. Then, $\widetilde{Q} \pitchfork_{\widetilde{I}} P_{\alpha_{0}}$ holds, and also $\widetilde{Q} \pitchfork_{I} P_{1}$; if we had $P_{1} \varsubsetneqq P_{\alpha_{0}}$, this would imply that $\widetilde{Q}$ would be $I$-transverse to the parent $\widetilde{P}_{1}$ of $P_{1}$ and $P$ would not be a child of $\widetilde{P}$. Therefore, we must have $P_{\alpha_{0}} \subset P_{1}$. By coherence (Proposition 6), we have that $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}(\widetilde{I})$. By parametric concavity (Proposition 7 ), from $\widetilde{Q} \pitchfork_{I} P_{1}$ and $\widetilde{Q} \pitchfork_{\widetilde{I}} P_{\alpha_{0}}$, we deduce that $\widetilde{Q} \pitchfork_{\tilde{I}} P_{1}$ also holds and $(P, Q, n) \in \mathcal{R}(\widetilde{I})$.

Proposition 19. Let $I$ be a parameter interval, and let $I_{1} \supset I$ be the largest parameter interval such that

$$
\begin{equation*}
\left|I_{1}\right|^{\beta}<\left(\frac{1}{2}|I|\right)^{\frac{1}{1-\eta}} . \tag{8.1}
\end{equation*}
$$

Let $(\widetilde{P}, \widetilde{Q}, \widetilde{n}),(P, Q, n)$ be elements of $\mathcal{R}(I)$ such that $P$ is a non-simple child of $\widetilde{P}$. Let $\left(P_{1}, Q_{1}, n_{1}\right)$, ( $\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}$ ) be the elements of $\mathcal{R}(I)$ such that

$$
(P, Q, n) \in(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \square\left(P_{1}, Q_{1}, n_{1}\right)
$$

and $\widetilde{P}_{1}$ is the parent of $P_{1}$.
Then, $\left(P_{1}, Q_{1}, n_{1}\right)$ belongs to $\mathcal{R}\left(I_{1}\right), \widetilde{P}_{1}$ is $I$-critical, $\widetilde{Q}_{1}$ is $I_{1}$-transverse and we have

$$
\begin{equation*}
2\left|\widetilde{P}_{1}\right|^{1-\eta}>|I| \tag{8.2}
\end{equation*}
$$

for all $t \in I$.
Remark. As parabolic composition is possible, we have $I \neq I_{0}$; then, as $\beta>(1-\eta)^{-1}$, we must have $I_{1} \supsetneqq I$ and $I_{1}$ is $\beta$-regular.

Proof. That $\widetilde{P}_{1}$ is $I$-critical has already been proved in Proposition 5. Also, as $\widetilde{Q} \pitchfork_{I} P_{1}$ holds but $\widetilde{Q} \pitchfork_{I} \widetilde{P}_{1}$ does not hold (because $P$ is a non-simple child of $\widetilde{P}$ ), we deduce (8.2) from Proposition 9 . Then, by definition of $I_{1}$, we have:

$$
\begin{equation*}
\left|\widetilde{P}_{1}\right|>\left|I_{1}\right|^{\beta} \tag{8.3}
\end{equation*}
$$

for all $t \in I$. Let us show that $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ belongs to $\mathcal{R}\left(I_{1}\right)$. Otherwise, there would exist $I_{2} \supset I$, with parent interval $\widetilde{I}_{2} \subset I_{1}$, such that ( $\left.\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ belongs to $\mathcal{R}\left(I_{2}\right)$ but not to $\mathcal{R}\left(\widetilde{I}_{2}\right)$. We apply the inequality (6.74) in the proof of Corollary 9 (Subsection 6.7) to get

$$
\begin{equation*}
\left|\widetilde{P}_{1}\right| \leqslant C\left|\widetilde{I}_{2}\right|^{\beta+\frac{1}{3}} \tag{8.4}
\end{equation*}
$$

in contradiction with (8.3). Therefore, ( $\left.\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ belongs to $\mathcal{R}\left(I_{1}\right)$. As $I_{1} \supsetneqq I, I_{1}$ is $\beta$-regular ; $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ cannot be $I_{1}$-bicritical in view of (8.3); $\widetilde{P}_{1}$ is $I_{1}$-critical and hence $\widetilde{Q}_{1}$ is $I_{1}$-transverse. Proposition 18 then shows that $\left(P_{1}, Q_{1}, n_{1}\right) \in \mathcal{R}\left(I_{1}\right)$.

Corollary 12. Let I be a parameter interval and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be an element of $\mathcal{R}(I)$. The number of $(P, Q, n) \in \mathcal{R}(I)$ such that $P$ is a child of $\widetilde{P}$ is finite.

Proof. We argue by induction on the level of the parameter interval.
If $I$ is the starting interval $I_{0}, \widetilde{P}$ has only simple children and the assertion is obvious. Assume that $I \varsubsetneqq I_{0}$. The number of simple children is finite, and we have to show that the same is true for the number of non-simple children. For every non-simple child $P$ of $\widetilde{P}$, let $I_{1},\left(P_{1}, Q_{1}, n_{1}\right)$, $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}\left(I_{1}\right)$ be as in Proposition 19. By the induction hypothesis, there is for each fixed $\widetilde{P}_{1}$ only a finite number of possibilities for $P_{1}$. On the other hand, in view of relation (8.2), there are obviously only a finite number of possibilities for $\widetilde{P}_{1}$. The induction step is complete, and this completes the proof.

We want to make the finiteness assertion quantitative, and will do that in two distinct ways. In each case, we have to estimate in the proof of Corollary 12 the number of possibilities for $\widetilde{P}_{1}$, and the number of possibilities for $P_{1}$ once $\widetilde{P}_{1}$ is fixed.

Proposition 20. Let $I$ be a parameter interval, and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be an element of $\mathcal{R}(I)$. The number of $(P, Q, n) \in \mathcal{R}(I)$ such that $P$ is a child of $\widetilde{P}$ is at most $|I|^{-c \eta}$, where $c$ is a constant depending only on $\beta$.

Proof. We argue again by induction on the level of $I$, following the proof of Corollary 12. When $I=I_{0}$, the number of (simple) children is at most the number of rectangles in the Markov partition, which is much smaller than $\varepsilon_{0}^{-c \eta}$ when $\varepsilon_{0}$ is small enough.

When $I \neq I_{0}$, the number of possibilities for $P_{1}$ when $\widetilde{P}_{1}$ is fixed in at most $\left|I_{1}\right|^{-c \eta}$ by the induction hypothesis. We have to estimate the number of possibilities for $\widetilde{P}_{1}$. We know that $\widetilde{Q} \pitchfork_{I} \widetilde{P}_{1}$ does not hold, but $\widetilde{Q} \pitchfork_{I} P_{1}$ holds.

As $\widetilde{P}_{1}$ is $I$-critical, it satisfies condition $(* s)$, defined just before Proposition 15 in Subsection 7.5. We have from (8.2) and Proposition 17

$$
\begin{equation*}
\max _{I}\left|\widetilde{P}_{1}\right| \leqslant C \min _{I}\left|\widetilde{P}_{1}\right| . \tag{8.5}
\end{equation*}
$$

Lemma 3. We have for all $t \in I$

$$
\begin{equation*}
\delta\left(\widetilde{Q}, \widetilde{P}_{1}\right) \leqslant C\left|\widetilde{P}_{1}\right|^{1-\eta} \tag{8.6}
\end{equation*}
$$

Proof. Let $(\widehat{P}, \widehat{Q}, \widehat{n})$ be the element of $\mathcal{R}(I)$ with smallest $\widehat{n}$ such that $\widetilde{Q} \subset \widehat{Q}$ and $\widehat{Q} \pitchfork_{I} P_{1}$ holds. The parent $Q^{*}$ of $\widehat{Q}$ is $I$-critical: otherwise, from an $I$-decomposition of $P_{s}$, one would find ( $P_{\alpha}, Q_{\alpha}, n_{\alpha}$ ) with $Q^{*} \pitchfork_{I} P_{\alpha}$ and $P_{\alpha} \cap P_{1} \neq \emptyset$; one would conclude that $Q^{*} \pitchfork_{I} P_{1}$ from Proposition 2 (if $P_{1} \subset P_{\alpha}$ ) or Proposition 4 (if $P_{\alpha} \subset P_{1}$ ). Thus, $\widehat{Q}$ satisfies condition $(* u)$. From Corollary 11 and Proposition 17, we get

$$
\begin{gather*}
\max _{I}|\widehat{Q}| \leqslant C \min _{I}|\widehat{Q}|  \tag{8.7}\\
\max _{I} \delta_{L R}\left(\widehat{Q}, \widetilde{P}_{1}\right)-\min _{I} \delta_{L R}\left(\widehat{Q}, \widetilde{P}_{1}\right) \leqslant 2|I| .
\end{gather*}
$$

As $\widetilde{Q} \pitchfork_{I} \widetilde{P}_{1}$ does not hold, $\widehat{Q} \overleftarrow{\pitchfork}_{I} \widetilde{P}_{1}$ does not hold either and at least one of the following three inequalities must hold:

$$
\begin{align*}
\delta_{L R}\left(\widehat{Q}, \widetilde{P}_{1}\right) & <2|I| \text { for some } t_{0} \in I  \tag{8.9}\\
\delta_{R}\left(\widehat{Q}, \widetilde{P}_{1}\right) & <2|\widehat{Q}|^{1-\eta} \text { for all } t \in I  \tag{8.10}\\
\delta_{L}\left(\widehat{Q}, \widetilde{P}_{1}\right) & <2\left|\widetilde{P}_{1}\right|^{1-\eta} \text { for all } t \in I \tag{8.11}
\end{align*}
$$

By Proposition 10, as $\widehat{Q} \pitchfork_{I} P_{1}$ holds but $\widehat{Q} \pitchfork_{I} \widetilde{P}_{1}$ does not hold, we have, for some $t_{1} \in I$ :

$$
\begin{equation*}
\left|\widetilde{P}_{1}\right|>\frac{1}{2}|\widehat{Q}| \tag{8.12}
\end{equation*}
$$

We can now prove (8.6).

If (8.11) holds, we have, for all $t \in I$ :

$$
\begin{equation*}
\delta\left(\widetilde{Q}, \widetilde{P}_{1}\right) \leqslant \delta_{L}\left(\widetilde{Q}, \widetilde{P}_{1}\right) \leqslant \delta_{L}\left(\widehat{Q}, \widetilde{P}_{1}\right)<2\left|\widetilde{P}_{1}\right|^{1-\eta} \tag{8.13}
\end{equation*}
$$

If (8.12) holds, we have from (5.12), (8.5), (8.7), (8.12):

$$
\begin{align*}
\delta\left(\widetilde{Q}, \widetilde{P}_{1}\right) & \leqslant \delta_{L R}\left(\widehat{Q}, \widetilde{P}_{1}\right)  \tag{8.14}\\
& \leqslant \delta_{R}\left(\widehat{Q}, \widetilde{P}_{1}\right)+c|\widehat{Q}| \\
& \leqslant 3|\widehat{Q}|^{1-\eta} \leqslant C\left|\widetilde{P}_{1}\right|^{1-\eta}
\end{align*}
$$

for all $t \in I$.

Finally, if (8.9) holds, we have from (8.5), (8.8), (8.2), for all $t \in I$ :

$$
\begin{align*}
\delta\left(\widetilde{Q}, \widetilde{P}_{1}\right) & \leqslant \delta_{L R}\left(\widehat{Q}, \widetilde{P}_{1}\right)  \tag{8.15}\\
& \leqslant 4|I| \leqslant 8\left|\widetilde{P}_{1}\right|^{1-\eta}
\end{align*}
$$

We are now able to estimate the number of possibilities for $\widetilde{P}_{1}$ and show that this number is at most

$$
\begin{equation*}
C|I|^{-\frac{\eta}{1-\eta}} \tag{8.16}
\end{equation*}
$$

This indeed follows from (8.6), (8.2) and the fact that if two distinct $\widetilde{P}_{1}$ are not disjoint, the ratio of their widths is bounded away from 1 (so the $\widetilde{P}_{1}$ at a given scale are disjoint; one then sums over scales). The total number of children is thus bounded by

$$
\begin{equation*}
C+2 C|I|^{-\frac{\eta}{1-\eta}}\left|I_{1}\right|^{-c \eta} \tag{8.17}
\end{equation*}
$$

where $I_{1}$ was the largest parameter interval satisfying

$$
\begin{equation*}
\left|I_{1}\right|^{\beta}<\left(\frac{1}{2}|I|\right)^{\frac{1}{1-\eta}} \tag{8.18}
\end{equation*}
$$

If $|I|>2 \varepsilon_{0}^{\beta(1-\eta)}$, we have $I_{1}=I_{0}$; in this case, the term $\left|I_{1}\right|^{-c \eta}$ in (8.17) is unnecessary because $\widetilde{P}_{1}$ has only simple children. If $|I| \leqslant 2 \varepsilon_{0}^{\beta(1-\eta)}$, we have

$$
\begin{equation*}
\left|I_{1}\right|^{\beta(1+\tau)^{-1}} \geqslant\left(\frac{1}{2}|I|\right)^{\frac{1}{1-\eta}} \tag{8.19}
\end{equation*}
$$

and the term in (8.17) is bounded by $|I|^{-c \eta}$ provided that

$$
\begin{equation*}
c \eta>\frac{\eta}{1-\eta}+c \eta \frac{1+\tau}{1-\eta} \beta^{-1} . \tag{8.20}
\end{equation*}
$$

As $\eta, \tau$ are very small, any choice of $c>\frac{\beta}{\beta-1}$ yields (8.20). Then, as $c>1$, such a choice is also convenient when $|I|>2 \varepsilon_{0}^{\beta(1-\eta)}$, and this concludes the proof of Lemma 3 .

In Proposition 20, we have estimated the total number of children in terms of the level of the parameter interval.

When $\widetilde{Q}$ is $I$-transverse, Proposition 18 guarantees that there will not be any new child of $\widetilde{Q}$ when we consider parameter intervals $\widehat{I} \subset I$ of higher level.

The same is true when for all $t \in I$

$$
\begin{equation*}
|\widetilde{Q}|^{1-\eta} \geqslant C|I| . \tag{8.21}
\end{equation*}
$$

with some large enough constant $C$. Indeed, let $I^{*}$ be a parameter interval strictly smaller than $I$, with parent interval $\widetilde{I}^{*}$, and let $(P, Q, n)$ be an element of $\mathcal{R}\left(I^{*}\right)$ such that $P$ is a non-simple child of $\widetilde{P}$. Let $\left(P_{1}, Q_{1}, n_{1}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ be as in the previous propositions. By Proposition 19 , the element ( $P_{1}, Q_{1}, n_{1}$ ) belongs to $\mathcal{R}\left(\widetilde{I}^{*}\right)$. It is then easy to deduce from $\widetilde{Q} \pitchfork_{I^{*}} P_{1}$ and (8.21), using (R7) and $\widetilde{I^{*}} \subset I$, that $\widetilde{Q} \pitchfork_{\tilde{I}^{*}} P_{1}$ also holds. This proves by induction that $(P, Q, n)$ belongs to $\mathcal{R}(I)$.

In the next proposition, we are interested, not in the total number of children, but in the number of children of a given width. The estimate is independent on the level of the parameter interval.

Proposition 21. Let $I$ be a parameter interval, and let $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ be an element of $\mathcal{R}(I)$. For any $\varepsilon>0$, the number of elements $(P, Q, n) \in \mathcal{R}(I)$ such that $P$ is a non-simple child of $\widetilde{P}$ satisfying $|P| \geqslant \varepsilon|\widetilde{P}|$ for some $t \in I$, is at most $\varepsilon^{-c^{\prime} \eta}$, where $c^{\prime}$ is a constant depending only on the regularity parameter $\beta$.

Proof. Let $\varepsilon>0$, and let $(P, Q, n)$ be an element of $\mathcal{R}(I)$ such that $P$ is a non-simple child of $\widetilde{P}$. We assume, for some $t_{0} \in I$, that:

$$
\begin{equation*}
|P| \geqslant \varepsilon|\widetilde{P}| . \tag{8.22}
\end{equation*}
$$

Let $\left(P_{1}, Q_{1}, n_{1}\right),\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}(I)$ be as in Proposition 19. From (3.27), we have, for all $t \in I$ :

$$
\begin{equation*}
|P| \leqslant C|\widetilde{P}|\left|P_{1}\right| \delta\left(\widetilde{Q}, P_{1}\right)^{-\frac{1}{2}} \tag{8.23}
\end{equation*}
$$

Property (R7) guarantees that, for all $t \in I$

$$
\begin{equation*}
\delta\left(\widetilde{Q}, P_{1}\right) \geqslant C^{-1}\left|P_{1}\right|^{1-\eta} \tag{8.24}
\end{equation*}
$$

Combining (8.22), (8.23), (8.24), we have, for some $t_{0} \in I$

$$
\begin{equation*}
\delta\left(\widetilde{Q}, P_{1}\right) \geqslant C^{-1} \varepsilon^{2 \frac{1-\eta}{1+\eta}} \tag{8.25}
\end{equation*}
$$

As we always have

$$
\begin{equation*}
\delta\left(\widetilde{Q}, P_{1}\right) \leqslant C \varepsilon_{0} \tag{8.26}
\end{equation*}
$$

there is no non-simple child satisfying (8.22) unless $\varepsilon<\varepsilon_{0}^{\frac{1}{2}}$; we will assume that this holds in the sequel.

From Lemma 3 above, we have, for all $t \in I$ :

$$
\begin{equation*}
\delta\left(\widetilde{Q}, \widetilde{P}_{1}\right) \leqslant C\left|\widetilde{P}_{1}\right|^{1-\eta} \tag{8.27}
\end{equation*}
$$

and thus, from (5.11), also

$$
\begin{align*}
\delta\left(\widetilde{Q}, P_{1}\right) & \leqslant \delta_{L}\left(\widetilde{Q}, \widetilde{P}_{1}\right)  \tag{8.28}\\
& \leqslant \delta\left(\widetilde{Q}, \widetilde{P}_{1}\right)+C\left|\widetilde{P}_{1}\right| \\
& \leqslant C\left|\widetilde{P}_{1}\right|^{1-\eta} .
\end{align*}
$$

Combining (8.25) and (8.28), we get, for some $t_{0} \in I$

$$
\begin{equation*}
\left|\widetilde{P}_{1}\right| \geqslant C^{-1} \varepsilon^{\frac{2}{1+\eta}} \tag{8.29}
\end{equation*}
$$

an inequality which actually holds for all $t \in I$ in view of (8.5).

As in the proof of Proposition 20, the number of $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ for which both (8.28), (8.29) hold is easily seen to be at most

$$
\begin{equation*}
C \varepsilon^{-\frac{2 \eta}{1+\eta}} \tag{8.30}
\end{equation*}
$$

On the other hand, let $\widehat{I}=I$ if $|I| \geqslant \varepsilon^{2(\beta+1 / 3)^{-1}}$; otherwise, let $\widehat{I}$ be the largest parameter interval which contains $I$ and satisfies $|\widehat{I}|<\varepsilon^{2(\beta+1 / 3)^{-1}}$. As $\widetilde{P}_{1}$ is $I$-critical, the same argument as in the proof of Corollary 9 in Subsection 6.7 shows, from (8.29), that we must have

$$
\begin{equation*}
\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) \in \mathcal{R}(\widehat{I}) \tag{8.31}
\end{equation*}
$$

Now, $\widetilde{P}_{1}$ is $\widehat{I}$-critical. If $\widehat{I} \neq I$, we have, from (8.29) and the definition of $\widehat{I}$, that

$$
\begin{equation*}
\left|\widetilde{P}_{1}\right|>|\widehat{I}|^{\beta}, \quad \text { for all } t \in I \tag{8.32}
\end{equation*}
$$

Therefore, $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$ cannot be $\widehat{I}$-bicritical, $Q_{1}$ is $\widehat{I}$-transverse and we conclude from Proposition 18 that $\left(P_{1}, Q_{1}, n_{1}\right)$ also belongs to $\mathcal{R}(\widehat{I})$. The same is also obviously true when $\widehat{I}=I$. We apply Proposition 20 : for each fixed $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$, the number of children $P_{1}$ is at most $|\widehat{I}|^{-c \eta}$. But we have

$$
\begin{array}{ll}
|\widehat{I}|=|I| \geqslant \varepsilon^{2(\beta+1 / 3)^{-1}} & \text { if }|I| \geqslant \varepsilon^{2(\beta+1 / 3)^{-1}} \\
|\widehat{I}|=\varepsilon_{0} & \text { if } \varepsilon^{2(\beta+1 / 3)^{-1}>\varepsilon_{0}} \\
|\widehat{I}| \geqslant \varepsilon^{2(1+\tau)(\beta+1 / 3)^{-1}} & \text { if } \varepsilon_{0} \geqslant \varepsilon^{2(\beta+1 / 3)^{-1}}>|I| \tag{8.35}
\end{array}
$$

In all cases, this gives

$$
\begin{equation*}
|\widehat{I}|^{-c \eta} \geqslant \varepsilon^{-c_{0} \eta} \tag{8.36}
\end{equation*}
$$

Combining Proposition 20 with the previous estimate in (8.30) for the number of possibilities for $\widetilde{P}_{1}$ gives therefore the required estimate.

### 8.2 A Dimension Estimate

The goal of this subsection is to obtain a bound on the number of elements $(P, Q, n)$ in $\mathcal{R}(I)$ with width $|P|$ bounded from below. This is a first step towards estimating the transverse dimension of the stable set $W^{s}(\Lambda)$, which is necessary in order to achieve our parameter selection in Section 9 .

Let $I$ be a parameter interval, and let $\left(P^{*}, Q^{*}, n^{*}\right)$ be an element of $\mathcal{R}(I)$. We introduce, in the spirit of Laplace, Dirichlet and Poincaré, the series

$$
\begin{equation*}
\Theta\left(P^{*}, I, s\right)=\sum|P|^{s} \tag{8.37}
\end{equation*}
$$

where the sum runs over elements $(P, Q, n) \in \mathcal{R}(I)$ such that $P \subset P^{*}$. Here $s$ is a complex variable and the series is at first a formal object, but we will soon see that it is uniformly convergent in a
half-plane $\left\{\operatorname{Re} s>\sigma_{0}\right\}$. The goal of this subjection is to obtain a nice estimate for $\sigma_{0}$ and for $\Theta$ in this half-plane.

The width $|P|$ and therefore also the series $\Theta$, depend on the parameter $t \in I$; but the estimate that we will get is uniform with respect to the parameter. The dependence of the estimate on $P^{*}$ is also quite straightforward, through a simple scaling factor.

Let us recall that we denote by $d_{s}^{0}$ the transverse Hausdorff dimension of the stable foliation $W^{s}(K)$ of the initial horseshoes $K$ for the value 0 of the parameter. It is well-known that this transverse Hausdorff dimension depends smoothly on the parameter, and it controls in a precise way the number of cylinders (for the Markov partition) of a given size; more precisely, as these cylinders correspond exactly to the elements of $\mathcal{R}\left(I_{0}\right)$, we know that, for all $t \in I_{0}$ and all $\varepsilon>0$ the number of $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$ such that $|P| \geqslant \varepsilon$ is at most

$$
\begin{equation*}
C \varepsilon^{-\left(d_{s}^{0}+C \varepsilon_{0}\right)} . \tag{8.38}
\end{equation*}
$$

This shows that for $\Theta\left(P^{*}, I_{0} ; s\right)$ we could take $\sigma_{0}=d_{s}^{0}+C \varepsilon_{0}$. For smaller parameter intervals, we have to allow a slightly larger margin with relation to the initial value $d_{s}^{0}$.

Proposition 22. The series $\Theta\left(P^{*}, I, s\right)$ is uniformly convergent in the half-plane $\left\{\right.$ Re $\left.s \geqslant d_{s}^{0}+\varepsilon_{0}^{\frac{1}{3} d_{s}^{0}}\right\}$. When $\left(P^{*}, Q^{*}, n^{*}\right) \in \mathcal{R}\left(I_{0}\right)$, we have for Re $s \geqslant d_{s}^{0}+\varepsilon_{0}^{\frac{1}{5}} d_{s}^{0}$

$$
\left|\Theta\left(P^{*}, I ; s\right)-\Theta\left(P^{*}, I_{0} ; s\right)\right| \geqslant\left|P^{*}\right|^{s} \varepsilon_{0}^{\frac{1}{20} d_{s}^{0}}
$$

Proof. Let $(P, Q, n)$ be an element of $\mathcal{R}(I)$ with $P \subset P^{*}$. Consider the intermediary rectangles

$$
\begin{equation*}
P^{*}=P(0) \subset P(1) \subset \cdots \subset P(\ell)=P \tag{8.39}
\end{equation*}
$$

with $P(i)$ the parent of $P(i+1)$. Let

$$
\begin{equation*}
\ell_{0}<\ell_{1}<\cdots<\ell_{k-1} \tag{8.40}
\end{equation*}
$$

be the indices such that $P\left(\ell_{j}+1\right)$ is a non-simple child of $P\left(\ell_{j}\right)$.

We also define for $0 \leqslant j \leqslant k$ elements $\left(P^{(j)}, Q^{(j)}, n^{(j)}\right) \in \mathcal{R}\left(I_{0}\right)$ by the following properties

$$
\begin{gather*}
\left(P\left(\ell_{0}\right), Q\left(\ell_{0}\right), n\left(\ell_{0}\right)\right)=\left(P(0), Q(0), n(0) *\left(P^{(0)}, Q^{(0)}, n^{(0)}\right)\right)  \tag{8.41}\\
\left(P\left(\ell_{j}\right), Q\left(\ell_{j}\right), n\left(\ell_{j}\right)\right)=\left(P\left(\ell_{j-1}+1\right), Q\left(\ell_{j-1}+1\right), n\left(\ell_{j-1}+1\right) *\left(P^{(j)}, Q^{(j)}, n^{(j)}\right)\right.  \tag{8.42}\\
(P, Q, n)=\left(P\left(\ell_{k-1}+1\right), Q\left(\ell_{k-1}+1\right), n\left(\ell_{k-1}+1\right) *\left(P^{(k)}, Q^{(k)}, n^{(k)}\right)\right. \tag{8.43}
\end{gather*}
$$

We now estimate the widths from (3.10):

$$
\begin{align*}
\left|P\left(\ell_{0}\right)\right| & \leqslant C\left|P^{*}\right|\left|P^{(0)}\right|  \tag{8.44}\\
\left|P\left(\ell_{j}\right)\right| & \leqslant C\left|P\left(\ell_{j-1}+1\right)\right|\left|P^{(j)}\right|  \tag{8.45}\\
|P| & \leqslant C\left|P\left(\ell_{k-1}+1\right)\right|\left|P^{(k)}\right| . \tag{8.46}
\end{align*}
$$

From (3.27) and property (R7), we also have:

$$
\begin{equation*}
\left|P\left(\ell_{j}+1\right)\right|<\varepsilon_{0}^{\frac{1}{2}}\left|P\left(\ell_{j}\right)\right| \tag{8.47}
\end{equation*}
$$

Define $m_{j}$ for $0 \leqslant j<k$ to be the largest integer such that, for all $t \in I$

$$
\begin{equation*}
\left|P\left(\ell_{j}+1\right)\right| \leqslant 2^{-m_{j}} \varepsilon_{0}^{\frac{1}{2}}\left|P\left(\ell_{j}\right)\right| \tag{8.48}
\end{equation*}
$$

From Proposition 21, for each fixed $P\left(\ell_{j}\right)$, the number of non-simple children $P\left(\ell_{j}+1\right)$ satisfying (8.48) is at most

$$
\begin{equation*}
\left(2^{m_{j}+1} \varepsilon_{0}^{-\frac{1}{2}}\right)^{c^{\prime} \eta} \tag{8.49}
\end{equation*}
$$

Combining (8.44), (8.45), (8.46) and (8.48), we also have

$$
\begin{equation*}
|P| \leqslant C^{k+1}\left|P^{*}\right|\left(\prod_{0}^{k}\left|P^{(j)}\right|\right) \varepsilon_{0}^{\frac{k}{2}} 2^{-\Sigma m_{j}} \tag{8.50}
\end{equation*}
$$

We will take this to the power $s$ and sum over $P$. We introduce (corresponding to the term $\left|P^{(j)}\right|^{s}$ )

$$
\begin{align*}
\Theta_{0}(s): & =\sum_{(\widehat{P}, \widehat{Q}, \widehat{n}) \in \mathcal{R}\left(I_{0}\right)}|\widehat{P}|^{s}  \tag{8.51}\\
& =\sum_{a} \Theta\left(R_{a}, I_{0}, s\right)
\end{align*}
$$

and also

$$
\begin{equation*}
\theta(s):=\sum_{m \geqslant 0}\left(C \varepsilon_{0}^{\frac{1}{2}} 2^{-m}\right)^{s}\left(2^{m+1} \varepsilon_{0}^{-\frac{1}{2}}\right)^{c^{\prime} \eta} \tag{8.52}
\end{equation*}
$$

The function $\Theta_{0}$ is controlled by (8.38), while $\theta$ satisfies

$$
\begin{equation*}
\theta(s)=2^{c^{\prime} \eta} C^{s} \varepsilon_{0}^{\frac{1}{2}\left(s-c^{\prime} \eta\right)}\left(1-2^{-\left(s-c^{\prime} \eta\right)}\right)^{-1} \tag{8.53}
\end{equation*}
$$

and therefore, for $C^{-1}<s<C$ :

$$
\begin{equation*}
C^{-1} \varepsilon_{0}^{\frac{1}{2}\left(s-c^{\prime} \eta\right)} \leqslant \theta(s) \leqslant C \varepsilon_{0}^{\frac{1}{2}\left(s-c^{\prime} \eta\right)} \tag{8.54}
\end{equation*}
$$

From (8.38), we have, for $s>d_{s}^{0}+C \varepsilon_{0}$ :

$$
\begin{equation*}
\Theta_{0}(s) \leqslant C\left(s-d_{s}^{0}-C \varepsilon_{0}\right)^{-1} \tag{8.55}
\end{equation*}
$$

In particular, for $s \geqslant d_{s}^{0}+\varepsilon_{0}^{1 / 3 d_{s}^{0}}$, we have

$$
\begin{array}{r}
\Theta_{0}(s) \leqslant C \varepsilon_{0}^{-\frac{1}{3} d_{s}^{0}} \\
\Theta_{0}(s) \theta(s) \leqslant C \varepsilon_{0}^{\frac{1}{10} d_{s}^{0}} \tag{8.57}
\end{array}
$$

But, from (8.50), we have for real $s$

$$
\begin{equation*}
\Theta\left(P^{*}, I, s\right) \leqslant C^{s}\left|P^{*}\right|^{s} \sum_{k \geqslant 0} \Theta_{0}^{k+1}(s) \theta^{k}(s), \tag{8.58}
\end{equation*}
$$

and therefore we deduce from (8.57) that the series defining $\Theta$ is uniformly convergent in the half plane $\left\{\operatorname{Re} s \geqslant d_{s}^{0}+\varepsilon_{0}^{\frac{1}{3} d_{s}^{0}}\right\}$.

Assume now that $\left(P^{*}, Q^{*}, n^{*}\right) \in \mathcal{R}\left(I_{0}\right) ;$ the difference $\Theta\left(P^{*}, I, s\right)-\Theta\left(P^{*}, I_{0}, s\right)$ consists of the sum of $|P|^{s}$ over those $P$ for which $k>0$. We get, for real $s$

$$
\begin{equation*}
\Theta\left(P^{*}, I, s\right)-\Theta\left(P^{*}, I_{0}, s\right) \leqslant C^{s}\left|P^{*}\right|^{s} \sum_{k>0} \Theta_{0}^{k+1}(s) \theta^{k}(s) . \tag{8.59}
\end{equation*}
$$

For $s>d_{s}^{0}+\varepsilon_{0}^{\frac{1}{5} d_{s}^{0}}$, we have, from (8.55), (8.54):

$$
\begin{align*}
\Theta_{0}(s) & \leqslant C \varepsilon_{0}^{-\frac{1}{5} d_{s}^{0}}  \tag{8.60}\\
\Theta_{0}^{2} \theta(s) & \leqslant C \varepsilon_{0}^{\frac{1}{15} d_{s}^{0}} \tag{8.61}
\end{align*}
$$

which gives the second part of the proposition.

### 8.3 Transfer to Parameter Space

8.3.1 Our goal in this subsection will be to prove the following result, which expresses a transfer of the dimension estimate of Subsection 8.2 to parameter space.

Proposition 23. Let $\widetilde{I}$ be a regular parameter interval. Let $\left(P^{*}, Q^{*}, n^{*}\right)$ be an element of $\mathcal{R}(\widetilde{I})$ such that $Q^{*}$ is $\widetilde{I}$-critical and

$$
\begin{equation*}
\left|Q^{*}\right| \leqslant \frac{1}{2}|\widetilde{I}|^{(1+\tau)(1-\eta)^{-1}} \tag{8.62}
\end{equation*}
$$

for all $t \in \widetilde{I}$. The, the number of candidates $I \subset \widetilde{I}$ of the next level, such that $Q^{*}$ is $I$-critical, is at most $|\widetilde{I}|^{-\tau d_{s}^{+}}$, where $d_{s}^{+}=d_{s}^{0}+C \eta \tau^{-1}$ can be made arbitrarily close to $d_{s}^{0}$.

Recall that the total number of candidates is $|\widetilde{I}|^{-\tau}$. Proposition 23 is the key estimate that will allow us in Section 9 to proceed with the selection process for parameters. The rest of the section is devoted to the proof of Proposition 23.
8.3.2 We start with some general observations, that could have been made much earlier, but are only useful now.

Let $I$ be a parameter interval. Let $(P, Q, n),\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{1}, Q_{1}, n_{1}\right)$, be elements of $\mathcal{R}(I)$ such that $Q \subset Q_{u}, P_{0} \subset P_{s}, P_{1} \subset P_{s}$ and $P_{0} \cap P_{1}=\emptyset$. Associated to the pair $\left(Q, P_{0}\right)\left(\operatorname{resp} .\left(Q, P_{1}\right)\right)$, we have defined in Subsection 3.5 a function $C_{0}(w, y, x)$ (resp. $C_{1}(w, y, x)$ ) which is the basis for the definition of $\delta\left(Q, P_{0}\right), \delta_{L R}\left(Q, P_{0}\right), \cdots\left(\right.$ resp. $\left.\delta\left(Q, P_{1}\right), \delta_{L R}\left(Q, P_{1}\right), \cdots\right)$. It follows immediately from the monotonicity of the function $\theta$ that, since $P_{0}$ and $P_{1}$ are disjoint, we have either

$$
\begin{align*}
& C_{0}\left(w, y, x_{0}\right)>C_{1}\left(w, y, x_{1}\right) \text { for all } w, y, x_{0}, x_{1}, t, \text { or }  \tag{8.63}\\
& C_{0}\left(w, y, x_{0}\right)<C_{1}\left(w, y, x_{1}\right) \text { for all } w, y, x_{0}, x_{1}, t \tag{8.64}
\end{align*}
$$

In the first case, we have, from the definitions in Subsection 5.4:

$$
\begin{align*}
\delta_{R}\left(Q, P_{0}\right) & <\delta\left(Q, P_{1}\right)  \tag{8.65}\\
\delta_{L R}\left(Q, P_{0}\right) & <\delta_{L}\left(Q, P_{1}\right)
\end{align*}
$$

while in the second the same inequalities hold exchanging $P_{0}$ and $P_{1}$.
We see in particular that if (8.63) holds and $Q, P_{1}$ are $I$-separated, then $Q, P_{0}$ are also $I$-separated.

On the other hand, we will prove (see figure 8).


Figure 8

Proposition 24. Assume that (8.63) holds and that $\left|P_{1}\right|^{1-\eta} \leqslant|I|$ for some $t \in I$. Then, if $Q$ and $P_{0}$ are $I$-transverse, $Q$ and $P_{1}$ are also $I$-transverse.

Proof. Let $\widetilde{I} \supset I$ and $(\widetilde{P}, \widetilde{Q}, \widetilde{n}),\left(\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right) \in \mathcal{R}(\widetilde{I})$ be such that $\widetilde{Q} \supset Q, \widetilde{P}_{0} \supset P_{0}$ and $\widetilde{Q} \widetilde{\Pi}_{\widetilde{I}} \widetilde{P}_{0}$ holds. If $P_{1} \subset \widetilde{P}_{0}$, we immediately conclude that $Q$ and $P_{1}$ are $I$-transverse. We assume, therefore, that $P_{1} \cap \widetilde{P}_{0}=\emptyset$; replacing $\left(P_{0}, Q_{0}, n_{0}\right)$ by ( $\left.\widetilde{P}_{0}, \widetilde{Q}_{0}, \widetilde{n}_{0}\right)$, and ( $P, Q, n$ ) by ( $\left.\widetilde{P}, \widetilde{Q}, \widetilde{n}\right)$, we can also assume that $(P, Q, n),\left(P_{0}, Q_{0}, n_{0}\right) \in \mathcal{R}(\widetilde{I})$ and $Q \widetilde{\Pi}_{\widetilde{I}} P_{0}$ holds.

Let $\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right)$, be the element of $\mathcal{R}(\widetilde{I})$ with $P_{1} \subset \widetilde{P}_{1}$ and smallest $\widetilde{P}_{1}$. We will prove that $Q$ and $\widetilde{P}_{1}$ are $\widetilde{I}$-transverse. We have, for all $t \in \widetilde{I}$, that

$$
\begin{equation*}
\delta_{L R}\left(Q, \widetilde{P}_{1}\right)>\delta_{L R}\left(Q, P_{0}\right) \geqslant 2|\widetilde{I}|, \tag{8.66}
\end{equation*}
$$

and also for some $t_{0} \in \widetilde{I}$,

$$
\begin{equation*}
\delta_{R}\left(Q, \widetilde{P}_{1}\right)>\delta_{R}\left(Q, P_{0}\right) \geqslant 2|Q|^{1-\eta} . \tag{8.67}
\end{equation*}
$$

Therefore, assuming that $Q \mathrm{C}_{\widetilde{I}} \widetilde{P}_{1}$ does not hold, we must have, for all $t \in \widetilde{I}$, that

$$
\begin{equation*}
\delta_{L}\left(Q, \widetilde{P}_{1}\right)<2\left|\widetilde{P}_{1}\right|^{1-\eta} . \tag{8.68}
\end{equation*}
$$

We cannot have in this case $\widetilde{P}_{1}=P_{1}$, because, for all $t \in I$,

$$
\begin{equation*}
\delta_{L}\left(Q, P_{1}\right)>\delta_{L R}\left(Q, P_{0}\right) \geqslant 2|\widetilde{I}|, \tag{8.69}
\end{equation*}
$$

and (8.68), (8.69) together would contradict the hypothesis of the proposition. Therefore, $\widetilde{P}_{1}$ strictly contains $P_{1}$ and $\widetilde{I}$ strictly contains $I$. But, then, applying the structure theorem of Subsection 6.7 to the child of $\widetilde{P}_{1}$ which contains $P_{1}$, we obtain that $\widetilde{Q}_{1}$ is $\widetilde{I}$-critical. As $\widetilde{I}$ is $\beta$-regular, it then follows from (8.68), (8.69) and (5.14) that $\widetilde{P}_{1}$ is $\widetilde{I}$-transverse. This implies that there exists $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in$ $\mathcal{R}(\widetilde{I})$ with $Q \cap Q^{\prime} \neq \emptyset$ such that $Q^{\prime} \pitchfork_{\widetilde{I}} \widetilde{P}_{1}$ holds. If $Q \subset Q^{\prime}$, it follows that $Q \pitchfork_{\widetilde{I}} \widetilde{P}_{1}$ holds. When $Q^{\prime} \subset Q$ we use both $Q \pitchfork_{\tilde{I}} P_{0}$ and $Q^{\prime} \pitchfork_{\tilde{I}} \widetilde{P}_{1}$ to conclude, as in the proof of Proposition 4, that $Q \pitchfork_{\tilde{I}} \widetilde{P}_{1}$.
8.3.3 We now switch back to the setting of Proposition 23.

Let $(P, Q, n) \in \mathcal{R}(\widetilde{I})$ with $P \subset P_{s}$. We say that $P$ is eventually $\widetilde{I}$-separated from $Q^{*}$ if there exists an $\widetilde{I}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $P$ such that $Q^{*}$ and $P_{\alpha}$ are $\widetilde{I}$-separated for every $\alpha$. We say that $P$ is eventually $\widetilde{I}$-transverse to $Q^{*}$ if there exists an $\widetilde{I}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $P$ such that $Q^{*}$ and $P_{\alpha}$ are $\widetilde{I}$-transverse for every $\alpha$. We say that $P$ is eventually $\widetilde{I}$ - $Q^{*}$-critical if it is neither eventually $\widetilde{I}$-separated from $Q^{*}$ nor eventually $\widetilde{I}$-transverse to $Q^{*}$.

Lemma 4. If $P$ is eventually $\widetilde{I}$-transverse to $Q^{*}$ and $2|P|^{1-\eta} \leqslant|\widetilde{I}|$ holds for some $t \in \widetilde{I}$, then $Q^{*} \pitchfork_{\tilde{I}} P$ holds.

Proof. This is an immediate consequence of Proposition 9.
Lemma 5. If $P$ is eventually $\widetilde{I}-Q^{*}$-critical, then $P$ is $\widetilde{I}$-critical.

Proof. If $P$ was $\widetilde{I}$-transverse, there would exist $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{R}(\widetilde{I})$ with $Q^{*} \cap Q_{\alpha} \neq \emptyset$ and $Q_{\alpha} \Pi_{\tilde{I}} P$. But, then, from Proposition 2 if $Q_{\alpha} \supset Q^{*}$ or Proposition 9 if $Q_{\alpha} \subset Q^{*}$, we would deduce that $Q^{*} \pitchfork_{\tilde{I}} P$ also holds.

Lemma 6. If $P$ is eventually $\widetilde{I}-Q^{*}$-critical and $|P|>|\widetilde{I}|^{\beta}$ holds for some $t \in \widetilde{I}$, then some child of $P$ is also eventually $\widetilde{I}-Q^{*}$-critical.

Proof. By Lemma 5 and Corollary 4 in Subsection $6.2, P$ is $\widetilde{I}$-decomposable. If all children of $P$ were eventually $\widetilde{I}$-transverse to $Q^{*}$ (resp. eventually $\widetilde{I}$-separated from $Q^{*}$ ), we would put together the corresponding $\widetilde{I}$-decompositions and obtain that $P$ is eventually $\widetilde{I}$-transverse to $Q^{*}$ (resp. eventually $\widetilde{I}$-separated from $Q^{*}$ ). Therefore, we shall assume that some child $P_{0}$ is eventually $\widetilde{I}$-transverse to $Q^{*}$, and some child $P_{1}$ is eventually $\widetilde{I}$-separated from $Q^{*}$.

By contradiction, we assume that none of the children is eventually $\widetilde{I}-Q^{*}$-critical and we will show that $Q^{*}$ is $\widetilde{I}$-transverse. We construct an $\widetilde{I}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $P_{s}$ such that every $P_{\alpha}$ is either $\widetilde{I}$-separated from $Q^{*}$ or $\widetilde{I}$-transverse to $Q^{*}$ in the following way.

Actually, it is sufficient to have an $\widetilde{I}$-decomposition such that every $P_{\alpha}$ is either eventually $\widetilde{I}$ separated from $Q^{*}$ or eventually $\widetilde{I}$-transverse to $Q^{*}$. Starting from the trivial decomposition of $P_{s}$, we have at step $i$ an $\widetilde{I}$-decomposition $\left(P_{\alpha}^{(i)}, Q_{\alpha}^{(i)}, n_{\alpha}^{(i)}\right)$. As long as there is one $\left(P_{\alpha}^{(i)}, Q_{\alpha}^{(i)}, n_{\alpha}^{(i)}\right)$ with $P \subset P_{\alpha}^{(i)}$, we observe that $P_{\alpha}^{(i)}$ is $\widetilde{I}$-critical and therefore $\widetilde{I}$-decomposable and break it into its children to go to step $i+1$.

After a finite number of steps, each $P_{\alpha}^{(i)}$ is either a child of $P$ or disjoint from $P$. Comparing $P_{\alpha}^{(i)}$ with $P_{0}$ and $P_{1}$, we conclude from Proposition 24 or from the remark before this proposition that $P_{\alpha}^{(i)}$ is either eventually $\widetilde{I}$-transverse to $Q^{*}$ or $\widetilde{I}$-separated from $Q^{*}$. Thus, we have constructed the required $\widetilde{I}$-decomposition and the lemma is proved.

Lemma 7. If $P_{0}, P_{1}$ are eventually $\widetilde{I}-Q^{*}$-critical and disjoint, then we have $\left|P_{0}\right| \leqslant C_{0}|\widetilde{I}|,\left|P_{1}\right| \leqslant$ $C_{0}|\widetilde{I}|$ for all $t \in \widetilde{I}$.

Proof. From Lemma 6, we can find $\left(\widehat{P}_{0}, \widehat{Q}_{0}, \widehat{n}_{0}\right),\left(\widehat{P}_{1}, \widehat{Q}_{1}, \widehat{n}_{1}\right)$ in $\mathcal{R}(\widetilde{I})$ with $\widehat{P}_{0} \subset P_{0}, \widehat{P}_{1} \subset P_{1}$, such that both $\widehat{P}_{0}, \widehat{P}_{1}$ are eventually $\widetilde{I}$ - $Q^{*}$-critical and we have

$$
\begin{equation*}
\left|\widehat{P}_{0}\right| \leqslant|\widetilde{I}|^{\beta}, \quad\left|\widehat{P}_{1}\right| \leqslant|\widetilde{I}|^{\beta} \quad \text { for all } t \in \widetilde{I} . \tag{8.70}
\end{equation*}
$$

If we had, for all $t \in \widetilde{I}$ and $i=0$ or 1 ,

$$
\begin{equation*}
\delta\left(Q^{*}, \widehat{P}_{i}\right) \geqslant 2|\widetilde{I}|, \tag{8.71}
\end{equation*}
$$

then, from (8.62) and (8.70), one would have that $Q^{*} \bar{\Pi}_{\tilde{I}} \widehat{P}_{i}$ holds and $\widehat{P}_{i}$ would not be eventually $\widetilde{I}$ - $Q^{*}$-critical. We have, therefore,

$$
\begin{equation*}
\delta_{L R}\left(Q^{*}, \widehat{P}_{0}\right)<2|\widetilde{I}| \quad \text { for some } t_{0} \in \widetilde{I} \text { and } \tag{8.72}
\end{equation*}
$$

$$
\begin{equation*}
\delta_{L R}\left(Q^{*}, \widehat{P}_{1}\right)<2|\widetilde{I}| \quad \text { for some } t_{1} \in I \tag{8.73}
\end{equation*}
$$

In the same way, we must have

$$
\begin{gather*}
\delta_{L R}\left(Q^{*}, \widehat{P}_{0}\right) \geqslant 0 \quad \text { for some } t_{0}^{\prime} \in \widetilde{I} \text { and also. }  \tag{8.74}\\
\delta_{L R}^{\prime}\left(Q^{*}, \widehat{P}_{1}\right) \geqslant 0 \quad \text { for some } t_{1}^{\prime} \in \widetilde{I} \tag{8.75}
\end{gather*}
$$

Let $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right),\left(P_{0}^{\prime}, Q_{0}^{\prime}, n_{0}^{\prime}\right),\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)$ be the elements of $\mathcal{R}(\widetilde{I})$ such that $Q^{*} \subset Q^{\prime}, \widehat{P}_{0} \subset P_{0}^{\prime}$, $\widehat{P}_{1} \subset P_{1}^{\prime}$, and

$$
\begin{equation*}
\left|Q^{\prime}\right|<|\widetilde{I}|, \quad\left|P_{0}^{\prime}\right|<|\widetilde{I}|, \quad\left|P_{1}^{\prime}\right|<|\widetilde{I}|, \quad \forall t \in \widetilde{I} \tag{8.76}
\end{equation*}
$$

and which are maximal with these properties.
Then, $P_{0}^{\prime}$ satisfies condition ( $* s$ ) of Subsection 7.6: this follows easily from (8.72). Similarly, $P_{1}^{\prime}$ satisfies condition $(* s)$ and $Q^{\prime}$ satisfies condition $(* u)$. We are therefore allowed to apply the estimate of Corollary 11 in Subsection 7.6 to conclude that

$$
\begin{equation*}
\max _{\widetilde{I}} \delta_{L R}\left(Q^{\prime}, P_{i}^{\prime}\right)-\min _{\widetilde{I}} \delta_{L R}\left(Q^{\prime}, P_{i}^{\prime}\right) \leqslant 2|\widetilde{I}| \tag{8.77}
\end{equation*}
$$

for $i=0,1$, and similarly for $\delta, \delta_{L}, \delta_{R}$. We either have (8.63) or (8.64). Assume, for instance, that (8.63) holds. If $P_{1} \subset P_{1}^{\prime}$, we have $\left|P_{1}\right| \leqslant|\widetilde{I}|$ from (8.76). If $P_{1}^{\prime} \subset P_{1}$, we have

$$
\begin{align*}
\max _{\widetilde{I}} \delta_{L}\left(Q^{*}, P_{1}\right) & \leqslant \max _{\widetilde{I}} \delta_{L}\left(Q^{\prime}, P_{1}^{\prime}\right)  \tag{8.78}\\
& \leqslant \min _{\widetilde{I}} \delta_{L}\left(Q^{\prime}, P_{1}^{\prime}\right)+2|\widetilde{I}| \\
& \leqslant \min _{\widetilde{\widetilde{I}}}\left(\delta_{L}\left(Q^{*}, P_{1}\right)+c\left|Q^{\prime}\right|\right)+2|\widetilde{I}| \\
& \leqslant c|\widetilde{I}| .
\end{align*}
$$

Similarly, we have $\left|P_{0}\right| \leqslant|\widetilde{I}|$ from (8.76) if $P_{0} \subset P_{0}^{\prime}$ and otherwise

$$
\begin{align*}
\min _{\widetilde{I}} \delta_{L R}\left(Q^{*}, P_{0}\right) & \geqslant \min _{\widetilde{I}} \delta_{L R}\left(Q^{\prime}, P_{0}^{\prime}\right)-c|\widetilde{I}|  \tag{8.79}\\
& \geqslant \max _{\widetilde{I}} \delta_{L R}\left(Q^{\prime}, P_{0}^{\prime}\right)-(c+2)|\widetilde{I}| \\
& \geqslant-c|\widetilde{I}| .
\end{align*}
$$

Observe that, for all $t \in \widetilde{I}$, the inequalities

$$
\begin{equation*}
c|\widetilde{I}| \geqslant \delta_{L}\left(Q^{*}, P_{1}\right)>\delta_{L R}\left(Q^{*}, P_{0}\right) \geqslant-c|\widetilde{I}| \tag{8.80}
\end{equation*}
$$

are also valid when $P_{1} \subset P_{1}^{\prime}$ or $P_{0} \subset P_{0}^{\prime}$. However, there is a constant $c>0$ such that, for any child $P_{1}^{*}$ of $P_{1}$, any child $P_{0}^{*}$ of $P_{0}$, we have, for all $t \in I$, that

$$
\begin{align*}
\delta_{L}\left(Q^{*}, P_{1}^{*}\right) & \geqslant \delta_{L}\left(Q^{*}, P_{1}\right)+c\left|P_{1}\right|  \tag{8.81}\\
\delta_{L R}\left(Q^{*}, P_{0}^{*}\right) & \leqslant \delta_{L R}\left(Q^{*}, P_{0}\right)-c\left|P_{0}\right| . \tag{8.82}
\end{align*}
$$

As we also must have, for some children $P_{1}^{*}, P_{0}^{*}$, that

$$
\begin{align*}
\delta_{L}\left(Q^{*}, P_{1}^{*}\right) & \leqslant c|\widetilde{I}|  \tag{8.83}\\
\delta_{L R}\left(Q^{*}, P_{0}^{*}\right) & \geqslant-c|\widetilde{I}| \tag{8.84}
\end{align*}
$$

because these children are eventually $\widetilde{I}-Q^{*}$-critical, we obtain the conclusion of the lemma.
8.3.4 Consider the set $\Pi$ of elements $(P, Q, n) \in \mathcal{R}(\widetilde{I})$ which are eventually $\widetilde{I}$ - $Q^{*}$-critical, satisfy

$$
\begin{equation*}
|P| \leqslant|\widetilde{I}|^{1+\tau} \tag{8.85}
\end{equation*}
$$

for all $t \in \widetilde{I}$ and are maximal (in $P$ ) with respect to these two properties.
Lemma 8. We have

$$
\# \Pi \leqslant \frac{1}{C_{1}}|\widetilde{I}|^{-\tau d_{s}^{+}}
$$

where $d_{s}^{+}=d_{s}^{0}+C \eta \tau^{-1}$ is as in the statement of Proposition 24.

Proof. From Lemma 7, there exists a unique element $\left(P_{0}, Q_{0}, n_{0}\right) \in \mathcal{R}(\widetilde{I})$ with the following properties:

- $P \subset P_{0}$ for all $(P, Q, n) \in \Pi$
- $\left|P_{0}\right|>C_{0}|\widetilde{I}|$ for some $t \in \widetilde{I}$
- every child $P_{1}$ of $P_{0}$ which contains a rectangle $P$ with $(P, Q, n) \in \Pi$ satisfies $\left|P_{1}\right| \leqslant C_{o}|\widetilde{I}|$ for all $t \in \widetilde{I}$.

As $P_{0}$ is eventually $\widetilde{I}$ - $Q^{*}$-critical and $\left|P_{0}\right|>C_{0}|\widetilde{I}|$ for some $t \in \widetilde{I}, P_{0}$ must be $\widetilde{I}$-critical. From Proposition 20, the number of children of $P_{0}$ is at most $|\widetilde{I}|^{-c \eta}$.

For every $(P, Q, n) \in \Pi$, either $P$ is a child of $P_{0}$ or the parent $\widetilde{P}$ of $P$ is contained in a child $P_{1}$ of $P_{0}$. Observe that we have by definition of $\Pi$

$$
\begin{equation*}
|\widetilde{P}|>|\widetilde{I}|^{1+\tau} \quad \text { for some } t \in \widetilde{I} \tag{8.86}
\end{equation*}
$$

As $\widetilde{P}$ satisfies condition $(* s)$ of Subsection 7.6 since $P_{0}$ is $\widetilde{I}$-critical, we also have

$$
\begin{equation*}
|\widetilde{P}| \geqslant c^{-1}|\widetilde{I}|^{1+\tau} \quad \text { for all } t \in \widetilde{I} \tag{8.87}
\end{equation*}
$$

As we also have

$$
\begin{equation*}
\left|P_{1}\right| \leqslant C_{0}|\widetilde{I}| \quad \text { for all } t \in \widetilde{I} \tag{8.88}
\end{equation*}
$$

the number of possible $\widetilde{P}$, given $P_{1}$, is from Proposition 22 at most,

$$
\begin{equation*}
\left(C|\widetilde{I}|^{\tau}\right)^{-d_{s}^{*}} \tag{8.89}
\end{equation*}
$$

with $d_{s}^{*}=d_{s}^{0}+\varepsilon_{0}^{\frac{1}{5} d_{s}^{0}}$. The number of children of $\widetilde{P}$ is at most $|\widetilde{I}|^{-c \eta}$. This gives a bound for the cardinality of $\Pi$ by

$$
\begin{equation*}
|\widetilde{I}|^{-c \eta}+|\widetilde{I}|^{-2 c \eta}\left(C|\widetilde{I}|^{\tau}\right)^{-d_{s}^{*}} \tag{8.90}
\end{equation*}
$$

in accordance with the statement of Lemma 8.
8.3.5 Proof of Proposition 23: By Lemma 5 and Corollary 4 in Subsection 6.2, every $(P, Q, n) \in$ $\mathcal{R}(\widetilde{I})$ such that $P$ is eventually $\widetilde{I}$ - $Q^{*}$-critical and $|P| \geqslant|\widetilde{I}|^{\beta}$ (for some $t \in \widetilde{I}$ ) is $\widetilde{I}$-decomposable.

Therefore, there exists an $\widetilde{I}$-decomposition $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ of $P_{s}$ such that every $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ is either eventually $\widetilde{I}$-separated from $Q^{*}$ or eventually $\widetilde{I}$-transverse to $Q^{*}$ or an element of $\Pi$.

Let $I \subset \widetilde{I}$ be a candidate interval of the next level, i.e. $|I|=|\widetilde{I}|^{1+\tau}$, such that $Q^{*}$ is $I$-critical.

We claim that there exists $(P, Q, n) \in \Pi$ such that $P$ is eventually $I$ - $Q^{*}$-critical.
Indeed, every $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ which is eventually $\widetilde{I}$-transverse to $Q^{*}$ (resp. eventually $\widetilde{I}$-separated from $Q^{*}$ ) is a fortiori $I$-transverse to $Q^{*}\left(\right.$ resp. $I$-separated from $\left.Q^{*}\right)$. If every ( $\left.P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \Pi$ was also either eventually $I$-transverse to $Q^{*}$ or eventually $I$-separated from $Q^{*}$, we would obtain a decomposition of $P_{s}$ which expresses that $Q^{*}$ is $I$-transverse.

On the other hand, fix $(P, Q, n) \in \Pi$. We show that there are at most $C_{1}$ candidates $I \subset \widetilde{I}$ such that $(P, Q, n)$ is $I$-critical. Together with Lemma 8, this will imply the statement of Proposition 23.

Choose $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in \mathcal{R}(\widetilde{I})$, with $Q^{*} \subset Q^{\prime},\left|Q^{\prime}\right|<|\widetilde{I}|$ for all $t \in \widetilde{I}$, and maximal with this property as in the proof of Lemma 7. Then, $Q^{\prime}$ satisfies condition $(* u)$ of Subsection 7.6 ; we already know that $P$ satisfies condition $(* s)$. Then, by Corollary 11 of Subsection 7.6, we have

$$
\begin{equation*}
\left|\frac{d}{d t} \delta_{L R}\left(Q^{\prime}, P\right)-1\right| \leqslant C \varepsilon_{0}^{\frac{1}{2}} \tag{8.91}
\end{equation*}
$$

But, if we have for all $t \in I$

$$
\begin{equation*}
\delta_{L R}\left(Q^{\prime}, P\right)<0 \tag{8.92}
\end{equation*}
$$

then $Q^{\prime}$ and $P$ are $I$-separated and a fortiori $Q^{*}$ and $P$ are $I$-separated.

On the other hand, if we have

$$
\begin{equation*}
\delta\left(Q^{\prime}, P\right)>3|I| \quad \text { for all } t \in I \tag{8.93}
\end{equation*}
$$

it is easy to conclude from (8.62) and Proposition 9 , that $P$ is eventually $I$-transverse to $Q^{*}$. But, by (5.11)-(5.14), we have

$$
\begin{equation*}
\delta_{L R}\left(Q^{\prime}, P\right)<\delta\left(Q^{\prime}, P\right)+c|I| \tag{8.94}
\end{equation*}
$$

and our claim then follows from (8.91). Proposition 23 is proved.

## 9 Strong Regularity and Parameter Selection

### 9.1 Partitions of the Critical Locus

As it was mentioned in Subsection 4.5, regularity is a rather qualitative property which is not appropriate for the quantitative estimates needed for parameter selection. Therefore, we will introduce later in this section, a stronger quantitative property, that we call strong regularity.

This property is built up from several bounds: the number of bicritical elements of a given size (including, of course, that there are no "fat" bicritical elements), and also sizes of the "critical locus" (see below), which should be of approximate dimension $d_{s}^{0}+d_{u}^{0}-1$. These last bounds are more elementary and will be taken care of first.

In all of Section 9, we fix a parameter interval $\widetilde{I}$ which is strongly regular, i.e. satisfies all bounds that will be stated shortly. We will prove at some point that the starting interval satisfies this strong regularity. As mentioned before, and as will be proved in Subsection 9.6, strong regularity implies $\beta$-regularity for some $\beta>1$. We denote by $I$ any parameter interval contained in $\widetilde{I}$ of the next level, i.e. $|I|=|\widetilde{I}|^{1+\tau}$; such a candidate interval is $\bar{\beta}$-regular (Corollary 9 in Subsection 6.7), with $\bar{\beta}=\beta(1+\tau)^{-1}$. The aim of this section is to estimate how many of the candidates $I$ fail to be strongly regular.

We denote by $\mathcal{C}_{+}(\widetilde{I})$ the set of $(P, Q, n)$ in $\mathcal{R}(\widetilde{I})$ such that $P$ is $\widetilde{I}$-critical, $|P| \leqslant|\widetilde{I}|^{1+\tau}$ for all $t \in \widetilde{I}$, and $P$ is maximal with this property: the parent $\widetilde{P}$ of $P$ satisfies $|\widetilde{P}|>|\widetilde{I}|^{1+\tau}$ for some $t \in \widetilde{I}$.

Obviously, if $(P, Q, n),\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ are distinct elements in $\mathcal{C}_{+}(\widetilde{I}), P$ and $P^{\prime}$ are disjoint. Moreover, if $(\widehat{P}, \widehat{Q}, \widehat{n})$ belongs to $\mathcal{R}(\widetilde{I}), \widehat{P}$ is $\widetilde{I}$-critical, and $|\widehat{P}| \leqslant|\widetilde{I}|^{1+\tau}$ for all $t \in \widetilde{I}$, there is a unique $(P, Q, n) \in \mathcal{C}_{+}(\widetilde{I})$ such that $\widehat{P} \subset P$.

Exchanging $P$ 's and $Q$ 's, we define $\mathcal{C}_{-}(\widetilde{I})$ in a similar way. The sets $\mathcal{C}_{+}(\widetilde{I}), \mathcal{C}_{-}(\widetilde{I})$ correspond to the $\widetilde{I}$-critical locus at the $|\widetilde{I}|^{1+\tau}$ scale. We will also need to consider this at the $|\widetilde{I}|$ scale, as follows. We define $\widehat{\mathcal{C}}_{+}(\widetilde{I})$ to be the set of $(P, Q, n) \in \mathcal{R}(\widetilde{I})$ such that $P$ contains some $P^{\prime}$ with $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in \mathcal{C}_{+}(\widetilde{I})$, $|P| \leqslant|\widetilde{I}|$ for all $t \in \widetilde{I}$, and $P$ is maximal with this property. We define similarly $\widehat{\mathcal{C}}_{-}(\widetilde{I})$.

We will need in the sequel to consider $\widehat{I}$-criticality (for some parameter interval $\widehat{I} \supset \widetilde{I}$ ) for rectangles in $\mathcal{R}(\widetilde{I})$ but not in $\mathcal{R}(\widehat{I})$. The following definition will be useful.

Definition. Let $I_{\alpha} \supset I$ be parameter intervals, and let $(P, Q, n) \in \mathcal{R}(I)$. We say that $P$ is thin $I_{\alpha}$-critical if there exists $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{R}\left(I_{\alpha}\right)$ with $P \subset P_{\alpha}, P_{\alpha}$ is $I_{\alpha}$-critical and

$$
\left|P_{\alpha}\right|^{1-\eta} \leqslant 2\left|I_{\alpha}\right|
$$

for some $t \in I_{\alpha}$.

The main justification for this definition is the following lemma, which is an immediate consequence of the structure theorem of Subsection 6.7 and Proposition 11 in Subsection 6.6.

Lemma 9. Let I be a parameter interval with parent $\widetilde{I}$, and let $(P, Q, n)$ be an element of $\mathcal{R}(I)$ which is not the restriction of an element of $\mathcal{R}(\widetilde{I})$. Let $\left(P_{i}, Q_{i}, n_{i}\right)$, for $0 \leqslant i \leqslant k$, be the elements of $\mathcal{R}(\widetilde{I})$ given by the structure theorem of Subsection 6.7. Then $P_{i}$ is thin $\widetilde{I}$-critical for $0<i \leqslant k$, and $Q_{i}$ is thin $\widetilde{I}$-critical for $0 \leqslant i<k$.

### 9.2 Size of the Critical Locus

We will state several inequalities related to the size of the sets $\mathcal{C}_{+}(I), \mathcal{C}_{-}(I), \widehat{\mathcal{C}}_{+}(I), \widehat{\mathcal{C}}_{-}(I)$. All these inequalities are part of the definition of strong regularity: they have to be satisfied by a strongly regular parameter interval. We will then see that if these inequalities are satisfied by the parent interval $\widetilde{I}$, most of the candidates $I \subset \widetilde{I}$ also satisfy these inequalities.

Recall that $d_{s}^{0}, d_{u}^{0}$ are the transverse Hausdorff dimensions of $W^{s}(K), W^{u}(K)$ at $t=0$. In Proposition 23 of Subsection 8.3, we introduced $d_{s}^{+}=d_{s}^{0}+C \eta \tau^{-1}$. Let also $d_{u}^{+}=d_{u}^{0}+C \eta \tau^{-1}$.

In Proposition 22 of Subsection 8.2, we have used $d_{s}^{*}:=d_{s}^{0}+\varepsilon_{0}^{\frac{1}{5}} d_{s}^{0}<d_{s}^{+}$. Define similarly

$$
d_{u}^{*}=d_{u}^{0}+\varepsilon_{0}^{\frac{1}{5}} d_{u}^{0} .
$$

We will control the cardinalities of $\mathcal{C}_{+}(I), \mathcal{C}_{-}(I), \widehat{\mathcal{C}}_{+}(I), \widehat{\mathcal{C}}_{-}(I)$, through:

$$
(\mathrm{SR} 1)_{u}
$$

$$
\begin{align*}
& \# \mathcal{C}_{+}(I) \leqslant C\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{s}^{+}-d_{u}^{+}-\tau} \varepsilon_{0}^{-\tau d_{s}^{0}},  \tag{SR1}\\
& \# \widehat{\mathcal{C}}_{+}(I) \leqslant C\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{s}^{+}-d_{u}^{+}-\tau},  \tag{SR1}\\
& \# \mathcal{C}_{-}(I) \leqslant C\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{s}^{+}-d_{u}^{+}-\tau} \varepsilon_{0}^{-\tau d_{u}^{0}}, \\
& \# \widehat{\mathcal{C}}_{-}(I) \leqslant C\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{s}^{+}-d_{u}^{+}-\tau} . \tag{SR1}
\end{align*}
$$

We will also control a weighted cardinality through

$$
\begin{equation*}
\sum_{\mathcal{C}_{+}(I)}|Q|^{d_{u}^{*}} \leqslant C\left|Q_{s}\right|^{d_{u}^{*}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{u}^{+}-\tau}, \tag{SR2}
\end{equation*}
$$

$(\mathrm{SR} 2)_{\widehat{s}}$
$(\mathrm{SR} 2)_{u}$
$(\mathrm{SR})_{\widehat{u}}$

$$
\begin{aligned}
& \sum_{\hat{\mathcal{C}}_{+}(I)}|Q|^{d_{u}^{*}} \leqslant C\left|Q_{s}\right|^{d_{u}^{*}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{u}^{+}-\tau}, \\
& \sum_{\mathcal{C}_{-}(I)}|P|^{d_{s}^{*}} \leqslant C\left|P_{u}\right|^{d_{s}^{*}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{s}^{+}-\tau}, \\
& \sum_{\hat{\mathcal{C}}_{-}(I)}|P|^{d_{s}^{*}} \leqslant C\left|P_{u}\right|^{d_{s}^{*}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{1-d_{s}^{+}-\tau} .
\end{aligned}
$$

The heuristics behind the second set of inequalities is the following: in the mean, one expects that elements of $\mathcal{R}(I)$ more or less satisfy

$$
\begin{equation*}
|P|^{d_{s}^{0}} \sim|Q|^{d_{u}^{0}} \tag{9.1}
\end{equation*}
$$

and, for $(P, Q, n) \in \widehat{\mathcal{C}}_{+}(I)$ (for instance), one should have

$$
\begin{equation*}
|P| \sim|I| \tag{9.2}
\end{equation*}
$$

which explains the relation between $(\mathrm{SR} 1)_{\widehat{s}}$ and $(\mathrm{SR} 2)_{\widehat{s}}$.

Let us check that the starting interval $I_{0}$ satisfies these eight inequalities.

Then $\mathcal{C}_{+}\left(I_{0}\right)$ (resp. $\left.\mathcal{C}_{-}\left(I_{0}\right)\right)$ consists of the elements $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$ with $P \cap P_{s} \neq \emptyset$ (resp. $Q \cap Q_{u} \neq$ $\emptyset),|P| \leqslant \varepsilon_{0}^{1+\tau}$ for all $t \in I_{0}$ (resp. $|Q| \leqslant \varepsilon_{0}^{1+\tau}$ for all $t \in I_{0}$ ) and maximal with this property. Then, $(\mathrm{SR} 1)_{s}$ follows from (8.38) and (SR1) $)_{u}$ is similar. Writing $(P, Q, n)=\left(P_{s}, Q_{s}, n_{s}\right) *\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$, the inequality $(\mathrm{SR} 2)_{s}$ becomes

$$
\begin{equation*}
\sum_{\mathcal{C}_{+}(I)}\left|Q^{\prime}\right|^{d_{u}^{0}+C \varepsilon_{0}} \leqslant C \tag{9.3}
\end{equation*}
$$

which is a standard property of uniformly hyperbolic horseshoes. The proof of (SR2) ${ }_{u}$ is similar. The case of $\widehat{\mathcal{C}}_{+}\left(I_{0}\right), \widehat{\mathcal{C}}_{-}\left(I_{0}\right)$ is even simpler. For the induction step, we have:

Proposition 25. If the parent interval $\widetilde{I}$ is $\beta$-regular and satisfies one of the eight inequalities (SR) above, then all candidates $I \subset \widetilde{I}$ satisfy the same inequality except perhaps for a proportion not larger than $C|\widetilde{I}|^{\tau^{2}}$.

Notation. Let $(P, Q, n) \in \mathcal{C}_{+}(\widetilde{I})$. We denote by $C r(P)$ the set of candidates $I \subset \widetilde{I}$ such that $P$ contains a thin $I$-critical rectangle.

Lemma 10. For any $(P, Q, n) \in \mathcal{C}_{+}(\widetilde{I})$, we have

$$
\# C r(P) \leqslant C|\widetilde{I}|^{-\tau d_{u}^{+}} .
$$

Proof. Let $\left(P^{*}, Q^{*}, n^{*}\right) \in \mathcal{R}(\widetilde{I})$ be an element such that $P^{*} \subset P$ and

$$
\begin{equation*}
\left|P^{*}\right| \leqslant \frac{1}{2}|\widetilde{I}|^{(1+\tau)(1-\eta)^{-1}} \tag{9.4}
\end{equation*}
$$

for all $t \in \widetilde{I}$. By Proposition 23 in Subsection 8.3 , there are at most $|\widetilde{I}|^{-\tau d_{u}^{+}}$candidates $I \subset \widetilde{I}$ such that $P^{*}$ is $I$-critical. Let $I \in C r(P)$. By definition, there exists $\left(P_{0}, Q_{0}, n_{0}\right) \in \mathcal{R}(I)$ with $P_{0} \subset P$, $P_{0} I$-critical and

$$
\begin{equation*}
\left|P_{0}\right|^{1-\eta} \leqslant 2|I| \text { for some } t_{0} \in I \tag{9.5}
\end{equation*}
$$

But if $P_{0}$ is $I$-critical and (9.5) holds, there must exist $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in \mathcal{R}(I)$ and $t_{1} \in I$ such that

$$
\begin{equation*}
0 \leqslant \delta_{L R}\left(Q^{\prime}, P_{0}\right) \leqslant 2|I| \tag{9.6}
\end{equation*}
$$

As $P_{0}, P^{*}$ are contained in $P$ and $|P| \leqslant|I|$ for all $t \in \widetilde{I}$, we also have

$$
\begin{equation*}
\left|\delta_{L R}\left(Q^{\prime}, P^{*}\right)\right| \leqslant C|I| . \tag{9.7}
\end{equation*}
$$

Proceeding as in the Proof of Lemma 7, Subsection 8.3, we deduce from Corollary 11 in Subsection 7.6, that there exists $I^{\prime} \subset \widetilde{I}$ at distance $c|I|$ of $I$ such that $P^{*}$ is $I^{\prime}$-critical (except perhaps when $I$ is very close to the boundary of $\widetilde{I})$. This proves the lemma.

Proof of Proposition 25. We will deal with $(\mathrm{SR} 1)_{s},(\mathrm{SR} 1)_{\widehat{s}},(\mathrm{SR} 2)_{s},(\mathrm{SR} 2)_{\widehat{s}}$, the other four being symmetric.

For each $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I})$, we consider the elements $\left(P_{\alpha, i}, Q_{\alpha, i}, n_{\alpha, i}\right) \in \mathcal{R}(\widetilde{I})$ which satisfy $P_{\alpha, i} \subset P_{\alpha},\left|P_{\alpha, i}\right| \leqslant|\widetilde{I}|^{(1+\tau)^{2}}$ for all $t \in \widetilde{I}$, and which are maximal with respect to this property.

This gives an $\widetilde{I}$-decomposition of $P_{\alpha}$ : indeed, it is easy to see that for each $\left(P_{\alpha}^{\prime}, Q_{\alpha}^{\prime}, n_{\alpha}^{\prime}\right) \in \mathcal{R}(\widetilde{I})$ with $P_{\alpha}^{\prime} \subset P_{\alpha}$ and $\left|P_{\alpha}^{\prime}\right|>|\widetilde{I}|^{(1+\tau)^{2}}$ for some $t \in \widetilde{I}, Q_{\alpha}^{\prime}$ is $\widetilde{I}$-transverse and therefore $P_{\alpha}^{\prime}$ is $\widetilde{I}$-decomposable.

Using Propositions 20 and 22, we argue as in the proof of Lemma 8, Subsection 8.3, to see that for each $P_{\alpha}$, the number of $P_{\alpha, i}$ is not larger than

$$
\begin{equation*}
|\widetilde{I}|^{-c \eta}|\widetilde{I}|^{-\tau(1+\tau) d_{s}^{*}} \tag{9.8}
\end{equation*}
$$

with $d_{s}^{*}=d_{s}^{0}+\varepsilon_{0}^{\frac{1}{5} d_{s}^{0}}$.

Obviously, if $(P, Q, n) \in \mathcal{C}_{+}(I)$, there exists $\alpha, i$ such that $P \supset P_{\alpha, i}$ and $I$ must belong to $C r\left(P_{\alpha}\right)$. We, therefore, have

$$
\begin{align*}
\sum_{I \subset \widetilde{I}} \# \mathcal{C}_{+}(I) & \leqslant C|\widetilde{I}|^{-c \eta-\tau(1+\tau) d_{s}^{*}} \sum_{\mathcal{C}_{+}(\widetilde{I})} \# C r\left(P_{\alpha}\right)  \tag{9.9}\\
& \leqslant \# \mathcal{C}_{+}(\widetilde{I})|\widetilde{I}|^{-c \eta-\tau(1+\tau) d_{s}^{*}} \max \# C r\left(P_{\alpha}\right)
\end{align*}
$$

We have

$$
\begin{equation*}
d_{s}^{+} \geqslant d_{s}^{*}(1+\tau)+C \eta \tau^{-1} \tag{9.10}
\end{equation*}
$$

and therefore, using also Lemma 10:

$$
\begin{equation*}
\sum_{I \subset \widetilde{I}} \# \mathcal{C}_{+}(I) \leqslant C \# \mathcal{C}_{+}(\widetilde{I})|\widetilde{I}|^{-\tau\left(d_{s}^{+}+d_{u}^{+}\right)} \tag{9.11}
\end{equation*}
$$

The induction step for $(\mathrm{SR} 1)_{s}$ follows immediately. In the same way as in (9.8), we obtain from Propositions 20 and 22 that

$$
\begin{equation*}
\# \mathcal{C}_{+}(\widetilde{I}) \leqslant|\widetilde{I}|^{-C \eta}|\widetilde{I}|^{-\tau d_{s}^{*}} \# \widehat{\mathcal{C}}_{+}(\widetilde{I}) \tag{9.12}
\end{equation*}
$$

On the other hand, for every $(P, Q, n) \in \widehat{\mathcal{C}}_{+}(I)$, there exists a $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I})$ with $P_{\alpha} \subset P$ and $I \in C r\left(P_{\alpha}\right)$. Therefore, we have

$$
\begin{equation*}
\sum_{I \subset \widetilde{I}} \# \widehat{\mathcal{C}}_{+}(I) \leqslant \sum_{\mathcal{C}_{+}(\widetilde{I})} \# C r\left(P_{\alpha}\right) \tag{9.13}
\end{equation*}
$$

Putting together $(9.12),(9.13)$ and Lemma 10, we get

$$
\begin{equation*}
\sum \# \widehat{\mathcal{C}}_{+}(I) \leqslant C|\widetilde{I}|^{-C \eta-\tau d_{s}^{*}-\tau d_{u}^{+}} \# \widehat{\mathcal{C}}_{+}(\widetilde{I}) \tag{9.14}
\end{equation*}
$$

and we conclude as above.

Let us now consider $(\mathrm{SR} 2)_{s}$. We must sum $|Q|^{d_{u}^{*}}$ over elements $(P, Q, n)$ of $\mathcal{C}_{+}(I)$. Fix $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ in $\mathcal{C}_{+}(\widetilde{I})$ such that $I \in C r\left(P_{\alpha}\right)$. When we consider the partial sum over those $Q$ contained in $Q_{\alpha}$, we get, from Proposition 22 in Subsection 8.2.

$$
\begin{equation*}
\sum_{Q \subset Q_{\alpha}}|Q|^{d_{u}^{*}} \leqslant C\left|Q_{\alpha}\right|^{d_{u}^{*}} \tag{9.15}
\end{equation*}
$$

As every $Q$ is contained in such a $Q_{\alpha}$, we will have:

$$
\begin{equation*}
\sum_{\mathcal{C}_{+}(I)}|Q|^{d_{u}^{*}} \leqslant C \sum_{I \in \operatorname{Cr}\left(P_{\alpha}\right)}\left|Q_{\alpha}\right|^{d_{u}^{*}} . \tag{9.16}
\end{equation*}
$$

Summing then over candidates $I$ and using Lemma 10, we get

$$
\begin{align*}
\sum_{I} \sum_{\mathcal{C}_{+}(I)}|Q|^{d_{u}^{*}} & \leqslant C \sum_{\mathcal{C}_{+}(\widetilde{I})}\left|Q_{\alpha}\right|^{d_{u}^{*}} \# C r\left(P_{\alpha}\right)  \tag{9.17}\\
& \leqslant C|\widetilde{I}|^{-\tau d_{u}^{+}} \sum_{\mathcal{C}_{+}(\widetilde{I})}\left|Q_{\alpha}\right|^{d_{u}^{*}}
\end{align*}
$$

which allow us to conclude as above the induction step for $(\mathrm{SR} 2)_{s}$. The argument for $(\mathrm{SR} 2)_{\widehat{s}}$ is similar and left to the reader.

### 9.3 Classes of Bicritical Rectangles

Now that the size of the critical locus is under control, we must pay attention to the number of bicritical rectangles, which represent the returns of the critical locus to itself under the dynamics.

In order to have an appropriate induction scheme, we need to bound the number of bicritical rectangles according to all width scales and also according to the level of criticality (i.e., the distance to critical locus) of both $P$ and $Q$. As we will see in the next subsection, the number of bicritical elements experiments a "phase transition" which is crucial for our argument but brings a lot of complications.

Let $I$ be a candidate interval as above, and let $I_{\alpha}, I_{\omega}$ be parameter intervals such that $I \subset I_{\alpha} \cap I_{\omega}$. Let also $x$ be a positive number.

Definition. We denote by $B i_{+}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ the set of elements $(P, Q, n) \in \mathcal{R}(I)$ such that $P$ is thin $I_{\alpha}$-critical, $Q$ is thin $I_{\omega}$-critical and $|P| \geqslant x$ for some $t \in I$.

Similarly, $B i_{-}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ is the set of elements $(P, Q, n) \in \mathcal{R}(I)$ such that $P$ is thin $I_{\alpha}$-critical, $Q$ is thin $I_{\omega}$-critical and $|Q| \geqslant x$ for some $t \in I$.

We denote by $B i_{ \pm}^{n e w}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ the set of elements $(P, Q, n) \in B i_{ \pm}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ that do not belong to $\mathcal{R}(\widetilde{I}), \widetilde{I}$ being as above the parent interval of $I$.

We will estimate the cardinality of all sets $B i_{ \pm}\left(I, I_{\alpha}, I_{\omega} ; x\right)$, by induction on the level of the parameter interval $I$. The easy case is when $I$ is strictly smaller than $I_{\alpha}$ and $I_{\omega}$. In this case, no parameter selection is needed. An element of $B i_{+}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ is either in $B i_{+}^{n e w}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ or in $B i_{+}\left(\widetilde{I}, I_{\alpha}, I_{\omega} ; x\right)$, where $\widetilde{I}$ is the parent interval of $I$. We will see that the cardinality of $B i_{+}^{n e w}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ can be estimated from the induction hypothesis and the structure theorem of Subsection 6.7 , and this cardinality is much smaller than the cardinality of $B i_{+}\left(\widetilde{I}, I_{\alpha}, I_{\omega} ; x\right)$ which is controlled by the induction hypothesis.

When $I$ is equal to $I_{\alpha}$, but strictly smaller than $I_{\omega}$ (or in the symmetric case $I=I_{\omega} \varsubsetneqq I_{\alpha}$ ), a parameter selection is needed in order to obtain satisfactory estimates. An element of $B i_{+}\left(I, I, I_{\omega} ; x\right)$ is either in $B i_{+}^{\text {new }}\left(I, I, I_{\omega} ; x\right)$, whose cardinality can be again estimated from the structure theorem and the induction hypothesis, or it belongs to $B i_{+}\left(\widetilde{I}, \widetilde{I}, I_{\omega} ; x\right)$. But in this last case, the vertical rectangle $P$, which is known to be $\widetilde{I}$-critical, is only $I$-critical for a small fraction of candidates $I$ which is controlled by Proposition 23 in Subsection 8.3. Averaging, like in Subsection 9.2, allows us to get the required estimate.

By far the most difficult case occurs when $I=I_{\alpha}=I_{\omega}$. When $x$ is large, we have a set of estimates, which is taken care of in the same way as for $I=I_{\alpha} \neq I_{\omega}$. But when $x$ is small, the parameter selection process is much more subtle and will be explained in Subsection 9.8.

### 9.4 Number of Bicritical Rectangles

In this subsection, we will state, and comment, the estimates for the cardinalities of the sets $B i_{ \pm}$ introduced above. The rest of Section 9 will then be devoted to the proof of these estimates under appropriate parameter selection.

At this point, we have to break the symmetry between past and future, $P$ 's and $Q$ 's, stable and unstable direction: the estimates are indeed not symmetric, except when $d_{s}^{0}=d_{u}^{0}$, i.e., in the conservative case of area-preserving diffeomorphisms.

We will assume that $d_{s}^{0} \geqslant d_{u}^{0}$ (and $d_{s}^{0}+d_{u}^{0}>1$ ). The case $d_{u}^{0} \geqslant d_{s}^{0}$ is obviously symmetric.
For $I, I_{\alpha}, I_{\omega}, x$ as above we want to have

$$
\begin{equation*}
\# B i_{+}\left(I, I_{\alpha}, I_{\omega} ; x\right) \leqslant C B, \tag{SR3}
\end{equation*}
$$

with

$$
\begin{align*}
B & =\max \left(B_{0}, B_{1}\right),  \tag{9.18}\\
B_{0} & =\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}\left(\frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}},  \tag{9.19}\\
B_{1} & =\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}\left(\min \left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}, \frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)\right)^{\sigma_{0}} . \tag{9.20}
\end{align*}
$$

Here $\left|P_{u}\right|$ denotes the supremum over $I_{0}$ of the width of $P_{u}$ : the exponents $\rho_{0}, \rho_{1}, \sigma_{0}, \sigma_{1}$ will be specified more precisely later, but anyway they satisfy

$$
\begin{align*}
\rho_{0} & =d_{s}^{0}+o(1),  \tag{9.21}\\
\rho_{1} & =\frac{d_{s}^{0}}{d_{s}^{0}+d_{u}^{0}}\left(2 d_{s}^{0}+d_{u}^{0}-1\right)+o(1),  \tag{9.22}\\
\sigma_{0} & =1-d_{s}^{0}+o(1),  \tag{9.23}\\
\sigma_{1} & =d_{s}^{0}-d_{u}^{0}+o(1) . \tag{9.24}
\end{align*}
$$

The meaning of the $o(1)$ terms in these formulas is that they become arbitrarily small when $\tau \gg$ $\eta \gg \varepsilon_{0}$ are small enough.

For the $B i_{-}$sets, we should have:
$(\mathrm{SR} 3)_{u}$

$$
\# B i_{-}\left(I, I_{\alpha}, I_{\omega} ; x\right) \leqslant C B^{\prime},
$$

with

$$
\begin{align*}
B^{\prime} & =\max \left(B_{0}^{\prime}, B_{1}^{\prime}\right)  \tag{9.25}\\
B_{0}^{\prime} & =\left(\frac{x}{\varepsilon_{0}\left|Q_{s}\right|}\right)^{-\rho_{0}^{\prime}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}\left(\frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}},  \tag{9.26}\\
B_{1}^{\prime} & =\left(\frac{x}{\varepsilon_{0}\left|Q_{s}\right|}\right)^{-\rho_{1}^{\prime}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}\left(\min \left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}, \frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)\right)^{\sigma_{0}},  \tag{9.27}\\
\rho_{0}^{\prime} & =\frac{d_{u}^{0}}{d_{s}^{0}} \rho_{0}=d_{u}^{0}+o(1), \\
\rho_{1}^{\prime} & =\frac{d_{u}^{0}}{d_{s}^{0}} \rho_{1}=\frac{d_{u}^{0}}{d_{s}^{0}+d_{u}^{0}}\left(2 d_{s}^{0}+d_{u}^{0}-1\right)+o(1)
\end{align*}
$$

Observe that the formulas (9.22), (9.29) for $\rho_{1}, \rho_{1}^{\prime}$ are not symmetric.

Definition. A parameter interval is strongly regular if it satisfies (SR3) ${ }_{s}$, (SR3) ${ }_{u}$ and the eight conditions (SR1), (SR2) of Subsection 8.2.

Remark. The definition will only be complete when we specify precisely the exponents $\rho_{0}, \rho_{1}, \cdots$.

We now comment on the inequalities above. First, observe that $B$ does not depend on $I$ : this reflects the fact mentioned above that $B i_{+}^{\text {new }}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ is small compared with $B i_{+}\left(\widetilde{I}, I_{\alpha}, I_{\omega} ; x\right)$.

From the formulas (9.21), (9.22), we have

$$
\begin{equation*}
\rho_{1}<\rho_{0} . \tag{9.30}
\end{equation*}
$$

Set

$$
\begin{equation*}
x_{c r}:=\varepsilon_{0}\left|P_{u}\right|\left(\max \left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}, \frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)\right)^{\frac{\sigma_{0}}{\rho_{0}-\rho_{1}}} . \tag{9.31}
\end{equation*}
$$

Then, we have $B=B_{0}$ for $x \leqslant x_{c r}$ and $B=B_{1}$ for $x \geqslant x_{c r}$ : this is the "phase transition" mentioned earlier. We have

$$
\begin{align*}
\rho_{0}-\rho_{1} & =\frac{d_{s}^{0}\left(1-d_{s}^{0}\right)}{d_{s}^{0}+d_{u}^{0}}+o(1),  \tag{9.32}\\
\frac{\sigma_{0}}{\rho_{0}-\rho_{1}} & =\frac{d_{s}^{0}+d_{u}^{0}}{d_{s}^{0}}+o(1)>1 . \tag{9.33}
\end{align*}
$$

For $x=x_{c r}$, we have

$$
\begin{equation*}
B=B_{c r}:=\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}\left(\frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}}\left(\max \left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}, \frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)\right)^{-\frac{\rho_{0} \sigma_{0}}{\rho_{0}-\rho_{1}}} . \tag{9.34}
\end{equation*}
$$

Assume $I_{\alpha}=I_{\omega}$; we then have

$$
\begin{equation*}
B_{c r}=\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}+\sigma_{0} \frac{\rho_{0}-2 \rho_{1}}{\rho_{0}-\rho_{1}}} \tag{9.35}
\end{equation*}
$$

Here, the exponent satisfies

$$
\begin{equation*}
\sigma_{1}+\sigma_{0} \frac{\rho_{0}-2 \rho_{1}}{\rho_{0}-\rho_{1}}=2-2 d_{s}^{0}-2 d_{u}^{0}+o(1)<0 . \tag{9.36}
\end{equation*}
$$

As $\left|I_{\alpha}\right| \leqslant \varepsilon_{0}$, we have $B_{c r} \geqslant 1$. As $B$ is a decreasing function of $x$, we have $B<1$ (in which case (SR3) $)_{s}$ means that the $B i_{+}$set is empty!) iff $B_{1}<1$ which corresponds to

$$
\begin{equation*}
x>\bar{x}:=\varepsilon_{0}\left|P_{u}\right|\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\frac{\sigma_{0}+\sigma_{1}}{\rho_{1}}} . \tag{9.37}
\end{equation*}
$$

The exponent here satisfies

$$
\begin{equation*}
\frac{\sigma_{0}+\sigma_{1}}{\rho_{1}}=\frac{1-d_{u}^{0}}{d_{s}^{0}} \frac{d_{s}^{0}+d_{u}^{0}}{2 d_{s}^{0}+d_{u}^{0}-1}+o(1) \tag{9.38}
\end{equation*}
$$

We are finally able to justify the assumption (H4) of our Main Theorem stated in Subsection 1.2! Indeed, with $d_{s}^{0} \geqslant d_{u}^{0}$, it means that

$$
\begin{equation*}
2\left(d_{s}^{0}\right)^{2}+\left(d_{u}^{0}\right)^{2}+2 d_{s}^{0} d_{u}^{0}<2 d_{s}^{0}+d_{u}^{0} \tag{H4}
\end{equation*}
$$

and this is exactly what is needed to guarantee that

$$
\begin{equation*}
\frac{\sigma_{0}+\sigma_{1}}{\rho_{1}}>1 \tag{9.39}
\end{equation*}
$$

We will choose the constant $\beta$ (related to the regularity property) in order to have

$$
\begin{equation*}
1<\beta<\frac{\sigma_{0}+\sigma_{1}}{\rho_{1}} \tag{9.40}
\end{equation*}
$$

and also

$$
\begin{equation*}
\left|P_{u}\right|<\varepsilon_{0}^{\beta-1} \tag{9.41}
\end{equation*}
$$

Then, from (9.37), we will have that

$$
\begin{equation*}
\bar{x}<\left|I_{\alpha}\right|^{\beta} \tag{9.42}
\end{equation*}
$$

Summarizing: if $(\mathrm{SR} 3)_{s}$ holds, and if $(P, Q, n) \in \mathcal{R}(I)$ is such that $P$ and $Q$ are thin $I_{\alpha}$-critical, we must have

$$
\begin{equation*}
|P|<\bar{x}<\left|I_{\alpha}\right|^{\beta} \tag{9.43}
\end{equation*}
$$

for all $t \in I$. This is almost what we need for $\beta$-regularity. The full proof is given below.

The discussion for $(\mathrm{SR} 3)_{u}$ is similar; the critical threshold is

$$
\begin{equation*}
x_{c r}^{\prime}:=\varepsilon_{0}\left|Q_{s}\right|\left(\max \left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}, \frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)\right)^{\frac{\sigma_{0}}{\rho_{0}^{\prime}-\rho_{1}^{\prime}}} \tag{9.44}
\end{equation*}
$$

with

$$
\begin{align*}
\rho_{0}^{\prime}-\rho_{1}^{\prime} & =\frac{d_{u}^{0}\left(1-d_{s}^{0}\right)}{d_{s}^{0}+d_{u}^{0}}+o(1)=\frac{d_{u}^{0}}{d_{s}^{0}}\left(\rho_{0}-\rho_{1}\right)  \tag{9.45}\\
\frac{\sigma_{0}}{\rho_{0}^{\prime}-\rho_{1}^{\prime}} & =\frac{\sigma_{0}}{\rho_{0}-\rho_{1}} \frac{d_{s}^{0}}{d_{u}^{0}}=\frac{d_{s}^{0}+d_{u}^{0}}{d_{u}^{0}}+o(1)>1 \tag{9.46}
\end{align*}
$$

When $I_{\alpha}=I_{\omega}$, we have

$$
\begin{equation*}
B_{c r}^{\prime}:=\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}+\sigma_{0} \frac{\rho_{0}^{\prime}-2 \rho_{1}^{\prime}}{\rho_{0}^{\prime}-\rho_{1}^{\prime}}}=B_{c r} \geqslant 1 \tag{9.47}
\end{equation*}
$$

Thus, we have $B^{\prime}<1$ if

$$
\begin{equation*}
x>\bar{x}^{\prime}:=\varepsilon_{0}\left|Q_{s}\right|\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\frac{\sigma_{0}+\sigma_{1}}{\rho_{1}^{\prime}}} \tag{9.48}
\end{equation*}
$$

We have here

$$
\begin{equation*}
\frac{\sigma_{0}+\sigma_{1}}{\rho_{1}^{\prime}}=\frac{\sigma_{0}+\sigma_{1}}{\rho_{1}} \frac{d_{s}^{0}}{d_{u}^{0}}>\beta \tag{9.49}
\end{equation*}
$$

and we choose $\beta$ close enough to one to have

$$
\begin{equation*}
\left|Q_{s}\right|<\varepsilon_{0}^{\beta-1} . \tag{9.50}
\end{equation*}
$$

Then, we are again able to conclude that, if $(\mathrm{SR} 3)_{u}$ holds, any $(P, Q, n) \in \mathcal{R}(I)$ such that both $P$ and $Q$ are thin $I_{\alpha}$-critical must satisfy, for all $t \in I$

$$
\begin{equation*}
|Q|<\bar{x}^{\prime}<\left|I_{\alpha}\right|^{\beta} . \tag{9.51}
\end{equation*}
$$

Proposition 26. If a candidate interval satisfies (SR3)s and (SR3) ${ }_{u}$, then it is $\beta$-regular. In particular, strong regularity implies regularity.

Proof. We argue by induction on the level of the parameter interval. For the starting interval $I_{0}$, we use (for the first time!) the assumption (H1) of Subsection 1.2 that the periodic points $p_{s}, p_{u}$ do not belong to the same periodic orbit. Then, if $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$ is such that $P \subset P_{s}, Q \subset Q_{u}$, we must have, for all $t \in I_{0}$

$$
\begin{equation*}
|P|<\varepsilon_{0}^{\beta}, \quad|Q|<\varepsilon_{0}^{\beta} \tag{9.52}
\end{equation*}
$$

if we take $\beta$ close enough to one. This proves that $I_{0}$ is $\beta$-regular (independently of (SR3) ${ }_{s}$, (SR3) ${ }_{u}$ ). Assume that $I \neq I_{0}$ satisfies $(\mathrm{SR} 3)_{s},(\mathrm{SR} 3)_{u}$ and that $(P, Q, n) \in \mathcal{R}(I)$ is $I$-bicritical. Assume also, for instance, that

$$
\begin{equation*}
\max _{I}|Q| \leqslant \max _{I}|P| \tag{9.53}
\end{equation*}
$$

and, by contradiction that

$$
\begin{equation*}
\max _{I}|P| \geqslant|I|^{\beta} . \tag{9.54}
\end{equation*}
$$

From the proof of Corollary 9 in Subsection 6.7, we know that $(P, Q, n) \in \mathcal{R}(\widetilde{I})(\widetilde{I}$ being the parent of $I$ ). As $(P, Q, n)$ is $\widetilde{I}$-bicritical, we must have

$$
\begin{equation*}
\max _{I}|P|<|\widetilde{I}|^{\beta} . \tag{9.55}
\end{equation*}
$$

Therefore, $P$ would be thin $I$-critical; similarly $Q$ would be thin $I$-critical. But (SR3) , (see (9.43), says that such a ( $P, Q, n$ ) satisfying (9.54) does not exist.

Remark. While there are only eight inequalities (SR1), (SR2) for each parameter interval I, the inequalities (SR3) ${ }_{s}$, (SR3) form a family parametrized not only by $I$, but also by the parameter intervals $I_{\alpha} \supset I$ and $I_{\omega} \supset I$ and the real number $x>0$. Because each inequality, at least when $I=I_{\alpha}$ or $I=I_{\omega}$, is only obtained after parameter selection, we will discretize the continuous variable $x$ by considering only the values $x=2^{-N}, N \geqslant 0$. There is still an infinite number of inequalities, but we will see that they are trivially satisfied if $N$ is large enough.

### 9.5 The Starting Interval

Our main purpose in this subsection is to prove the following fact.
Proposition 27. The starting interval is strongly regular.

Proof. We have already checked in Subsection 8.2 that $I_{0}$ satisfies the eight inequalities (SR1), (SR2). We have therefore to prove that $(\mathrm{SR} 3)_{s},(\mathrm{SR} 3)_{u}$ are also satisfied. We clearly have $I_{\alpha}=$ $I_{\omega}=I_{0}$. Then $B i_{+}\left(I_{0}, I_{0}, I_{0} ; x\right)$ is the set of $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$ such that $P \subset P_{s}, Q \subset Q_{u}$ and $|P| \geqslant x$ for some $t \in I_{0}$. Writing $(P, Q, n)$ as a simple composition

$$
\begin{equation*}
\left.(P, Q, n)=\left(P_{s}, Q_{s}, n_{s}\right) *\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) *\left(P_{u}, Q_{u}, n_{u}\right)\right) \tag{9.56}
\end{equation*}
$$

(here we use again assumption (H1)), we, then, use the standard estimate (8.38) to obtain:

$$
\begin{equation*}
\# B i_{+}\left(I_{0}, I_{0}, I_{0} ; x\right) \leqslant C\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\left(d_{s}^{0}+C \varepsilon_{0}\right)} \tag{9.57}
\end{equation*}
$$

and similarly

$$
\begin{equation*}
\# B i_{-}\left(I_{0}, I_{0}, I_{0} ; x\right) \leqslant C\left(\frac{x}{\varepsilon_{0}\left|Q_{s}\right|}\right)^{-\left(d_{u}^{0}+C \varepsilon_{0}\right)} . \tag{9.58}
\end{equation*}
$$

Therefore, we obtain $(\mathrm{SR} 3)_{s}$ and $(\mathrm{SR} 3)_{u}$ if we have:

$$
\begin{align*}
\rho_{0} & >d_{s}^{0}+C \varepsilon_{0},  \tag{9.5}\\
\rho_{0}^{\prime} & >d_{u}^{0}+C \varepsilon_{0}, \tag{9.60}
\end{align*}
$$

which is compatible with (9.21), (9.28).

The other part of $(\mathrm{SR} 3)_{s},(\mathrm{SR} 3)_{u}$ which can be taken care of right now is the case where $x$ is extremely small.

In this case, we will just forget about the criticality conditions for $P$ and $Q$ and bound the cardinality of $B i_{+}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ by the cardinality of the set of $(P, Q, n) \in \mathcal{R}(I)$ for which $|P| \geqslant x$ for at least some $t \in I$.

This cardinality was estimated in Proposition 22 of Subsection 8.2. Actually, the estimation was given for fixed parameter but it is easy to check that the same proof gives the same estimate as in Proposition 22 for the set we are considering. One obtains

$$
\begin{equation*}
\# B i_{+}\left(I, I_{0}, I_{0} ; x\right) \leqslant\left(\frac{x}{\varepsilon_{0}}\right)^{-d_{s}^{*}} \tag{9.61}
\end{equation*}
$$

with $d_{s}^{*}=d_{s}^{0}+\varepsilon_{0}^{\frac{1}{5} d_{s}^{0}}$ as above.

We want to have (for all $\left.I_{\alpha}, I_{\omega} \supset I\right)$

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}}\right)^{-d_{s}^{*}} \leqslant B_{0} \tag{9.62}
\end{equation*}
$$

which is equivalent to

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}}\right)^{\rho_{0}-d_{s}^{*}} \leqslant\left|P_{u}\right|^{\rho_{0}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}} . \tag{9.63}
\end{equation*}
$$

This will be satisfied if

$$
\begin{equation*}
x \leqslant x_{\min }:=|\widetilde{I}|^{C\left(\rho_{0}-d_{s}^{*}\right)^{-1}} \tag{9.64}
\end{equation*}
$$

So, we need to have

$$
\begin{equation*}
\rho_{0}>d_{s}^{*} \tag{9.65}
\end{equation*}
$$

which is compatible with $(9.59),(9.21)$. The exponent $C\left(\rho_{0}-d_{s}^{*}\right)^{-1}$ will be very large. We have proved

Proposition 28. The estimate (SR3)s is satisfied for all candidates $I$, all $I_{\alpha}, I_{\omega} \supset I$, as soon as $x \leqslant x_{\min }$. A similar statement holds for (SR3) ${ }_{u}$, with a threshold

$$
\begin{equation*}
x_{\min }^{\prime}:=|\widetilde{I}|^{C\left(\rho_{0}^{\prime}-d_{u}^{*}\right)^{-1}} \tag{9.66}
\end{equation*}
$$

### 9.6 New Bicritical Rectangles

We consider in this section the set $B i_{+}^{n e w}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ of bicritical rectangles which were not defined over the parent $\widetilde{I}$ of $I$ : cf. Subsection 9.3 . We assume that $\widetilde{I} \subset I_{\alpha} \cap I_{\omega}$. We apply to each element $(P, Q, n)$ in this set the structure theorem (Theorem 1) of Subsection 6.7. We obtain an integer $k>0$, elements $\left(P_{0}, Q_{0}, n_{0}\right), \cdots,\left(P_{k}, Q_{k}, n_{k}\right)$ of $\mathcal{R}(\widetilde{I})$ such that

$$
\begin{equation*}
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{k}, Q_{k}, n_{k}\right) \tag{9.67}
\end{equation*}
$$

Moreover, $P_{i}$ is $\widetilde{I}$-critical for $0<i \leqslant k$ and $Q_{i}$ is $\widetilde{I}$-critical for $0 \leqslant i<k$. On the other hand, $P_{0}$ is $I_{\alpha}$-critical because $P$ is $I_{\alpha}$-critical, and $Q_{k}$ is $I_{\omega}$-critical because $Q$ is $I_{\omega}$-critical. Denote by $x_{i}=2^{-n_{i}}$ the largest integral negative power of 2 such that

$$
\begin{equation*}
\left|P_{i}\right| \geqslant x_{i} \quad \text { for some } t \in \widetilde{I} \tag{9.68}
\end{equation*}
$$

Lemma 11. We have

$$
\begin{aligned}
\left(P_{0}, Q_{0}, n_{0}\right) & \in B i_{+}\left(\widetilde{I}, I_{\alpha}, \widetilde{I} ; x_{0}\right) \\
\left(P_{k}, Q_{k}, n_{k}\right) & \in B i_{+}\left(\widetilde{I}, \widetilde{I}, I_{\omega} ; x_{k}\right) \\
\left(P_{i}, Q_{i}, n_{i}\right) & \in B i_{+}\left(\widetilde{I}, \widetilde{I}, \widetilde{I} ; x_{i}\right)
\end{aligned}
$$

for $0<i<k$.

Remark. This lemma is the reason why we need to consider different levels of criticality for $P$ and $Q$.

Proof. For $0<i \leqslant k, P_{i}$ is $\widetilde{I}$-critical and it is even thin $\widetilde{I}$-critical by Corollary 7 of Subsection 6.7. Similarly, for $Q_{j}$ when $0 \leqslant j<k$. Finally, $P_{0}$ is thin $P_{\alpha}$-critical because $P$ is thin $I_{\alpha}$-critical and $P_{0}$ is the thinnest rectangle containing $P$ which is defined over $\widetilde{I}$ (which is contained in $I_{\alpha}$ ), so the thinnest rectangle containing $P$ and defined over $I_{\alpha}$ also contains $P_{0}$. Similarly, $Q_{k}$ is thin $I_{\omega}$-critical.

The widths $\left|P_{i}\right|$ are related to the width of $P$ by Corollary 6 in Subsection 6.7 which gives

$$
\begin{equation*}
x \leqslant C^{k}|I|^{-\frac{k}{2}} \prod_{0}^{k} x_{i} \tag{9.69}
\end{equation*}
$$

Let us write

$$
\begin{aligned}
\#\left(x_{0}\right) & :=\# B i_{+}\left(\widetilde{I}, I_{\alpha}, \widetilde{I} ; x_{0}\right), \\
\#\left(x_{k}\right) & :=\# B i_{+}\left(\widetilde{I}, \widetilde{I}, I_{w} ; x_{k}\right), \\
\#\left(x_{i}\right) & :=\# B i_{+}\left(\widetilde{I}, \widetilde{I}, \widetilde{I} ; x_{i}\right) \text { for } 0<i<k .
\end{aligned}
$$

Then, as each parabolic composition produces two elements, we have

$$
\begin{equation*}
\# B i_{+}^{\text {new }}\left(I, I_{\alpha}, I_{\omega} ; x\right) \leqslant \sum_{k>0} 2^{k} \sum_{x_{0}, \cdots, x_{k}} \prod_{0}^{k} \#\left(x_{i}\right) . \tag{9.70}
\end{equation*}
$$

The term $\#\left(x_{i}\right)$ is estimated by the induction hypothesis $(\mathrm{SR} 3)_{s}$ for $\widetilde{I}$. In view of the threshold (9.31), we divide $B i_{+}^{\text {new }}$ into two parts. In the first, we put the elements for which every $x_{i}$ is above the threshold $x_{i, c r}$ given by (9.31). In the second, at least one of the $x_{i}$ is below $x_{i, c r}$.

Let us consider the first part. Then all $\#\left(x_{i}\right)$ are estimated by $B_{1}$ and we have

$$
\begin{equation*}
\prod_{0}^{k} \#\left(x_{i}\right) \leqslant C^{k+1}\left(\prod_{0}^{k} x_{i}\right)^{-\rho_{1}}\left(\varepsilon_{0}\left|P_{u}\right|\right)^{(k+1) \rho_{1}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{(k+1) \sigma_{0}+k \sigma_{1}} \tag{9.71}
\end{equation*}
$$

In view of (9.69), the right-hand side is bounded by

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}} Z^{k} \tag{9.72}
\end{equation*}
$$

with

$$
\begin{equation*}
Z:=\left(C \varepsilon_{0}\left|P_{u}\right||I|^{-\frac{1}{2}}\right)^{\rho_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}} . \tag{9.73}
\end{equation*}
$$

For all $0 \leqslant i \leqslant k$, from (9.31) we have

$$
\begin{equation*}
x_{i, c r} \geqslant \varepsilon_{0}\left|P_{u}\right|\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\frac{\sigma_{0}}{\rho_{0}-\rho_{1}}} \tag{9.74}
\end{equation*}
$$

(with equality when $i \neq 0, k)$. Therefore, the number of $(k+1)$-tuples $\left(n_{0}, \cdots, n_{k}\right)$ such that $2^{-n_{i}} \geqslant x_{i, c r}$ for $0 \leqslant i \leqslant k$ is at most $\left(C \log |\widetilde{I}|^{-1}\right)^{k+1} \leqslant\left(C \log |\widetilde{I}|^{-1}\right)^{2 k}$. We conclude that the cardinality of the first part of $B i_{+}^{n e w}$ is bounded by

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}} \sum_{k>0} Z_{1}^{k}, \tag{9.75}
\end{equation*}
$$

with

$$
\begin{equation*}
Z_{1}:=2\left(C \log |\widetilde{I}|^{-1}\right)^{2} Z \tag{9.76}
\end{equation*}
$$

In view of (9.73), we have

$$
\begin{equation*}
Z_{1}<\frac{1}{2}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}-\rho_{1}}|\widetilde{I}|^{\frac{1}{3} \rho_{1}} \tag{9.77}
\end{equation*}
$$

with $\sigma_{0}+\sigma_{1}-\rho_{1}>0($ cf. (9.39)).

As $Z_{1}<\frac{1}{2}$, the bound (9.75) is smaller than

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}-\rho_{1}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}|\widetilde{I}|^{\frac{1}{3} \rho_{1}} \tag{9.78}
\end{equation*}
$$

Thus, (9.78) is a bound for the cardinality of the first part of $B i_{+}^{n e w}\left(I, I_{\alpha}, I_{\omega} ; x\right)$. Let us turn to the second part. Let $J$ be the non-empty subset of indices $i \in\{0, \cdots, k\}$ for which $x_{i}<x_{i, c r}$, and write $j=\# J$.

We first estimate the product $\prod_{i} \#\left(x_{i}\right)$. As $\rho_{0}>\rho_{1}$, we have from (9.69)

$$
\begin{equation*}
\prod_{J} x_{i}^{-\rho_{0}} \prod_{J^{c}} x_{i}^{-\rho_{1}} \leqslant\left(C^{-k}|I|^{\frac{k}{2}} x\right)^{-\rho_{0}} \tag{9.79}
\end{equation*}
$$

As we also have $\widetilde{I} \subset I_{\alpha}, \widetilde{I} \subset I_{\omega}$, we obtain

$$
\begin{equation*}
\prod_{i} \#\left(x_{i}\right) \leqslant C^{k+1}\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}\left(\frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}} Y_{0}^{j-1} Y_{1}^{k} \tag{9.80}
\end{equation*}
$$

with

$$
\begin{align*}
Y_{0} & =\left(\varepsilon_{0}\left|P_{u}\right|\right)^{\rho_{0}-\rho_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}}  \tag{9.81}\\
Y_{1} & =\left(C^{-1}|I|^{\frac{1}{2}}\right)^{-\rho_{0}}\left(\varepsilon_{0}\left|P_{u}\right|\right)^{\rho_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}} \tag{9.82}
\end{align*}
$$

We have

$$
\begin{equation*}
Y_{0}<1 \tag{9.83}
\end{equation*}
$$

and we can rewrite $Y_{1}$ as

$$
\begin{equation*}
Y_{1}=C^{\rho_{0}}\left(\frac{\widetilde{I}}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}-\frac{1}{2} \rho_{0}(1+\tau)} \varepsilon_{0}^{\rho_{1}-\frac{1}{2} \rho_{0}(1+\tau)}\left|P_{u}\right|^{\rho_{1}} \tag{9.84}
\end{equation*}
$$

The exponents satisfy

$$
\begin{gather*}
\sigma_{0}+\sigma_{1}-\frac{1}{2} \rho_{0}(1+\tau)=1-d_{u}^{0}-\frac{1}{2} d_{s}^{0}+o(1)  \tag{9.85}\\
\rho_{1}-\frac{1}{2} \rho_{0}(1+\tau)=\frac{d_{s}^{0}\left(3 d_{s}^{0}+d_{u}^{0}-2\right)}{2\left(d_{s}^{0}+d_{u}^{0}\right)}+o(1) \tag{9.86}
\end{gather*}
$$

From (H4), we have $1-d_{u}^{0}-\frac{1}{2} d_{s}^{0}>0$; from $d_{s}^{0}+d_{u}^{0} \geqslant 1, d_{s}^{0} \geqslant d_{u}^{0}$, we have $3 d_{s}^{0}+d_{u}^{0}-2 \geqslant 0$. We can, therefore, find $\sigma_{2}>0$ such that

$$
\begin{equation*}
Y_{1} \leqslant|\widetilde{I}|^{2 \sigma_{2}} \tag{9.87}
\end{equation*}
$$

Consider now the number of $(k+1)$-tuples $\left(n_{0}, \cdots, n_{k}\right)$ for this second part. From (9.69), we have

$$
\begin{equation*}
\prod_{0}^{k} x_{i} \geqslant C^{-k} x|I|^{\frac{k}{2}} \tag{9.88}
\end{equation*}
$$

and we have only to consider the case

$$
\begin{equation*}
x \geqslant x_{\min }:=|\widetilde{I}|^{C\left(\rho_{0}-d_{s}^{*}\right)^{-1}} \tag{9.89}
\end{equation*}
$$

From Corollary 7 in Subsection 6.7, we also have, for $0<i \leqslant k$,

$$
\begin{equation*}
x_{i}<|\widetilde{I}| \tag{9.90}
\end{equation*}
$$

We conclude that the number of $\left(n_{0}, \cdots, n_{k}\right)$ is smaller than

$$
\begin{equation*}
\left(C\left(\rho_{0}-d_{s}^{*}\right)^{-1} \log |\widetilde{I}|^{-1}\right)^{k+1} \tag{9.91}
\end{equation*}
$$

Using $k+1 \leqslant 2 k$ for $k>0$, we carry this to (9.70) to obtain a bound for the second part of $B i_{+}^{\text {new }}$, which is equal to

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}\left(\frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}} \sum_{k>0} Y^{k} \tag{9.92}
\end{equation*}
$$

with

$$
\begin{equation*}
Y=2 C\left(\rho_{0}-d_{s}^{*}\right)^{-2}\left(\log |\widetilde{I}|^{-1}\right)^{2} Y_{1} \tag{9.93}
\end{equation*}
$$

Here, $\left(\rho_{0}-d_{s}^{*}\right)^{-2}$ is large but independent of $\varepsilon_{0}$, which is always assumed to be as small as necessary.
In view of (9.87), we have

$$
\begin{equation*}
\sum_{k>0} Y^{k}<|\widetilde{I}|^{\sigma_{2}} \tag{9.94}
\end{equation*}
$$

We have now estimated the cardinalities of the two parts of $B i_{+}^{\text {new }}$. Taking $\sigma_{2}$ smaller if necessary, we have

$$
\begin{equation*}
0<\sigma_{2} \leqslant \frac{1}{3} \rho_{1} \tag{9.95}
\end{equation*}
$$

In (9.78), as $\sigma_{0}+\sigma_{1}-\rho_{1}>0$, we have,

$$
\begin{equation*}
\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}-\rho_{1}} \leqslant\left(\min \frac{\left|I_{\alpha}\right|}{\varepsilon_{0}}, \frac{\left|I_{\omega}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}} \tag{9.96}
\end{equation*}
$$

Thus, we have proved

Proposition 29. Assume that (SR3)s is satisfied by the parent interval $\widetilde{I}$. Then, for all $x \geqslant x_{\text {min }}$ and all candidates $I \subset \widetilde{I}$, all $I_{\alpha}, I_{\omega}$ containing $\widetilde{I}$, we have

$$
\# B i_{+}^{\text {new }}\left(I, I_{\alpha}, I_{\omega} ; x\right) \leqslant 2 B|\widetilde{I}|^{\sigma_{2}}
$$

with $B$ given by (9.18).

The presence of the term $|\widetilde{I}|^{\sigma_{2}}$ puts $B i_{+}^{\text {new }}$ well under control.
Corollary 13. Assume that $\widetilde{I} \subset I_{\alpha} \cap I_{\omega}$ and that (SR3)s holds for $\widehat{I}=I_{\alpha} \cap I_{\omega}$. Then, (SR3)s holds for all candidates $I \subset \widetilde{I}$.

Proof. An element $(P, Q, n) \in B i_{+}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ either belongs to $B i_{+}\left(\widehat{I}, I_{\alpha}, I_{\omega} ; x\right)$ or to $B i_{+}^{\text {new }}\left(I^{*}, I_{\alpha}, I_{\omega} ; x\right)$ for some interval $I^{*}$ with $\widehat{I} \varsubsetneqq I^{*} \subset I$. As the series $\sum_{I^{*}}\left|I^{*}\right|^{\sigma_{2}}$ is bounded (actually very small), the Corollary follows.

The conclusions of Proposition 29 and Corollary 13 also hold for the $B i_{-}$sets; let us review briefly the proof of Proposition 29.

In the formulas for $B_{0}^{\prime}, B_{1}^{\prime}$ the exponents $\sigma_{0}, \sigma_{1}$ are the same as for $B_{0}, B_{1}$ but we have to replace the exponents $\rho_{0}, \rho_{1}$ by $\rho_{0}^{\prime}=\frac{d_{u}^{0}}{d_{s}^{0}} \rho_{0}, \rho_{1}^{\prime}=\frac{d_{u}^{0}}{d_{s}^{0}} \rho_{1}$ (with $d_{s}^{0} \geqslant d_{u}^{0}$ ).

Therefore, we have $\sigma_{0}+\sigma_{1}-\rho_{1}^{\prime}>0$ (cf. (9.77), (9.78)), $\sigma_{0}+\sigma_{1}-\frac{1}{2} \rho_{0}^{\prime}>0$ (cf. (9.85)), $\rho_{1}^{\prime}-\frac{1}{2} \rho_{0}^{\prime} \geqslant 0$ (cf. (9.86). Therefore, Proposition 29 holds for the set $B i_{-}^{\text {new }}\left(I, I_{\alpha}, I_{\omega} ; x\right)$ with an appropriate choice of $\sigma_{2}$ and Corollary 13 holds for $(\mathrm{SR} 3)_{u}$.

The induction step in the case where $I$ is distinct from $I_{\alpha}$ and $I_{\omega}$ is complete.

### 9.7 The Case $I=I_{\alpha} \neq I_{\omega}$

We consider here the set $B i_{+}\left(I, I, I_{\omega} ; x\right)$ with $I_{\omega} \supset \widetilde{I}$. We will estimate the size of this set for most candidates $I \subset \widetilde{I}$.

Let $(P, Q, n)$ be an element of $B i_{+}\left(I, I, I_{\omega} ; x\right)$. Either it belongs to $B i_{+}^{\text {new }}\left(I, I, I_{\omega} ; x\right)$ or to $B i_{+}\left(\widetilde{I}, \widetilde{I}, I_{\omega} ; x\right)$. In the second case, as $P$ in thin $I$-critical, there exists $\left(P^{*}, Q^{*}, n^{*}\right) \in \mathcal{C}_{+}(\widetilde{I})$ such that $P \subset P^{*}$ and, moreover, we have $I \in C r\left(P^{*}\right)$ by definition of this set (Subsection 9.1). We will denote by $B i_{+}\left(P^{*}\right)$ the set of $(P, Q, n) \in B i_{+}\left(\widetilde{I}, \widetilde{I}, I_{\omega} ; x\right)$ such that $P \subset P^{*}$. We, thus, have

$$
\begin{equation*}
\# B i_{+}\left(I, I, I_{\omega} ; x\right) \leqslant \# B i_{+}^{\text {new }}\left(I, I, I_{\omega} ; x\right)+\sum_{\substack{\left(P^{*}, Q^{*}, n^{*}\right) \in \mathcal{C}_{+}(\widetilde{I}) \\ I \in C r\left(P^{*}\right)}} \# B i_{+}\left(P^{*}\right) \tag{9.97}
\end{equation*}
$$

As the sets $B i_{+}\left(P^{*}\right)$ are disjoint, we have

$$
\begin{equation*}
\sum_{\mathcal{C}_{+}(\widetilde{I})} \# B i_{+}\left(P^{*}\right) \leqslant \# B i_{+}\left(\widetilde{I}, \widetilde{I}, I_{\omega} ; x\right) . \tag{9.98}
\end{equation*}
$$

We also have

$$
\begin{equation*}
B i_{+}^{\text {new }}\left(I, I, I_{\omega} ; x\right) \subset B i_{+}^{\text {new }}\left(I, \widetilde{I}, I_{\omega} ; x\right) . \tag{9.99}
\end{equation*}
$$

We assume that $(\mathrm{SR} 3)_{s}$ holds for parameter intervals containing $\widetilde{I}$. Write $\widetilde{B}=\max \left(\widetilde{B}_{0}, \widetilde{B}_{1}\right)$ for the values of (9.18)-(9.20) with $I_{\alpha}=\widetilde{I}$. Thus, we have

$$
\begin{equation*}
\# B i_{+}\left(\widetilde{I}, \widetilde{I}, I_{\omega} ; x\right) \leqslant C \widetilde{B} \tag{9.100}
\end{equation*}
$$

and, from Proposition 29,

$$
\begin{equation*}
\# B i_{+}^{\text {new }}\left(I, \widetilde{I}, I_{\omega} ; x\right) \leqslant 2 \widetilde{B}|\widetilde{I}|^{\sigma_{2}} \tag{9.101}
\end{equation*}
$$

We now sum over candidates $I \subset \widetilde{I}$ the estimate (9.97). We obtain

$$
\begin{equation*}
\sum_{I \subset \widetilde{I}} \# B i_{+}\left(I, I, I_{\omega} ; x\right) \leqslant 2 \widetilde{B}|\widetilde{I}|^{\sigma_{2}}|\widetilde{I}|^{-\tau}+\sum_{\mathcal{C}_{+}(\widetilde{I})} \# B i_{+}\left(P^{*}\right) \# C r\left(P^{*}\right) \tag{9.102}
\end{equation*}
$$

and then, using Lemma 10 and (9.100), the same sum is bounded by

$$
\begin{equation*}
2 \widetilde{B}|\widetilde{I}|^{\sigma_{2}-\tau}+C^{\prime} \widetilde{B}|\widetilde{I}|^{-\tau d_{u}^{+}} \leqslant C^{\prime \prime} \widetilde{B}|\widetilde{I}|^{-\tau d_{u}^{+}} \tag{9.103}
\end{equation*}
$$

Write $B=\max \left(B_{0}, B_{1}\right)$ for the values of (9.18)-(9.20) with $I_{\alpha}=I$. The number of candidates $I$ such that $\# B i_{+}\left(I, I, I_{\omega} ; x\right)>B$ is at most

$$
\begin{equation*}
C^{\prime \prime} \widetilde{B} B^{-1}|\widetilde{I}|^{-\tau d_{u}^{+}} . \tag{9.104}
\end{equation*}
$$

From formulas (9.18)-(9.20), we see that

$$
\begin{equation*}
\widetilde{B} B^{-1}=|\widetilde{I}|^{-\tau\left(\sigma_{0}+\sigma_{1}\right)} \tag{9.105}
\end{equation*}
$$

The exponents $\sigma_{0}, \sigma_{1}$ will be chosen in order to have

$$
\begin{equation*}
\sigma_{0}+\sigma_{1}+d_{u}^{+}<1-2 \tau \tag{9.106}
\end{equation*}
$$

which is compatible with $(9.23),(9.24)$. Then, we obtain that (SR3) ${ }_{s}$ holds except for a proportion of candidates no greater than $C|\widetilde{I}|^{2 \tau^{2}}$. This assertion has been proved for a fixed value of $x$ and a fixed parameter interval $I_{\omega}$. But (in view of the constant $C$ in $\left.(\mathrm{SR} 3)_{s}\right)$, it is sufficient to prove $(\mathrm{SR} 3)_{s}$ when $x=2^{-n}, n$ a nonnegative integer, and $x \geqslant x_{\min }$; and the number of intervals $I_{\omega} \supset \widetilde{I}$ is at most

$$
\begin{equation*}
C \tau^{-1} \log \left(\frac{\log |\widetilde{I}|}{\log \varepsilon_{0}}\right) \tag{9.107}
\end{equation*}
$$

Thus, the total number of cases that we have to consider is much smaller than $|\widetilde{I}|^{-\tau^{2}}$. We obtain

Proposition 30. Assume that (SR3)s holds for parameter intervals containing $\widetilde{I}$. Then, (SR3)s holds for $B i_{+}\left(I, I, I_{\omega} ; x\right)$, for all $I_{\omega} \supset \widetilde{I}$ and all x, except perhaps for a proposition of candidates no greater than $|\widetilde{I}|^{-\tau^{2}}$.

Proposition 30 is also valid exchanging $I_{\alpha}$ and $I_{\omega}$. There are also two similar statements involving $(\mathrm{SR} 3)_{u}$ and $B i_{-}$. We review briefly the slight difference in the proof.

For $B i_{-}\left(I, I, I_{\omega} ; x\right)$ (with $\left.I_{\omega} \supset \widetilde{I}\right)$, we proceed exactly as for $B i_{+}\left(I, I, I_{\omega} ; x\right)$. For $B i_{+}\left(I, I_{\alpha}, I ; x\right)$, we use $\mathcal{C}_{-}(\widetilde{I})$ instead of $\mathcal{C}_{+}(\widetilde{I})$ to subdivide the set of bicritical elements.

We now have, by the dual version of Lemma 10

$$
\begin{equation*}
\# C r\left(Q^{*}\right) \leqslant C|\widetilde{I}|^{-\tau d_{s}^{+}} \tag{9.108}
\end{equation*}
$$

On the other hand, from formulas (9.18)-(9.20), we must replace (9.105) by

$$
\begin{equation*}
\widetilde{B} B^{-1}=|\widetilde{I}|^{-\tau \sigma_{0}} \tag{9.109}
\end{equation*}
$$

Thus, we will choose $\sigma_{0}$ in order to have

$$
\begin{equation*}
\sigma_{0}+d_{s}^{+}<1-2 \tau \tag{9.110}
\end{equation*}
$$

Obviously, this is compatible with $(9.23),(9.24),(9.106)$. The end of the argument is the same as before.

This completes the proof of the induction step except for the case $I=I_{\alpha}=I_{\omega}$.

In this last case, we can actually apply the same argument as above to complete the induction step when $x$ is large. When $I_{\alpha}=I_{\omega}$ is equal to $I$ or $\widetilde{I}$, indeed we have

$$
\begin{align*}
B_{0}(x) & =\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}}  \tag{9.111}\\
B_{1}(x) & =\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}  \tag{9.112}\\
\widetilde{B}_{0}(x) & =\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}}  \tag{9.113}\\
\widetilde{B}_{1}(x) & =\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}} \tag{9.114}
\end{align*}
$$

For $x \geqslant \widetilde{x}_{c r}$, we, therefore, have

$$
\begin{equation*}
\widetilde{B}(x) B(x)^{-1}=|\widetilde{I}|^{-\tau\left(\sigma_{0}+\sigma_{1}\right)} \tag{9.115}
\end{equation*}
$$

and the same argument as before applies.

For $x \leqslant x_{c r}$, we have

$$
\begin{equation*}
\widetilde{B}(x) B(x)^{-1}=|\widetilde{I}|^{-\tau\left(2 \sigma_{0}+\sigma_{1}\right)} \tag{9.116}
\end{equation*}
$$

with $2 \sigma_{0}+\sigma_{1}=2-d_{s}^{0}-d_{u}^{0}+o(1)$ : this case will be the object of the rest of Section 9.
In the intermediate range $x_{c r} \leqslant x \leqslant \widetilde{x}_{c r}$, we write $x=\widetilde{x}_{c r} y$, where we have, in view of (9.31),

$$
\begin{equation*}
1 \geqslant y \geqslant|\widetilde{I}|^{\frac{\tau \sigma_{0}}{\rho_{0}-\rho_{1}}} \tag{9.117}
\end{equation*}
$$

We now have in this range:

$$
\begin{align*}
\widetilde{B}(x) B(x)^{-1} & =\widetilde{B}_{0}(x) B_{1}(x)^{-1}  \tag{9.118}\\
& =y^{\rho_{1}-\rho_{0}}|\widetilde{I}|^{-\tau\left(\sigma_{0}+\sigma_{1}\right)}
\end{align*}
$$

Therefore, the proportion of bad candidates will not be greater than $C|\widetilde{I}|^{2 \tau^{2}}$, as long as we have

$$
\begin{equation*}
y^{\rho_{0}-\rho_{1}} \geqslant|\widetilde{I}|^{\tau\left(1-\sigma_{0}-\sigma_{1}-d_{u}^{+}-2 \tau\right)} \tag{9.119}
\end{equation*}
$$

We recall that $\rho_{0}-\rho_{1}>0$ and that $1-\sigma_{0}-\sigma_{1}-d_{u}^{+}-2 \tau$ is positive according to (9.106), but it has to be small according to $(9.23),(9.24)$. We state

Proposition 31. Assume that (SR3)s holds for all parameter intervals containing $\widetilde{I}$. Then, it holds for $B i_{+}(I, I, I ; x)$ in the range

$$
\begin{equation*}
x \geqslant \widetilde{x}_{c r}|\widetilde{I}|^{\frac{\tau}{\rho_{0}-\rho_{1}}}\left(1-\sigma_{0}-\sigma_{1}-d_{u}^{+}-2 \tau\right) \tag{9.120}
\end{equation*}
$$

except for a proportion of candidates no greater than $|\widetilde{I}|^{\tau^{2}}$.
For $B i_{-}(I, I, I ; x)$, the corresponding range is

$$
\begin{equation*}
x \geqslant \widetilde{x}_{c r}^{\prime}|\widetilde{I}|^{\frac{\tau}{\rho_{0}^{\prime}-\rho_{1}^{\prime}}\left(1-\sigma_{0}-\sigma_{1}-d_{u}^{+}-2 \tau\right)} \tag{9.121}
\end{equation*}
$$

### 9.8 The Case $I=I_{\alpha}=I_{\omega}$ : General Overview

We explain in this subsection the strategy which will be pursued in the case $I=I_{\alpha}=I_{\omega}, x$ small.

The exponent $2 \sigma_{0}+\sigma_{1}$ in (9.116) means that we have take into account the criticality of both $P$ and $Q$ in the selection process.

We start as before. An element $(P, Q, n)$ in $B i_{+}(I, I, I ; x)$ is either in $B i_{+}^{\text {new }}(I, I, I ; x)$ or in $B i_{+}(\widetilde{I}, \widetilde{I}, \widetilde{I} ; x)$. The cardinality of the first set is bounded as before by

$$
\begin{equation*}
\# B i_{+}^{n e w}(I, I, I ; x) \leqslant 2 \widetilde{B}|\widetilde{I}|^{\sigma_{2}} \tag{9.122}
\end{equation*}
$$

On the other hand, let $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I})$ and $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{C}_{-}(\widetilde{I})$. We denote by $B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$ the set of $(P, Q, n) \in B i_{+}(\widetilde{I}, \widetilde{I}, \widetilde{I} ; x)$ such that $P \subset P_{\alpha}$ and $Q \subset Q_{\omega}$. These subsets are disjoint and their union contain the elements $(P, Q, n)$ of $B i_{+}(I, I, I ; x)$ which belong to $\mathcal{R}(\widetilde{I})$ because $P$ and
$Q$ are thin $I$-critical. Moreover, only those $P_{\alpha}, Q_{\omega}$ such that $I \in C r\left(P_{\alpha}\right) \cap C r\left(Q_{\omega}\right)$ will actually appear.

Thus, in analogy with (9.97), we have

$$
\begin{equation*}
\# B i_{+}(I, I, I ; x) \leqslant 2 \widetilde{B}|\widetilde{I}|^{\sigma_{2}}+\sum \# B i_{+}\left(P_{\alpha}, Q_{\omega}\right) \tag{9.123}
\end{equation*}
$$

where the sum is over pairs $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I}),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{C}_{-}(\widetilde{I})$ such that $I \subset C r\left(P_{\alpha}\right)$ and $I \subset C r\left(Q_{\omega}\right)$.

Unfortunately, we are not able to estimate the size of the intersection $\operatorname{Cr}\left(P_{\alpha}\right) \cap \operatorname{Cr}\left(Q_{\omega}\right)$ directly in a satisfactory way: while these two sets seem "independent" in the naive sense, it is another matter to translate this intuition in a quantitative estimate in the style of Lemma 10.

Instead of this approach, we will use some degree of independence, when $x$ is small, of the variables $P_{\alpha}$ and $Q_{\omega}$ in $B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$. To explain the technique, consider first the unrealistic model case where we would have

$$
\begin{equation*}
\# B i_{+}\left(P_{\alpha}, Q_{\omega}\right)=b_{+}\left(P_{\alpha}\right) b_{-}\left(Q_{\omega}\right) \tag{9.124}
\end{equation*}
$$

for some functions $b_{+}, b_{-}$on $\mathcal{C}_{+}(\widetilde{I}), \mathcal{C}_{-}(\widetilde{I})$, respectively.

The formula (9.123) now gives

$$
\begin{equation*}
\# B i_{+}(I, I, I ; x) \leqslant 2 \widetilde{B}|\widetilde{I}|^{\sigma_{2}}+\phi_{+}(I) \phi_{-}(I) \tag{9.125}
\end{equation*}
$$

with

$$
\begin{align*}
& \phi_{+}(I)=\sum_{I \in C r\left(P_{\alpha}\right)} b_{+}\left(P_{\alpha}\right),  \tag{9.126}\\
& \phi_{-}(I)=\sum_{I \in C r\left(Q_{\omega}\right)} b_{-}\left(Q_{\omega}\right) . \tag{9.127}
\end{align*}
$$

We now average separately $\phi_{+}$and $\phi_{-}$. We obtain

$$
\begin{align*}
\sum_{I} \phi_{+}(I) & \leqslant\left(\max _{\mathcal{C}_{+}(\tilde{I})} \# \operatorname{Cr}\left(P_{\alpha}\right)\right) \sum_{\mathcal{C}_{+}(\widetilde{I})} b_{+}\left(P_{\alpha}\right),  \tag{9.128}\\
\sum_{I} \phi_{-}(I) & \leqslant\left(\max _{\mathcal{C}_{-}(\widetilde{I})} \# \operatorname{Cr}\left(Q_{\omega}\right)\right) \sum_{\mathcal{C}_{-}(\widetilde{I})} b_{-}\left(Q_{\omega}\right), \tag{9.129}
\end{align*}
$$

where $C r\left(P_{\alpha}\right)$ and $C r\left(Q_{\omega}\right)$ are estimated by Lemma 10 and we have

$$
\begin{equation*}
\sum b_{+}\left(P_{\alpha}\right) \sum b_{-}\left(Q_{\omega}\right)=\sum \# B i_{+}\left(P_{\alpha}, Q_{\omega}\right) \leqslant \widetilde{B} \tag{9.130}
\end{equation*}
$$

It is then sufficient to eliminate candidates for which either $\phi_{+}$or $\phi_{-}$is much above its average value to be able to conclude the proof.

As (9.124) does not hold, we will find an appropriate substitute as follows.
We will subdivide each class $B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$ into subclasses $B i_{+}\left(P_{\alpha}, Q_{\omega}, \ell\right)$; the index $\ell$ runs through a finite large set $L$ dependent on $I$ and $x$ but independent on $P_{\alpha}$ and $Q_{\omega}$. Moreover, we will have functions $b_{+}\left(P_{\alpha}, \ell\right), b_{-}\left(Q_{\omega}, \ell\right)$ on $\mathcal{C}_{+}(\widetilde{I}) \times L, \mathcal{C}_{-}(\widetilde{I}) \times L$, respectively, such that,

$$
\begin{equation*}
\# B i_{+}\left(P_{\alpha}, Q_{\omega}, \ell\right) \leqslant b_{+}\left(P_{\alpha}, \ell\right), b_{-}\left(Q_{\omega}, \ell\right) \tag{9.131}
\end{equation*}
$$

We then set, for each $\ell \in L$ :

$$
\begin{align*}
\phi_{+, \ell}(I) & =\sum_{I \in C r\left(P_{\alpha}\right)} b_{+}\left(P_{\alpha}, \ell\right),  \tag{9.132}\\
\phi_{-, \ell}(I) & =\sum_{I \in C r\left(Q_{\omega}\right)} b_{-}\left(Q_{\omega}, \ell\right) . \tag{9.133}
\end{align*}
$$

We average each of these functions to get, in view of Lemma 10,

$$
\begin{align*}
\sum_{I} \phi_{+, \ell}(I) & \leqslant C|\widetilde{I}|^{-\tau d_{u}^{+}} b_{+}(\ell),  \tag{9.134}\\
\sum_{I} \phi_{-, \ell}(I) & \leqslant C|\widetilde{I}|^{-\tau d_{s}^{+}} b_{-}(\ell), \tag{9.135}
\end{align*}
$$

with

$$
\begin{align*}
b_{+}(\ell) & =\sum_{\mathcal{C}_{+}(\widetilde{I})} b_{+}\left(P_{\alpha}, \ell\right),  \tag{9.136}\\
b_{-}(\ell) & =\sum_{\mathcal{C}_{-}(\widetilde{I})} b_{-}\left(Q_{\omega}, \ell\right) . \tag{9.137}
\end{align*}
$$

For each $\ell$, we will have

$$
\begin{align*}
& \phi_{+, \ell}(I) \leqslant|\widetilde{I}|^{\tau\left(1-d_{u}^{+}-3 \tau\right)} b_{+}(\ell)  \tag{9.138}\\
& \phi_{-, \ell}(I) \leqslant|\widetilde{I}|^{\tau\left(1-d_{s}^{+}-3 \tau\right)} b_{-}(\ell) \tag{9.139}
\end{align*}
$$

except for a proportion of candidates not greater than $C|\widetilde{I}|^{3 \tau^{2}}$. Set

$$
\begin{equation*}
\widehat{B}=\sum_{L} b_{+}(\ell) b_{-}(\ell) . \tag{9.140}
\end{equation*}
$$

Because we need to eliminate candidates for each $\ell, L$ should not be too large. We will have, see Proposition 33 in Subsection 9.10, that

$$
\begin{equation*}
\# L \leqslant C|\widetilde{I}|^{-\tau^{2}} \tag{9.141}
\end{equation*}
$$

Taking into account that we must eliminate candidates for each $x=2^{-n} \geqslant x_{\min }$, the total proportion of the failed candidates is at most $|\widetilde{I}|^{\tau^{2}}$. On the other hand, for the surviving candidates, the discussion above gives

$$
\begin{align*}
\sum_{I \in C r\left(P_{\alpha}\right) \cap C r\left(Q_{\omega}\right)} \# B i_{+}\left(P_{\alpha}, Q_{\omega}\right) & \leqslant \sum_{L} \sum_{I \in C r\left(P_{\alpha}\right)} \sum_{I \in C r\left(Q_{\omega}\right)} b_{+}\left(P_{\alpha}, \ell\right) b_{-}\left(Q_{\omega}, \ell\right)  \tag{9.142}\\
& =\sum_{L} \phi_{+, \ell}(I) \phi_{-, \ell}(I) \leqslant|\widetilde{I}|^{\tau\left(2-d_{s}^{\left.+-d_{\omega}^{+}-6 \tau\right)} \widehat{B}\right.}
\end{align*}
$$

$$
\begin{equation*}
\# B i_{+}(I, I, I ; x) \leqslant 2 \widetilde{B}|\widetilde{I}|^{\sigma_{2}}+\widehat{B}|\widetilde{I}|^{\tau\left(2-d_{s}^{+}-d_{u}^{+}-6 \tau\right)} \tag{9.143}
\end{equation*}
$$

We definitely have $\widetilde{B}|\widetilde{I}|^{\sigma_{2}} \ll B$. In order to obtain (SR3) ${ }_{s}$, we need

$$
\begin{equation*}
\widehat{B}|\widetilde{I}|^{\tau\left(2-d_{s}^{+}-d_{u}^{+}-6 \tau\right)} \leqslant C B \tag{9.144}
\end{equation*}
$$

As $2-d_{s}^{+}-d_{u}^{+}-6 \tau=2 \sigma_{0}+\sigma_{1}+o(1)$, we see from (9.116) that $\widehat{B}$ cannot be much larger than $\widetilde{B}$.
In the next three subsections, we will

- define precisely $L$ and the subclasses $B i_{+}\left(P_{\alpha}, Q_{\omega}, \ell\right)$;
- bound the cardinality of $L$ (to obtain (9.141));
- obtain an appropriate estimate for $\widehat{B}$ (cf. (9.144)).


### 9.9 Subclasses of Bicritical Elements

### 9.9.1 Bound elements.

Definition. Let $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I}),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{C}_{-}(\widetilde{I})$. An element $(P, Q, n) \in B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$ is bound if $n \leqslant n_{\alpha}+n_{\omega}$. Otherwise, we say that $(P, Q, n)$ is free. We will denote by $B i_{+}\left(P_{\alpha}, Q_{\alpha}, \diamond\right)$ the subset of bound elements of $B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$.

Thus, $\diamond$ is an element of $L$. On the other hand, free elements will correspond to many elements of $L$. Recall that we have $x \leqslant \widetilde{x}_{c r}$. When $x \ll \widetilde{x}_{c r}$, most elements are free. When $x \gg \widetilde{x}_{c r}$, on the opposite, most elements are bound.

Proposition 32. For any $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{R}(\widetilde{I})$, and any $n \leqslant n_{\alpha}+n_{\omega}$, there is at most one element $(P, Q, n) \in \mathcal{R}(\widetilde{I})$ of length $n$ such that $P \subset P_{\alpha}, Q \subset Q_{\omega}$.

Proof. We argue by induction on the level of the parameter interval.
When $\widetilde{I}$ is the starting interval $I_{0}$, no parabolic composition is involved and the result follows from usual symbolic dynamics: as $n \leqslant n_{\alpha}+n_{\omega}$, the word associated to a bound element is determined by its initial and final parts.

Assume that the result holds for parameter intervals strictly larger than $\widetilde{I}$. Denote by $\widetilde{I}_{1}$ the parent interval of $\widetilde{I}$.

Assume first that both $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ and $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$ belong to $\mathcal{R}\left(\widetilde{I}_{1}\right)$. We claim that any bound element also belongs to $\mathcal{R}\left(\widetilde{I}_{1}\right)$, which allow us to conclude the proof by the induction hypothesis. Indeed, if $(P, Q, n)$ satisfies $P \subset P_{\alpha}, Q \subset Q_{\omega}$ and does not belong to $\mathcal{R}\left(\widetilde{I}_{1}\right)$, we apply the structure
theorem in Subsection 6.7: it gives elements $\left(P_{0}, Q_{0}, n_{0}\right),\left(P_{k}, Q_{k}, n_{k}\right) \in \mathcal{R}\left(\widetilde{I}_{1}\right)$ such that $P_{0}$ is the thinnest rectangle containing $P$ defined over $\widetilde{I}_{1}, Q_{k}$ is the thinnest rectangle containing $Q$ defined over $\widetilde{I}_{1}$, and $n \geqslant n_{0}+n_{k}+N_{0}$. Therefore, $n_{0} \geqslant n_{\alpha}, n_{k} \geqslant n_{\omega}$ and $n>n_{\alpha}+n_{\omega}$.

We now consider the case when, for instance, $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ does not belong to $\mathcal{R}\left(\widetilde{I}_{1}\right)$. We now apply the structure theorem of Subsection 6.7 to $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ and also to an element $(P, Q, n) \in \mathcal{R}(\widetilde{I})$ with $P \subset P_{\alpha}, Q \subset Q_{\omega}, n \leqslant n_{\alpha}+n_{\omega}$. We obtain integers $0<j \leqslant k$, elements $\left(P_{i}, Q_{i}, n_{i}\right) \in \mathcal{R}\left(\widetilde{I}_{1}\right)$ for $0 \leqslant i \leqslant k$ such that

$$
\begin{equation*}
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{k}, Q_{k}, n_{k}\right) \tag{9.145}
\end{equation*}
$$

and also $\left(\widetilde{P}_{j}, \widetilde{Q}_{j}, \widetilde{n}_{j}\right) \in \mathcal{R}\left(\widetilde{I}_{1}\right)$ such that

$$
\begin{equation*}
\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{j-1}, Q_{j-1}, n_{j-1}\right) \square\left(\widetilde{P}_{j}, \widetilde{Q}_{j}, \widetilde{n}_{j}\right) \tag{9.146}
\end{equation*}
$$

There also exists $m$ with $0 \leqslant m \leqslant j$ and $\left(\widetilde{P}_{m}^{\prime}, \widetilde{Q}_{m}^{\prime}, \widetilde{n}_{m}^{\prime}\right)$ in $\mathcal{R}\left(\widetilde{I}_{1}\right)$ such that

- either $m=j=k,\left(\widetilde{P}_{m}^{\prime}, \widetilde{Q}_{m}^{\prime}, \widetilde{n}_{m}^{\prime}\right)=\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$;
- or $m<k$ and we have,

$$
\begin{equation*}
\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in\left(\widetilde{P}_{m}^{\prime}, \widetilde{Q}_{m}^{\prime}, \widetilde{n}_{m}^{\prime}\right) \square \cdots \square\left(P_{k}, Q_{k}, n_{k}\right) . \tag{9.147}
\end{equation*}
$$

If $m<j$, the sequence $\left(P_{i}, Q_{i}, n_{i}\right)$ and the choice of the result (out of two possibilities) in each parabolic composition are completely determined by $P_{\alpha}$ and $Q_{\omega}$ : the assertion of the proposition follows.

When $m=j$, the sequence $\left(P_{i}, Q_{i}, n_{i}\right)$ for $i \neq m=j$ is determined by $P_{\alpha}, Q_{\omega}$; but we also have $P_{j} \subset \widetilde{P}_{j}, Q_{j}=Q_{m} \subset \widetilde{Q}_{m}$ and $n_{j} \leqslant \widetilde{n}_{j}+\widetilde{n}_{m}^{\prime}$, so by the induction hypothesis ( $P_{j}, Q_{j}, n_{j}$ ) is also determined by $P_{\alpha}, Q_{\omega}$. Again, the choices of the results in the parabolic compositions are also determined by $P_{\alpha}, Q_{\omega}$. The proof of the proposition is complete.

Recall that, by Proposition 13 in Subsection 7.1, we have

$$
\begin{equation*}
|P| \leqslant C \exp \left(-n^{\gamma}\right) \tag{9.148}
\end{equation*}
$$

for any $(P, Q, n) \in \mathcal{R}(\widetilde{I})$, any $t \in \widetilde{I}$, with $\gamma=\log \frac{3}{2} / \log 2$. Any bound element in $B i_{+}\left(P_{\alpha}, Q_{\omega}, \diamond\right)$ must satisfy $|P| \geqslant x$ for some $t \in \widetilde{I}$, and we therefore have

$$
\begin{equation*}
n \leqslant\left(\log \frac{C}{x}\right)^{\frac{1}{\gamma}} \tag{9.149}
\end{equation*}
$$

For $x \geqslant x_{\min }$, we have

$$
\begin{equation*}
\left(\log \frac{C}{x}\right)^{\frac{1}{\gamma}} \leqslant(\log |\widetilde{I}|)^{2} \tag{9.150}
\end{equation*}
$$

We thus shall define

$$
\begin{equation*}
b_{+}\left(P_{\alpha}, \diamond\right)=b_{-}\left(Q_{\omega}, \diamond\right)=\log |\widetilde{I}|^{-1} \tag{9.151}
\end{equation*}
$$

and we will indeed have, from Proposition 32,

$$
\begin{equation*}
\# B i_{+}\left(P_{\alpha}, Q_{\omega}, \diamond\right) \leqslant b_{+}\left(P_{\alpha}, \diamond\right) b_{-}\left(Q_{\omega}, \diamond\right) \tag{9.152}
\end{equation*}
$$

### 9.9.2 Decomposition of a free element.

Let $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I}),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{C}_{-}(\widetilde{I})$ and let $(P, Q, n) \in B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$ be a free element. We will analyze with respect to the structure theorem of Subsection 6.7 the way in which ( $P, Q, n$ ), $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$ have been created. This will allow us in the sequel to define various subclasses of free elements.

Denote by $\widehat{I}_{0}$ the largest parameter interval such that $(P, Q, n) \in \mathcal{R}\left(\widehat{I}_{0}\right)$. Elements $(P, Q, n)$ for which $\widehat{I}_{0}$ is the starting interval $I_{0}$ are said to have depth 0 . They form a first subclass of $B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$ denoted by $B i_{+}\left(P_{\alpha}, Q_{\omega}, 0\right)$.

We now assume that $\widehat{I}_{0} \neq I_{0}$ and denote by $\widetilde{I}_{0}$ the parent interval of $\widehat{I}_{0}$. We apply the structure theorem of Subsection 6.7. We obtain an integer $k>0$, elements $\left(P_{0}, Q_{0}, n_{0}\right), \cdots,\left(P_{k}, Q_{k}, n_{k}\right)$ in $\mathcal{R}\left(\widetilde{I}_{0}\right)$ such that

$$
\begin{equation*}
(P, Q, n) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(P_{k}, Q_{k}, n_{k}\right) . \tag{9.153}
\end{equation*}
$$

As in the proof of Proposition 32, we find $0 \leqslant j \leqslant k$ and $\left(\widetilde{P}_{j}, \widetilde{Q}_{j}, \widetilde{n}_{j}\right) \in \mathcal{R}\left(\widetilde{I}_{0}\right)$ such that either $j=0,\left(\widetilde{P}_{j}, \widetilde{Q}_{j}, \widetilde{n}_{j}\right)=\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ (if $\left.\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{R}\left(\widetilde{I}_{0}\right)\right)$ or $j>0$ and

$$
\begin{equation*}
\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in\left(P_{0}, Q_{0}, n_{0}\right) \square \cdots \square\left(\widetilde{P}_{j}, \widetilde{Q}_{j}, \widetilde{n}_{j}\right) \tag{9.154}
\end{equation*}
$$

Similarly, we find $0 \leqslant m \leqslant k$ and $\left(\widetilde{P}_{m}^{\prime}, \widetilde{Q}_{m}^{\prime}, \widetilde{n}_{m}^{\prime}\right) \in \mathcal{R}\left(\widetilde{I}_{0}\right)$ such that either $m=k,\left(\widetilde{P}_{m}^{\prime}, \widetilde{Q}_{m}^{\prime}, \widetilde{n}_{m}^{\prime}\right)=$ ( $P_{\omega}, Q_{\omega}, n_{\omega}$ ) or $m<k$ and

$$
\begin{equation*}
\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in\left(\widetilde{P}_{m}^{\prime}, \widetilde{Q}_{m}^{\prime}, \widetilde{n}_{m}^{\prime}\right) \square \cdots \square\left(P_{k}, Q_{k}, n_{k}\right) . \tag{9.155}
\end{equation*}
$$

We also must have $P_{j} \subset \widetilde{P}_{j}, Q_{m} \subset \widetilde{Q}_{m}^{\prime}$. Moreover, as $(P, Q, n)$ is free, we must have $j \leqslant m$ and, when $j=m$, we must also have $n_{j}=n_{m}>\widetilde{n}_{j}+\widetilde{n}_{m}^{\prime}$.

We say that $(P, Q, n)$ is fully decomposed if one has here $j<m$ or $j=m$ and $\left(P_{j}, Q_{j}, n_{j}\right) \in \mathcal{R}\left(I_{0}\right)$. Such elements are said to have depth one.

Assume that $(P, Q, n)$ is not fully decomposed. Then, we have $j=m, P_{j} \subset \widetilde{P}_{j}, Q_{j} \subset \widetilde{Q}_{j}^{\prime}$ and the largest parameter interval $\widehat{I}_{1}$ for which $\left(P_{j}, Q_{j}, n_{j}\right) \in \mathcal{R}\left(\widehat{I}_{1}\right)$ is not the starting interval $I_{0}$. We denote by $\widetilde{I}_{1}$ the parent interval. We rewrite

$$
\begin{align*}
\left(P^{1}, Q^{1}, n^{1}\right) & :=\left(P_{j}, Q_{j}, n_{j}\right),  \tag{9.156}\\
\left(P_{\alpha}^{1}, Q_{\alpha}^{1}, n_{\alpha}^{1}\right) & :=\left(\widetilde{P}_{j}, \widetilde{Q}_{j}, \widetilde{n}_{j}\right), \\
\left(P_{\omega}^{1}, Q_{\omega}^{1}, n_{\omega}^{1}\right) & :=\left(\widetilde{P}_{j}^{\prime}, \widetilde{Q}_{j}^{\prime}, \widetilde{n}_{j}^{\prime}\right),
\end{align*}
$$

and proceed with these elements as we did with $(P, Q, n),\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$ : we will find integers $0 \leqslant j_{1} \leqslant m_{1} \leqslant k_{1}$ (with $k_{1}>0$ ), elements $\left(P_{i}^{1}, Q_{i}^{1}, n_{i}^{1}\right)$ for $0 \leqslant i \leqslant k_{1}$ and also ( $P_{\alpha}^{2}, Q_{\alpha}^{2}, n_{\alpha}^{2}$ ), $\left(P_{\omega}^{2}, Q_{\omega}^{2}, n_{\omega}^{2}\right)$, all in $\mathcal{R}\left(\widetilde{I}_{1}\right)$, such that

$$
\begin{align*}
\left(P^{1}, Q^{1}, n^{1}\right) & \in\left(P_{0}^{1}, Q_{0}^{1}, n_{0}^{1}\right) \square \cdots \square\left(P_{k_{1}}^{1}, Q_{k_{1}}^{1}, n_{k_{1}}^{1}\right),  \tag{9.157}\\
\left(P_{\alpha}^{1}, Q_{\alpha}^{1}, n_{\alpha}^{1}\right) & \in\left(P_{0}^{1}, Q_{0}^{1}, n_{0}^{1}\right) \square \cdots \square\left(P_{j_{1}-1}^{1}, Q_{j_{1}-1}^{1}, n_{j_{1}-1}^{1}\right) \square\left(P_{\alpha}^{2}, Q_{\alpha}^{2}, n_{\alpha}^{2}\right), \\
\left(P_{\omega}^{1}, Q_{\omega}^{1}, n_{\omega}^{1}\right) & \in\left(P_{\omega}^{2}, Q_{\omega}^{2}, n_{\omega}^{2}\right) \square\left(P_{m_{1}+1}^{1}, Q_{m_{1}+1}^{1}, n_{m_{1}+1}^{1}\right) \square \cdots \square\left(P_{k_{1}}^{1}, Q_{k_{1}}^{1}, n_{k_{1}}^{1}\right) .
\end{align*}
$$

Again, we say that $\left(P^{1}, Q^{1}, n^{1}\right)$ is fully decomposed if either $j_{1}<m_{1}$ or $j_{1}=m_{1}$ and $\left(P_{j_{1}}^{1}, Q_{j_{1}}^{1}, n_{j_{1}}^{1}\right)$ is defined over the starting interval $I_{0}$; otherwise we set

$$
\begin{equation*}
\left(P^{2}, Q^{2}, n^{2}\right):=\left(P_{j_{1}}^{1}, Q_{j_{1}}^{1}, n_{j_{1}}^{1}\right), \tag{9.158}
\end{equation*}
$$

and we go on. The sequence of parameter intervals $\widehat{I}_{0} \subset \widehat{I}_{1} \subset \cdots$ is strictly increasing and therefore the process will stop. We define inductively the depth of $(P, Q, n)$ to be the depth of $\left(P^{1}, Q^{1}, n^{1}\right)$ plus one.

### 9.9.3 Size of the subclass of depth 0

We will define in this subsection $b_{+}\left(P_{\alpha}, 0\right), b_{-}\left(Q_{\omega}, 0\right)$ in order to have

$$
\begin{equation*}
\# B i_{+}\left(P_{\alpha}, Q_{\omega}, 0\right) \leqslant b_{+}\left(P_{\alpha}, 0\right) b_{-}\left(Q_{\omega}, 0\right) \tag{9.159}
\end{equation*}
$$

Let $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I}),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{C}_{-}(\widetilde{I})$ and let $(P, Q, n) \in B i_{+}\left(P_{\alpha}, Q_{\omega}, 0\right)$. Then $(P, Q, n)$ is obtained by a simple composition

$$
\begin{equation*}
(P, Q, n)=\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) *\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) *\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \tag{9.160}
\end{equation*}
$$

We have here, for all $t \in \widetilde{I}$

$$
\begin{align*}
|P| & \leqslant C\left|P_{\alpha}\right|\left|P^{\prime}\right|\left|P_{\omega}\right|  \tag{9.161}\\
\left|P_{\alpha}\right| & \leqslant|\widetilde{I}|^{1+\tau}(\text { cf. Subsection } 9.1)
\end{align*}
$$

and also, for some $t_{0} \in \widetilde{I}$

$$
\begin{equation*}
|P| \geqslant x \tag{9.163}
\end{equation*}
$$

This gives, for this value $t_{0}$ :

$$
\begin{equation*}
\left|P^{\prime}\right| \geqslant C^{-1}|\widetilde{I}|^{-(1+\tau)}\left(\max _{\widetilde{I}}\left|P_{\omega}\right|\right)^{-1} x \tag{9.164}
\end{equation*}
$$

We observe that, as $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$ belongs to $\mathcal{R}\left(I_{0}\right)$ and $\left|Q_{\omega}\right|$ is of the order of $|\widetilde{I}|^{1+\tau}$, we have

$$
\begin{equation*}
\max _{\widetilde{I}}\left|P_{\omega}\right| \leqslant C \min _{\widetilde{I}}\left|P_{\omega}\right| . \tag{9.165}
\end{equation*}
$$

From the estimate (8.38) in Subsection 8.2, we can thus define, as $d_{s}^{0}+C \varepsilon_{0}<d_{s}^{*}$,

$$
b_{+}\left(P_{\alpha}, 0\right)= \begin{cases}\left(C|\widetilde{I}|^{1+\tau} x^{-1}\right)^{d_{s}^{*}} & \text { if }\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{R}\left(I_{0}\right),  \tag{9.166}\\ 0 & \text { otherwise },\end{cases}
$$

$$
b_{-}\left(Q_{\omega}, 0\right)= \begin{cases}\left(\min _{\widetilde{I}}\left|P_{\omega}\right|\right)^{d_{s}^{*}} & \text { if }\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{R}\left(I_{0}\right),  \tag{9.167}\\ 0 & \end{cases}
$$

Then, (9.159) is satisfied.

### 9.9.4 Subclasses of higher depth

Let $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{C}_{+}(\widetilde{I}),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \mathcal{C}_{-}(\widetilde{I})$ and let $(P, Q, n) \in B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$ be an element of depth $s>0$.

Let us first restate and extend somewhat the notations and the setting of 9.9.2. We set

$$
\begin{align*}
\left(P^{0}, Q^{0}, n^{0}\right) & :=(P, Q, n),  \tag{9.168}\\
\left(P_{\alpha}^{0}, Q_{\alpha}^{0}, n_{\alpha}^{0}\right) & :=\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right), \\
\left(P_{\omega}^{0}, Q_{\omega}^{0}, n_{\omega}^{0}\right) & :=\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) .
\end{align*}
$$

We have

- a strictly increasing sequence of parameter intervals

$$
\begin{equation*}
\widehat{I}_{0} \subset \widehat{I}_{1} \subset \cdots \subset \widehat{I}_{s-1} \subset \widehat{I}_{s}=I_{0} \tag{9.169}
\end{equation*}
$$

with $\widetilde{I} \subset \widehat{I}_{0}$; we denote by $\widetilde{I}_{r}$ the parent interval of $\widehat{I}_{r}$ for $0 \leqslant r<s$;

- a sequence $\left(P^{r}, Q^{r}, n^{r}\right), 0 \leqslant r \leqslant s$ such that $\left(P^{r}, Q^{r}, n^{r}\right)$ belongs to $\mathcal{R}\left(\widehat{I}_{r}\right)$ but not to $\mathcal{R}\left(\widetilde{I}_{r}\right)$ for $r<s$; also $\left(P^{s}, Q^{s}, n^{s}\right) \in \mathcal{R}\left(\widehat{I}_{s-1}\right) ;$
- two sequences $\left(P_{\alpha}^{r}, Q_{\alpha}^{r}, n_{\alpha}^{r}\right),\left(P_{\omega}^{r}, Q_{\omega}^{r}, n_{\omega}^{r}\right), 0 \leqslant r \leqslant s$; for each $r<s$, resp. $r=s$, the two elements belonging to $\mathcal{R}\left(\widehat{I}_{r}\right)$, resp. $\mathcal{R}\left(\widetilde{I}_{s-1}\right)$;
- two sequences $\left(P_{+}^{r}, Q_{+}^{r}, n_{+}^{r}\right),\left(P_{-}^{r}, Q_{-}^{r}, n_{-}^{r}\right), 0<r \leqslant s$; for each $r$, the two elements belonging to $\mathcal{R}\left(\widehat{I}_{r-1}\right)$.

These data are related by the following properties: for each $0<r \leqslant s$, we have

$$
\begin{align*}
& \left(P^{r-1}, Q^{r-1}, n^{r-1}\right) \in\left(P_{-}^{r}, Q_{-}^{r}, n_{-}^{r}\right) \square\left(P^{r}, Q^{r}, n^{r}\right) \square\left(P_{+}^{r}, Q_{+}^{r}, n_{+}^{r}\right),  \tag{9.170}\\
& \left(P_{\alpha}^{r-1}, Q_{\alpha}^{r-1}, n_{\alpha}^{r-1}\right) \in\left(P_{-}^{r}, Q_{-}^{r}, n_{-}^{r}\right) \square\left(P_{\alpha}^{r}, Q_{\alpha}^{r}, n_{\alpha}^{r}\right),  \tag{9.171}\\
& \left(P_{\omega}^{r-1}, Q_{\omega}^{r-1}, n_{\omega}^{r-1}\right) \in\left(P_{\omega}^{r}, Q_{\omega}^{r}, n_{\omega}^{r}\right) \square\left(P_{+}^{r}, Q_{+}^{r}, n_{+}^{r}\right) . \tag{9.172}
\end{align*}
$$

The process stops at step $s$ because of one of the two following cases occur
a) $\left(P^{s}, Q^{s}, n^{s}\right)$ does not belong to $\mathcal{R}\left(\widetilde{I}_{s-1}\right)$; then, by the structure theorem of Subsection 6.7 , there exists an integer $h>0$, elements $\left(P_{0}^{s}, Q_{0}^{s}, n_{0}^{s}\right) \cdots\left(P_{h}^{s}, Q_{h}^{s}, n_{h}^{s}\right)$ in $\mathcal{R}\left(\widetilde{I}_{s-1}\right)$ with

$$
\begin{equation*}
\left(P^{s}, Q^{s}, n^{s}\right) \in\left(P_{0}^{s}, Q_{0}^{s}, n_{0}^{s}\right) \square \cdots \square\left(P_{h}^{s}, Q_{h}^{s}, n_{h}^{s}\right) \tag{9.173}
\end{equation*}
$$

and also

$$
\begin{equation*}
P_{0}^{s} \subset P_{\alpha}^{s}, \quad Q_{h}^{s} \subset Q_{\omega}^{s} \tag{9.174}
\end{equation*}
$$

b) $\left(P^{s}, Q^{s}, n^{s}\right)$ belongs to $\mathcal{R}\left(I_{0}\right)$; in this case we set $h=0$.

We also observe that the parabolic compositions in (9.170) through (9.172) take place in $\mathcal{R}\left(\widehat{I}_{r-1}\right)$ but not in $\mathcal{R}\left(\widetilde{I}_{r-1}\right)$; in (9.173), they take place in $\mathcal{R}\left(\widehat{I}_{s-1}\right)$ but not in $\mathcal{R}\left(\widetilde{I}_{s-1}\right)$.

A subclass $B i_{+}\left(P_{\alpha}, Q_{\omega}, \ell\right)$, i.e. an element of $L$, distinct from the two $(\diamond, 0)$ that we already know is determined by the following data

- the depth $s(>0)$;
- the sequence $\widehat{I}_{0} \subset \cdots \subset \widehat{I}_{s}=I_{0}$;
- the integer $h \geqslant 0$;
- for each $0<i<h$, the smallest integer $u_{i}$ such that $\left|P_{i}^{s}\right| \geqslant 2^{-u_{i}}=: x_{i}$ for some $t \in \widetilde{I}$ when $h>1$;
- when $h>0$, the smallest integers $u_{0}, u_{h}$ such that $\left|P_{-}\right| \geqslant 2^{-u_{0}}=: x_{0}$ for some $t_{-} \in \widetilde{I}$, $\left|P_{+}\right| \geqslant 2^{-u_{h}}=: x_{h}$ for some $t_{+} \in \widetilde{I}$; here, the elements $\left(P_{-}, Q_{-}, n_{-}\right),\left(P_{+}, Q_{+}, n_{+}\right)$are determined by $P \subset P_{-}, Q \subset Q_{+}$and

$$
\begin{align*}
& \left(P_{-}, Q_{-}, n_{-}\right) \in\left(P_{-}^{1}, Q_{-}^{1}, n_{-}^{1}\right) \square \cdots \square\left(P_{-}^{s}, Q_{-}^{s}, n_{-}^{s}\right) \square\left(P_{0}^{s}, Q_{0}^{s}, n_{0}^{s}\right)  \tag{9.175}\\
& \left(P_{+}, Q_{+}, n_{+}\right) \in\left(P_{h}^{s}, Q_{h}^{s}, n_{h}^{s}\right) \square\left(P_{+}^{s}, Q_{+}^{s}, n_{+}^{s}\right) \square \cdots \square\left(P_{+}^{1}, Q_{+}^{1}, n_{+}^{1}\right) \tag{9.176}
\end{align*}
$$

Thus, we group together in a subclass $B i_{+}\left(P_{\alpha}, Q_{\omega}, \ell\right)$ the elements of $B i_{+}\left(P_{\alpha}, Q_{\omega}\right)$ who share the same data; the elements of $L$, distinct from $\diamond, 0$, are the sets of data for which at least one subclass $B i_{+}\left(P_{\alpha}, Q_{\omega}, \ell\right)$ is non-empty, for some $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ in $\mathcal{C}_{+}(\widetilde{I}),\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$ in $\mathcal{C}_{-}(\widetilde{I})$.

The definition of the set $L$ is now complete.

### 9.9.5 Sizes of subclasses of higher depth

The context and notations are the same as above. We want to define $b_{+}\left(P_{\alpha}, \ell\right)$ and $b_{-}\left(Q_{\omega}, \ell\right)$ in order to satisfy (9.131) in Subsection 9.8.

We first observe that $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ determines $\left(P_{-}^{1}, Q_{-}^{1}, n_{-}^{1}\right), \cdots,\left(P_{-}^{s}, Q_{-}^{s}, n_{-}^{s}\right),\left(P_{\alpha}^{s}, Q_{\alpha}^{s}, n_{\alpha}^{s}\right)$ and the result of parabolic compositions between these elements. Similarly, $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$ determines $\left(P_{+}^{1}, Q_{+}^{1}, n_{+}^{1}\right), \cdots,\left(P_{+}^{s}, Q_{+}^{s}, n_{+}^{s}\right),\left(P_{\omega}^{s}, Q_{\omega}^{s}, n_{\omega}^{s}\right)$ and the result of parabolic compositions between these elements. Therefore, the only "freedom" for the element $(P, Q, n)$ in the subclass $B i_{+}\left(P_{\alpha}, Q_{\omega}, \ell\right)$ is through $\left(P^{s}, Q^{s}, n^{s}\right)$, and this freedom is constrained by the relations $P^{s} \subset P_{\alpha}^{s}, Q^{s} \subset Q_{\omega}^{s}$.

Consider first a subclass with $h=0$, i.e., $\left(P^{s}, Q^{s}, n^{s}\right) \in \mathcal{R}\left(I_{0}\right)$. The widths of the strips are related as follows: for every $t \in \widetilde{I}$, we have

$$
\begin{equation*}
C^{-1} \frac{\left|P^{s}\right|}{\left|P_{\alpha}^{s}\right|\left|P_{\omega}^{s \mid}\right|} \leqslant \frac{|P|}{\left|P_{\alpha}\right|\left|P_{\omega}\right|} \leqslant C \frac{\left|P^{s}\right|}{\left|P_{\alpha}^{s}\right|\left|P_{\omega}^{s}\right|} \tag{9.177}
\end{equation*}
$$

This allows us to take, as in the case of depth 0 ,

$$
b_{+}\left(P_{\alpha}, \ell\right)= \begin{cases}\left(C|\widetilde{I}|^{1+\tau} x^{-1}\right)^{d_{s}^{*}} & \text { if }\left(P_{\alpha}^{s}, Q_{\alpha}^{s}, n_{\alpha}^{s}\right) \in \mathcal{R}\left(I_{0}\right)  \tag{9.178}\\ 0 & \text { otherwise }\end{cases}
$$

$$
b_{-}\left(Q_{\omega}, \ell\right)= \begin{cases}\left|P_{\omega}\right|^{d_{s}^{*}} & \text { if }\left(P_{\omega}^{s}, Q_{\omega}^{s}, n_{\omega}^{s}\right) \in \mathcal{R}\left(I_{0}\right), \\ 0 & \end{cases}
$$

Consider now a subclass with $h>0$, i.e., case a) in Subsection 9.9.4 above.

By the structure theorem in Subsection 6.7, see Lemma 11 in Subsection 9.6, for $0<i<h$, the element $\left(P_{i}^{s}, Q_{i}^{s}, n_{i}^{s}\right)$ belongs to $B i_{+}\left(\widetilde{I}_{s-1}, \widetilde{I}_{s-1}, \widetilde{I}_{s-1} ; x_{i}\right)$. From Corollary 6 in Subsection 6.7 , we have

$$
\begin{equation*}
x \leqslant C\left(C\left|\widehat{I}_{s-1}\right|^{-\frac{1}{2}}\right)^{h} x_{0} x_{1} \cdots x_{h} \tag{9.180}
\end{equation*}
$$

Thus, the data of every subclass must satisfy (9.180). Assuming that (9.180) holds, we set $b_{+}\left(P_{\alpha}, \ell\right)=0$ if $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \notin \mathcal{R}\left(\widehat{I}_{0}\right)$. When $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \mathcal{R}\left(\widehat{I}_{0}\right)$, we set

$$
\begin{equation*}
b_{+}\left(P_{\alpha}, \ell\right)=2^{h}\left(\prod_{0<i<h} \# B i_{+}\left(\widetilde{I}_{s-1}, \widetilde{I}_{s-1}, \widetilde{I}_{s-1} ; x_{i}\right)\right) \# B i_{+}\left(P_{\alpha}, \widetilde{I}_{s-1} ; x_{0}\right) \tag{9.181}
\end{equation*}
$$

Here, $B i_{+}\left(P_{\alpha}, \widetilde{I}_{s-1}, x_{0}\right)$ is by definition the set of elements $\left(P_{-}, Q_{-}, n_{-}\right)$in $\mathcal{R}(\widetilde{I})$ such that $P_{-} \subset P_{\alpha}$, $Q_{-}$is thin $\widetilde{I}_{s-1}$-critical and $\left|P_{-}\right| \geqslant x_{0}$ for some $t \in \widetilde{I}$.

Similarly, when (9.180) holds, we set $b_{-}\left(Q_{\omega}, \ell\right)=0$ if $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \notin \mathcal{R}\left(\widehat{I}_{0}\right)$. When $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in$ $\mathcal{R}\left(\widehat{I}_{0}\right)$, we set

$$
\begin{equation*}
b_{-}\left(Q_{\omega}, \ell\right)=\# B i_{+}\left(\widetilde{I}_{s-1}, Q_{\omega} ; x_{h}\right) \tag{9.182}
\end{equation*}
$$

where now $B i_{+}\left(\widetilde{I}_{s-1}, Q_{\omega} ; x_{h}\right)$ is the set of elements $\left(P_{+}, Q_{+}, n_{+}\right)$in $\mathcal{R}(\widetilde{I})$ such that $Q_{+} \subset Q_{\omega}, P_{+}$ is thin $\widetilde{I}_{s-1}$-critical and $\left|P_{+}\right| \geqslant x_{h}$ for some $t \in \widetilde{I}$.

The factor $2^{h}$ in (9.181) takes care of the possible results of the "free" parabolic compositions, i.e., those compositions which are not constrained by $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right)$ or $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)$.

The definition of $L, b_{+}, b_{-}$is now complete, and relation (9.131) is satisfied.

### 9.10 The Size of the Index set $L$

It is not difficult from (9.180) to see that the index set $L$ is finite, but we need an explicit bound on its cardinality (cf. (9.141)).

Proposition 33. The index set $L$ satisfies

$$
\# L \leqslant C|\widetilde{I}|^{-\tau^{2}}
$$

Proof. In the first part of the proof, we fix the depth $s$ and the sequence of intervals $\widehat{I}_{0} \subset \cdots \subset$ $\widehat{I}_{s-1} \subset \widehat{I}_{s}=I_{0}$. There is one subclass with $h=0$ and we will estimate the number of subclasses with $h>0$, i.e., the number of $(h+1)$-tuples $\left(u_{0}, \cdots, u_{h}\right)$ such that $(9.180)$ is satisfied; the integer $h$ itself is not fixed.

By Corollary 7 in Subsection 6.7, we have

$$
\begin{align*}
& x_{i}<\left|\widetilde{I}_{s-1}\right|^{\beta} \text { for } 0<i<h  \tag{9.183}\\
& x_{h}<C\left|\widetilde{I}_{s-1}\right|^{(1-\eta)^{-1}} \tag{9.184}
\end{align*}
$$

As $P_{-} \subset P_{\alpha}$, we also have, for a non-empty subclass

$$
\begin{equation*}
x_{0}<|\widetilde{I}|^{1+\tau} \tag{9.185}
\end{equation*}
$$

We rewrite (9.180) as

$$
\begin{equation*}
\frac{x_{0}}{|\widetilde{I}|^{1+\tau}}\left(\prod_{0<i<h} \frac{x_{i}}{\left|\widetilde{I}_{s-1}\right|^{\beta}}\right) \frac{x_{h}}{C\left|\widetilde{I}_{s-1}\right|^{\frac{1}{1-\eta}}} \geqslant \frac{x}{C|\widetilde{I}|^{1+\tau}\left|\widetilde{I}_{s-1}\right|^{\beta(h-1)+(1-\eta)^{-1}}}\left(C^{-1}\left|\widehat{I}_{s-1}\right|^{\frac{1}{2}}\right)^{h} \tag{9.186}
\end{equation*}
$$

Using $\beta>1$, and taking base-two logarithms, it is sufficient to bound the number of non-negative integral solutions of

$$
\begin{equation*}
n_{0}+\cdots+n_{h} \leqslant A_{0}-A_{1} h \tag{9.187}
\end{equation*}
$$

with

$$
\begin{align*}
A_{0} & =\log _{2}\left(|\widetilde{I}| x^{-1}\right)  \tag{9.188}\\
A_{1} & =\frac{1}{3} \log _{2}\left|\widehat{I}_{s-1}\right|^{-1} \tag{9.189}
\end{align*}
$$

As $x \leqslant \widetilde{x}_{c r} \ll|\widetilde{I}|$, both $A_{0}$ and $A_{1}$ are large; by taking $A_{0}$ slightly larger and $A_{1}$ slightly smaller, we can assume that both $A_{0}, A_{1}$ are integers. The number of non-negative integral solutions of (9.187) is then the coefficient of $z^{A_{0}}$ in the power series for

$$
\begin{equation*}
\chi(z):=\sum_{h \geqslant 0} z^{A_{1} h}(1-z)^{-h-2}=(1-z)^{-1}\left(1-z-z^{A_{1}}\right)^{-1} \tag{9.190}
\end{equation*}
$$

We estimate this coefficient by a Cauchy integral on the circle $|z|=1-2 A_{1}^{-1} \log A_{1}$. On this circle, we have

$$
\begin{align*}
\left|z^{A_{1}}\right| & <A_{1}^{-1}  \tag{9.191}\\
|\chi(z)| & <\frac{1}{2} A_{1}^{2}\left(\log A_{1}\right)^{-2} \tag{9.192}
\end{align*}
$$

The number of solutions of (9.187) is, therefore, not greater than

$$
\begin{equation*}
A_{1}^{2}\left(\log A_{1}\right)^{-2}\left(1-2 A_{1}^{-1} \log A_{1}\right)^{-A_{0}} . \tag{9.193}
\end{equation*}
$$

In view of (9.189), this quantity is smaller than

$$
\begin{equation*}
\left(\log \left|\widehat{I}_{s-1}\right|\right)^{2} \exp \left(C A_{0} \frac{\log |\log | \widehat{I}_{s-1}| |}{|\log | \widehat{I}_{s-1}| |}\right) \tag{9.194}
\end{equation*}
$$

This is a bound for the number of subclasses with fixed depth $s$ and sequence $\widehat{I}_{0} \subset \cdots \subset \widehat{I}_{s-1}$. We have now to sum over these remaining data. Observe that (9.194) depends on $\left|\widehat{I}_{s-1}\right|$, not on the depth $s$ and the intervals $\widehat{I}_{r}, 0 \leqslant r<s-1$.

Fix an interval $\widehat{I}$ with $\widetilde{I} \subset \widehat{I} \subset I_{0}, \widehat{I} \neq I_{0}$. Let $S(\widehat{I})$ be the number of parameter intervals $I^{*}$ with $\widetilde{I} \subset I^{*} \subset \widehat{I}, I^{*} \neq \widetilde{I}$. Every $I^{*}$ in this range may or may not be one of the $\widehat{I}_{r}$, for a sequence $\widehat{I}_{0} \subset \cdots \subset \widehat{I}_{s-1}$ terminating with $\widehat{I}_{s-1}=\widehat{I}$; in other terms, there are exactly $2^{S(\widehat{I})}$ such sequences (of various lengths). This means that the total number of subclasses is bounded by

$$
\begin{equation*}
\sum_{\widehat{I}} 2^{S(\hat{I})}(\log |\widehat{I}|)^{2} \exp \left(C A_{0} \frac{\log |\log | \widehat{I}| |}{|\log | \widehat{I}| |}\right) . \tag{9.195}
\end{equation*}
$$

We have here

$$
\begin{equation*}
\frac{\log |\log | \widehat{I}|\mid}{|\log | \widehat{I}|\mid} \leqslant \frac{\log \log \varepsilon_{0}^{-1}}{\log \varepsilon_{0}^{-1}} \tag{9.196}
\end{equation*}
$$

$$
\begin{equation*}
|\log | \widetilde{I}\left|=(1+\tau)^{S(\widehat{I})} \log \right| \widehat{I} \mid, \tag{9.197}
\end{equation*}
$$

$$
\begin{equation*}
S(\widehat{I}) \leqslant 2 \tau^{-1} \log \left(\frac{\log |\widetilde{I}|^{-1}}{\log \varepsilon_{0}^{-1}}\right)=: S_{\max } \tag{9.198}
\end{equation*}
$$

The sum (9.195) is thus bounded by

$$
\begin{align*}
& C 2^{S_{\max }}\left(\log \varepsilon_{0}^{-1}\right)^{2} \exp \left(C A_{0} \frac{\log \log \varepsilon_{0}^{-1}}{\log \varepsilon_{0}^{-1}}\right)  \tag{9.199}\\
& \leqslant\left(\log |\widetilde{I}|^{-1}\right)^{2 \tau^{-1}} \exp \left(C A_{0} \frac{\log \log \varepsilon_{0}^{-1}}{\log \varepsilon_{0}^{-1}}\right) .
\end{align*}
$$

As $x \geqslant x_{\text {min }}:=|\widetilde{I}|^{C\left(\rho_{0}-d_{s}^{*}\right)^{-1}}$ (cf. (9.64)), we have

$$
\begin{equation*}
A_{0} \leqslant C\left(\rho_{0}-d_{s}^{*}\right)^{-1} \log |\widetilde{I}|^{-1} . \tag{9.200}
\end{equation*}
$$

We choose the exponent $\rho_{0}$ in order to have

$$
\begin{equation*}
\rho_{0}>d_{s}^{*}+\tau . \tag{9.201}
\end{equation*}
$$

As $\varepsilon_{0}$ can be chosen arbitrarily small, we have

$$
\begin{equation*}
C A_{0} \frac{\log \log \varepsilon_{0}^{-1}}{\log \varepsilon_{0}^{-1}}<\frac{1}{2} \tau^{2} \log |\widetilde{I}|^{-1} \tag{9.202}
\end{equation*}
$$

We conclude, then, that with $\varepsilon_{0}$ small enough, the term in (9.199) is indeed smaller than $|\widetilde{I}|^{-\tau^{2}}$.

### 9.11 The Size of $\widehat{B}$

According to the roadmap exposed in Subsection 9.8, we have now to estimate the quantity set in Subsection 9.8

$$
\begin{equation*}
\widehat{B}=\sum_{L} b_{+}(\ell) b_{-}(\ell) \tag{9.140}
\end{equation*}
$$

with

$$
\begin{align*}
& b_{+}(\ell)=\sum_{\mathcal{C}_{+}(\widetilde{I})} b_{+}\left(P_{\alpha}, \ell\right),  \tag{9.136}\\
& b_{-}(\ell)=\sum_{\mathcal{C}_{-}(\widetilde{I})} b_{-}\left(Q_{\omega}, \ell\right) . \tag{9.137}
\end{align*}
$$

Consider first the bound elements. In view of (9.150), we have:

$$
\begin{align*}
b_{+}(\diamond) & =\# \mathcal{C}_{+}(\widetilde{I}) \log |\widetilde{I}|^{-1}  \tag{9.203}\\
b_{-}(\diamond) & =\# \mathcal{C}_{-}(\widetilde{I}) \log |\widetilde{I}|^{-1}
\end{align*}
$$

Consider next the class of depth 0 , and also the classes of higher depth with $h=0$ : in view of (9.166)-(9.167) and (9.178)-(9.179), we have in these cases

$$
\begin{align*}
b_{+}(\ell) & \leqslant\left(C|\widetilde{I}|^{1+\tau} x^{-1}\right)^{d_{s}^{*}} \# \mathcal{C}_{+}(\widetilde{I}),  \tag{9.204}\\
b_{-}(\ell) & \leqslant \sum_{\mathcal{C}_{-}(\widetilde{I})}\left(\max _{\widetilde{I}}\left|P_{\omega}\right|\right)^{d_{s}^{*}} \tag{9.205}
\end{align*}
$$

Also, the number of such classes, according to the discussion in the proof of Proposition 33 is not larger than

$$
\begin{equation*}
2^{S_{\max }} \leqslant\left(\frac{\log |\widetilde{I}|^{-1}}{\log \varepsilon_{0}^{-1}}\right)^{2 \tau^{-1}} \tag{9.206}
\end{equation*}
$$

The remaining subclasses are more complicated! Formulas (9.181), (9.182) suggest an induction. We thus assume that $(\mathrm{SR} 3)_{s}$ is satisfied for all parameter intervals containing $\widetilde{I}$. We have, for a class of depth $s>0$ with $h>0$ :

$$
\begin{gather*}
b_{+}(\ell) \leqslant 2^{h}\left(\prod_{0<i<h}\left(\# B i_{+}\left(\widetilde{I}_{s-1}, \widetilde{I}_{s-1}, \widetilde{I}_{s-1} ; x_{i}\right)\right)\right) \# B i_{+}\left(\widetilde{I}, \widetilde{I}, \widetilde{I}_{s-1} ; x_{0}\right)  \tag{9.207}\\
b_{-}(\ell) \leqslant \# B i_{+}\left(\widetilde{I}, \widetilde{I}_{s-1}, \widetilde{I} ; x_{h}\right) . \tag{9.208}
\end{gather*}
$$

Observe that, from (9.31), the critical value $x_{c r}$ in each of the $B i_{+}$sets above in the same and equal to

$$
\begin{equation*}
x_{c r}:=\varepsilon_{0}\left|P_{u}\right|\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\frac{\sigma_{0}}{\rho_{0}-\rho_{1}}} \tag{9.209}
\end{equation*}
$$

As in Subsection 9.6, we separate the subclasses into two parts: those for which every $x_{i}$ is above the critical value $x_{c r}$ and the others. In the first case, we have from (SR3) $s$

$$
\begin{equation*}
\# B i_{+}\left(\widetilde{I}_{s-1}, \widetilde{I}_{s-1}, \widetilde{I}_{s-1} ; x_{i}\right) \leqslant C\left(\frac{x_{i}}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}, \text { for } 0<i<h \tag{9.210}
\end{equation*}
$$

$$
\begin{gather*}
\# B i_{+}\left(\widetilde{I}, \widetilde{I}, \widetilde{I}_{s-1} ; x_{0}\right) \leqslant C\left(\frac{x_{0}}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}  \tag{9.211}\\
\# B i_{+}\left(\widetilde{I}, \widetilde{I}_{s-1}, \widetilde{I} ; x_{h}\right) \leqslant C\left(\frac{x_{h}}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}} . \tag{9.212}
\end{gather*}
$$

Multiplying these inequalities, we obtain, taking (9.180) into account

$$
\begin{equation*}
b_{+}(\ell) b_{-}(\ell) \leqslant A_{2}^{h} A_{3} \tag{9.213}
\end{equation*}
$$

with

$$
\begin{gather*}
A_{2}=2\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}\left(C \varepsilon_{0}\left|P_{u}\right|\left|\widehat{I}_{s-1}\right|^{-\frac{1}{2}}\right)^{\rho_{1}}  \tag{9.214}\\
A_{3}=C^{\rho_{1}}\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}} \tag{9.215}
\end{gather*}
$$

In the second case, as $\rho_{0}>\rho_{1}$, we have

$$
\begin{equation*}
\# B i_{+}\left(\widetilde{I}_{s-1}, \widetilde{I}_{s-1}, \widetilde{I}_{s-1} ; x_{i}\right) \leqslant C\left(\frac{x_{i}}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}} \tag{9.216}
\end{equation*}
$$

$$
\begin{equation*}
\# B i_{+}\left(\widetilde{I}, \widetilde{I}, \widetilde{I}_{s-1} ; x_{0}\right) \leqslant C\left(\frac{x_{0}}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}} \tag{9.217}
\end{equation*}
$$

$$
\begin{equation*}
\# B i_{+}\left(\widetilde{I}, \widetilde{I}_{s-1}, \widetilde{I} ; x_{h}\right) \leqslant C\left(\frac{x_{h}}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\sigma_{1}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\sigma_{0}} \tag{9.218}
\end{equation*}
$$

Moreover, comparing the $\sigma$ exponents in (9.19) and (9.20), we see that if $x_{i} \leqslant x_{c r}, 0 \leqslant i \leqslant h$, the corresponding inequality is still true after multiplying the right-hand side by $\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}}$. As this happens at least once, we get, by multiplying the three inequalities together:

$$
\begin{equation*}
b_{+}(\ell) b_{-}(\ell) \leqslant \widetilde{A}_{2}^{h} \widetilde{A}_{3} \tag{9.219}
\end{equation*}
$$

now with

$$
\begin{gather*}
\widetilde{A}_{2}=2\left(\frac{\left|\widetilde{I}_{s-1}\right|}{\varepsilon_{0}}\right)^{\sigma_{0}+\sigma_{1}}\left(C \varepsilon_{0}\left|P_{u}\right|\left|\widehat{I}_{s-1}\right|^{-\frac{1}{2}}\right)^{\rho_{0}}  \tag{9.220}\\
\widetilde{A}_{3}=C^{\rho_{0}}\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}}
\end{gather*}
$$

With $\widetilde{B}$ as in Subsection 9.8 , we have $\max \left(A_{3}, \widetilde{A}_{3}\right) \leqslant C \widetilde{B}$.
We observe that in both (9.213) and (9.219), our estimate for $b_{+}(\ell), b_{-}(\ell)$ depends on the class $\ell$ only through $\widehat{I}_{s-1}$ and $h$. We first sum over subclasses with a fixed depth $s$ and sequence $\widehat{I}_{0} \subset \cdots \subset \widehat{I}_{s-1}$, using the same method of generating series as in the proof of Proposition 33. To deal with the two cases at the same time, we first observe that

$$
\begin{equation*}
\sigma_{0}+\sigma_{1}-\frac{1}{2} \rho_{0}(1+\tau)=1-d_{u}^{0}-\frac{1}{2} d_{s}^{0}+o(1)>0 \tag{9.222}
\end{equation*}
$$

under (H4), and a fortiori $\sigma_{0}+\sigma_{1}-\frac{1}{2} \rho_{1}(1+\tau)>0$. Thus, $A_{2}$ and $\widetilde{A}_{2}$ are larger when $\widetilde{I}_{s-1}$ is larger; the largest case is $\widetilde{I}_{s-1}=I_{0}$, which gives

$$
\begin{equation*}
\max \left(A_{2}, \widetilde{A}_{2}\right) \leqslant \widehat{A}_{2}:=2\left(C \varepsilon_{0}^{\frac{1}{2}(1-\tau)}\left|P_{u}\right|\right)^{\rho_{1}} \tag{9.223}
\end{equation*}
$$

We, then, set

$$
\begin{align*}
\chi_{1}(z) & =\sum_{h>0} \widehat{A}_{2}^{h} z^{A_{1} h}(1-z)^{-h-2}  \tag{9.224}\\
& =\widehat{A}_{2} z^{A_{1}}(1-z)^{-2}\left(1-z-\widehat{A}_{2} z^{A_{1}}\right)^{-1}
\end{align*}
$$

The (partial) sum of $b_{+}(\ell) b_{-}(\ell)$ is, thus, not larger than $C \widetilde{B}$ times the coefficient of $z^{A_{0}}$ in the power series for $\chi_{1}(z)$. Recall that $A_{0}, A_{1}$ were defined in (9.188), (9.189).

We estimate this coefficient by Cauchy integration on the circle $\left\{|z|=1-A_{0}^{-1}-\widehat{A}_{2}\right\}$, on which we have

$$
\begin{align*}
& |1-z|^{-2} \leqslant\left(\widehat{A}_{2}+A_{0}^{-1}\right)^{-2} \leqslant A_{0}^{2}  \tag{9.225}\\
& \left|1-z-\widehat{A}_{2} z^{A_{1}}\right|^{-1} \leqslant A_{0}  \tag{9.226}\\
& \left|\chi_{1}(z)\right| \leqslant \widehat{A}_{2} A_{0}^{3}  \tag{9.227}\\
& \left|z^{-A_{0}}\right| \leqslant C\left(1+\widehat{A}_{2}\right)^{A_{0}} \tag{9.228}
\end{align*}
$$

The (partial) sum of $b_{+}(\ell) b_{-}(\ell)$ is therefore dominated by

$$
\begin{equation*}
C\left(1+\widehat{A}_{2}\right)^{A_{0}} \widehat{A}_{2} A_{0}^{3} \widetilde{B} \tag{9.229}
\end{equation*}
$$

We now have to sum over sequences $\widehat{I}_{0} \subset \cdots \subset \widehat{I}_{s-1}$ and depth $s$; but (9.229) is independent of these data and the same remarks as in the proof of Proposition 33 apply. So, we finally obtain for the sum of $b_{+}(\ell) b_{-}(\ell)$ over subclasses with $s>0$ and $h>0$, a bound by

$$
\begin{equation*}
C\left(1+\widehat{A}_{2}\right)^{A_{0}} \widehat{A}_{2} A_{0}^{3} 2^{S_{\max }} \widetilde{B} \tag{9.230}
\end{equation*}
$$

with $S_{\max }:=2 \tau^{-1} \log \left(\frac{\log \left(\left.\widetilde{I}\right|^{-1}\right.}{\log \varepsilon_{0}^{-1}}\right)$ (cf. (9.198).
We have here, from (9.200), (9.201)

$$
\begin{align*}
A_{0} & \leqslant C \tau^{-1} \log |\widetilde{I}|^{-1},  \tag{9.231}\\
\left(1+\widehat{A}_{2}\right) A_{0} & \leqslant|\widetilde{I}|^{-C \tau^{-1} \widehat{A}_{2}},  \tag{9.232}\\
\widehat{A}_{2} & \leqslant \varepsilon_{0}^{\frac{1}{2} \rho_{1}} . \tag{9.233}
\end{align*}
$$

As $\varepsilon_{0}$ can be made as small as we want with regard to $\tau$, the term in (9.230) is bounded by

$$
\begin{equation*}
\varepsilon_{0}^{\sigma}|\widetilde{I}|^{-\varepsilon_{o}^{\sigma}}\left(\log |\widetilde{I}|^{-1}\right)^{2 \tau^{-1}} \widetilde{B} \tag{9.234}
\end{equation*}
$$

for some fixed $\sigma>0$.
We summarize the calculations in this subsection in
Proposition 34. The quantity $\widehat{B}=\sum_{L} b_{+}(\ell) b_{-}(\ell)$ is bounded by $\widehat{B}_{1}+\widehat{B}_{2}+\widehat{B}_{3}$, with

$$
\begin{aligned}
\widehat{B}_{1} & =\left(\# \mathcal{C}_{+}(\widetilde{I})\right)\left(\# \mathcal{C}_{-}(\widetilde{I})\right)\left(\log |\widetilde{I}|^{-1}\right)^{2} \\
\widehat{B}_{2} & =\left(\frac{\log |\widetilde{I}|^{-1}}{\log \varepsilon_{0}^{-1}}\right)^{2 \tau^{-1}}\left(C|\widetilde{I}|^{1+\tau} x^{-1}\right)^{d_{s}^{*}}\left(\# \mathcal{C}_{+}(\widetilde{I})\right) \sum_{\mathcal{C}_{-}(\widetilde{I})}\left(\max _{\widetilde{I}}\left|P_{\omega}\right|\right)^{d_{s}^{*}}, \\
\widehat{B}_{3} & =\varepsilon_{0}^{\sigma}|\widetilde{I}|^{-\varepsilon_{0}^{\sigma}}\left(\log |\widetilde{I}|^{-1}\right)^{2 \tau^{-1}} \widetilde{B} .
\end{aligned}
$$

### 9.12 End of the Induction Step for (SR3)s

Of the two inequalities that were assumed in Subsection 9.8 to make work the argument, the first has been the subject of Proposition 33. The second is

$$
\begin{equation*}
\widehat{B}|\widetilde{I}|^{\tau\left(2-d_{s}^{+}-d_{u}^{+}-6 \tau\right)} \leqslant C B \tag{9.144}
\end{equation*}
$$

From Proposition 34, it is sufficient to prove the same inequality with $\widehat{B}$ replaced by $\widehat{B}_{i}, i=1,2,3$.
First consider $\widehat{B}_{3}$. For $x \leqslant \widetilde{x}_{c r}$, we have

$$
\begin{equation*}
\left.\widetilde{B}=\widetilde{B}_{0}=|\widetilde{I}|^{-\tau\left(2 \sigma_{0}+\sigma_{1}\right.}\right) B_{0} \leqslant|\widetilde{I}|^{-\tau\left(2 \sigma_{0}+\sigma_{1}\right)} B, \tag{9.235}
\end{equation*}
$$

and, therefore,

$$
\begin{equation*}
\widehat{B}_{3}|\widetilde{I}|^{\tau\left(2-d_{s}^{+}-d_{u}^{+}-6 \tau\right)} \leqslant \varepsilon_{0}^{\sigma}\left(\log |\widetilde{I}|^{-1}\right)^{2 \tau^{-1}}|\widetilde{I}|^{\omega} B \tag{9.236}
\end{equation*}
$$

with $\omega=\tau\left(2-d_{s}^{+}-d_{u}^{+}-6 \tau-2 \sigma_{0}-\sigma_{1}\right)-\varepsilon_{0}^{\sigma}$.

We choose the exponents $\sigma_{0}, \sigma_{1}$ in order to have

$$
\begin{equation*}
\omega>\tau^{2} \tag{9.237}
\end{equation*}
$$

which means

$$
\begin{equation*}
2 \sigma_{0}+\sigma_{1}<2-d_{s}^{+}-d_{u}^{+}-7 \tau-\varepsilon_{0}^{\sigma} \tau^{-1} \tag{9.238}
\end{equation*}
$$

which is compatible with the previous conditions on $\sigma_{0}, \sigma_{1}$.
As $\varepsilon_{0}$ can be chosen arbitrarily small with respect to $\tau,(9.144)$ holds for $\widehat{B}_{3}$.
Next consider $\widehat{B}_{1}$. We assume that the sizes of $\mathcal{C}_{+}(\widetilde{I}), \mathcal{C}_{-}(\widetilde{I})$ are controlled by $(\mathrm{SR} 1)_{s},(\mathrm{SR} 1)_{u}$. Then, we have

$$
\begin{equation*}
\widehat{B}_{1}|\widetilde{I}|^{\tau\left(2-d_{s}^{+}-d_{u}^{+}-6 \tau\right)} \leqslant C\left(\log |\widetilde{I}|^{-1}\right)^{2}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{2-2 d_{s}^{+}-2 d_{u}^{+}-2 \tau} \varepsilon_{0}^{-\tau\left(d_{s}^{0}+d_{u}^{0}\right)}|\widetilde{I}|^{\tau\left(2-d_{s}^{+}-d_{u}^{+}-6 \tau\right)} . \tag{9.239}
\end{equation*}
$$

The right-hand will be smaller than $C B_{0}$ as soon as

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{\rho_{0}} \leqslant\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}-2+2 d_{s}^{+}+2 d_{u}^{+}+A \tau} \varepsilon_{0}^{A \tau}\left(\log |\widetilde{I}|^{-1}\right)^{-2} \tag{9.240}
\end{equation*}
$$

for some fixed constant $A>0$.

We would like (9.240) to be a consequence of $x \leqslant \widetilde{x}_{c r}$, but unfortunately this is only true if $\widetilde{I}$ is not too large. Observe that

$$
\begin{equation*}
\rho_{0}^{-1}\left(2 \sigma_{0}+\sigma_{1}-2+2 d_{s}^{+}+2 d_{u}^{+}+A \tau\right)-\left(\rho_{0}-\rho_{1}\right)^{-1} \sigma_{0}=o(1) \tag{9.241}
\end{equation*}
$$

where $\left(\rho_{0}-\rho_{1}\right)^{-1} \sigma_{0}$ is the exponent appearing in the definition (9.31) of $\widetilde{x}_{c r}$. Write

$$
\begin{equation*}
\sigma:=2 \sigma_{0}+\sigma_{1}-2+2 d_{s}^{+}+2 d_{u}^{+}+A \tau=d_{s}^{+}+d_{u}^{+}+o(1) \tag{9.242}
\end{equation*}
$$

We choose the exponent $\rho_{1}$ in order to have

$$
\begin{equation*}
\rho_{1}>\rho_{0}\left(1-\frac{\sigma_{0}}{\sigma}\right)+\kappa, \tag{9.243}
\end{equation*}
$$

with $\kappa>0$ small to guarantee that (9.22) holds. Then, we have

$$
\begin{equation*}
\rho_{0}^{-1} \sigma<\frac{\sigma_{0}}{\rho_{0}-\rho_{1}+\kappa}<\frac{\sigma_{0}}{\left(\rho_{0}-\rho_{1}\right)}-C^{-1} \kappa \tag{9.244}
\end{equation*}
$$

and therefore (9.240) holds as soon as

$$
\begin{equation*}
x \leqslant \widetilde{x}_{c r}\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{-C^{-1} \kappa} \varepsilon_{0}^{A \tau}\left(\log |\widetilde{I}|^{-1}\right)^{-\frac{2}{\rho_{0}}} \tag{9.245}
\end{equation*}
$$

Keeping $\kappa \gg \tau$, the right-hand side is larger than $\widetilde{x}_{c r}$ when $|\widetilde{I}|<\varepsilon_{0}^{1+C \kappa^{-1} \tau}$, $C$ fixed large enough. Thus, we are able to conclude that $(9.144)$ holds for $\widehat{B}_{1}$ except in the range

$$
\begin{align*}
& \varepsilon_{0} \geqslant|\widetilde{I}| \geqslant \varepsilon_{0}^{1+C \kappa^{-1} \tau}  \tag{9.246}\\
& \widetilde{x}_{c r} \geqslant x \geqslant \widetilde{x}_{c r} \varepsilon_{0}^{2 A \tau}
\end{align*}
$$

We shall deal directly with this case below. Before that, we consider (9.144) for $\widehat{B}_{2}$. In this case, we assume that $(\mathrm{SR} 1)_{s}$ and $(\mathrm{SR} 2)_{u}$ hold. A small calculation shows that (9.144) holds as soon as

$$
\begin{equation*}
\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{\rho_{0}-d_{s}^{*}} \leqslant\left(\frac{|\widetilde{I}|}{\varepsilon_{0}}\right)^{\widehat{\sigma}} \varepsilon_{0}^{\widehat{A} \tau}\left(\frac{\log |\widetilde{I}|^{-1}}{\log \varepsilon_{0}^{-1}}\right)^{-2 \tau^{-1}} \tag{9.247}
\end{equation*}
$$

with

$$
\begin{equation*}
\widehat{\sigma}=-2+2 d_{s}^{+}+d_{u}^{+}-d_{s}^{*}+2 \sigma_{0}+\sigma_{1}+\widehat{A} \tau \tag{9.248}
\end{equation*}
$$

and $\widehat{A}$ a fixed positive constant. Here, both $\rho_{0}-d_{s}^{*}$ and $\widehat{\sigma}$ are $o(1)$.

With $\kappa$ as above, i.e., $\kappa=o(1), \tau=o(\kappa)$, we choose the exponents in order to have

$$
\begin{align*}
\rho_{0} & >d_{s}^{*}+\kappa^{2}  \tag{9.249}\\
\widehat{\sigma} & <-2 \kappa
\end{align*}
$$

which corresponds to

$$
\begin{equation*}
2 \sigma_{0}+\sigma_{1}<2-2 d_{s}^{+}-d_{u}^{+}+d_{s}^{*}-\widehat{A} \tau-2 \kappa \tag{9.251}
\end{equation*}
$$

If $(\widetilde{I}, x)$ is not in the range

$$
\begin{align*}
& \varepsilon_{0} \geqslant|\widetilde{I}| \geqslant \varepsilon_{0}^{1+C \kappa^{-1} \tau},  \tag{9.252}\\
& \widetilde{x}_{c r} \geqslant x \geqslant \widetilde{x}_{c r} \varepsilon_{0}^{C \kappa^{-1} \tau} \text {, }
\end{align*}
$$

then, (9.247) follows from $x \leqslant \widetilde{x}_{c r}$.

The final step in the inductive proof of $(\mathrm{SR} 3)_{s}$ is, therefore, the proof of
Proposition 35. Assume that

$$
|\widetilde{I}| \leqslant \varepsilon_{0}^{1+C \tau \kappa^{-1}}
$$

and that $(S R 1)_{\widehat{s}},(S R 1)_{\widehat{u}},(S R 2)_{\widehat{u}}$ hold for some candidate $I \subset \widetilde{I}$. Then $(S R 3)_{s}$ holds for $I$ in the range

$$
\widetilde{x}_{c r} \geqslant x \geqslant \widetilde{x}_{c r} \varepsilon_{0}^{C \tau \kappa^{-1}}
$$

Proof. Any element $(P, Q, n)$ in $B i_{+}(I, I, I ; x)$ satisfies $P \subset P_{s}, Q \subset Q_{u}$. Using (H1), this implies, for all $t \in I$ :

$$
\begin{equation*}
|P| \leqslant C \varepsilon_{0}\left|P_{u}\right| \tag{9.253}
\end{equation*}
$$

It, then, follows from Corollary 6 in Subsection 6.7 that for $x \geqslant \widetilde{x}_{c r} \varepsilon_{0}^{C \tau \kappa^{-1}}$, one must have $(P, Q, n) \in \mathcal{R}\left(I_{0}\right)$.

As $P$ is thin $I$-critical, there exists $\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) \in \widehat{\mathcal{C}}_{+}(I)$ with $P \subset P_{\alpha}$. Similarly, there exists $\left(P_{\omega}, Q_{\omega}, n_{\omega}\right) \in \widehat{\mathcal{C}}_{-}(I)$ with $Q \subset Q_{\omega}$; see Subsection 9.1 for the definitions of $\widehat{\mathcal{C}}_{-}(I), \widehat{\mathcal{C}}_{+}(I)$.

Let us estimate the number of possible $(P, Q, n)$ for fixed $P_{\alpha}, Q_{\omega}$.
If $n \leqslant n_{\alpha}+n_{\omega}$, there is at most one for each value of $n$. If $n \geqslant n_{\alpha}+n_{\omega}$, we can write (as $\left.(P, Q, n) \in \mathcal{R}\left(I_{0}\right)\right),(P, Q, n)$ as a simple composition

$$
(P, Q, n)=\left(P_{\alpha}, Q_{\alpha}, n_{\alpha}\right) *\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) *\left(P_{\omega}, Q_{\omega}, n_{\omega}\right)
$$

and conclude that there are no more than

$$
\begin{equation*}
C\left(\frac{x}{|I|\left|P_{\omega}\right|}\right)^{-d_{s}^{0}} \tag{9.254}
\end{equation*}
$$

possible $P^{\prime}$ (at this scale, the dependence of dimension on the parameter is not relevant).
The total number is, therefore, at most

$$
\begin{equation*}
C \log \left(\frac{\varepsilon_{0}\left|P_{u}\right|}{x}\right)\left(\# \widehat{\mathcal{C}}_{-}(I)\right)\left(\# \widehat{\mathcal{C}}_{+}(I)\right)+C\left(\# \widehat{\mathcal{C}}_{+}(I)\right)\left(\frac{x}{|I|}\right)^{-d_{s}^{*}} \sum_{\widehat{\mathcal{C}}_{-}(I)}\left|P_{\omega}\right|^{d_{s}^{*}} \tag{9.255}
\end{equation*}
$$

In view of $(\mathrm{SR} 1)_{\widehat{s}},(\mathrm{SR} 1)_{\widehat{u}},(\mathrm{SR} 2)_{\widehat{u}}$, this is not greater than

$$
\begin{equation*}
C\left(\log \frac{\varepsilon_{0}\left|P_{u}\right|}{x}\right)\left(\frac{|I|}{\varepsilon_{0}}\right)^{2-2 d_{s}^{+}-2 d_{u}^{+}-2 \tau}+C\left(\frac{|I|}{\varepsilon_{0}}\right)^{2-2 d_{s}^{+}-d_{u}^{+}-2 \tau+d_{s}^{*}}\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-d_{s}^{*}} \tag{9.256}
\end{equation*}
$$

and this should be smaller than $C B_{0}$ with

$$
\begin{equation*}
B_{0}=\left(\frac{x}{\varepsilon_{0}\left|P_{u}\right|}\right)^{-\rho_{0}}\left(\frac{|I|}{\varepsilon_{0}}\right)^{2 \sigma_{0}+\sigma_{1}} \tag{9.257}
\end{equation*}
$$

and

$$
\begin{equation*}
\frac{x}{\varepsilon_{0}\left|P_{u}\right|} \leqslant\left(\frac{|I|}{\varepsilon_{0}}\right)^{\frac{\sigma_{0}}{\rho_{0}-\rho_{1}}} \tag{9.258}
\end{equation*}
$$

The second part of (9.256) is as required as $\rho_{0}>d_{s}^{*}$ (cf. (9.65) and we ask that

$$
\begin{equation*}
2 \sigma_{0}+\sigma_{1}<2-2 d_{s}^{+}-d_{u}^{+}-2 \tau+d_{s}^{*} \tag{9.259}
\end{equation*}
$$

In the first part of (9.256), we bound the logarithmic term by a small negative power of $\frac{x}{\varepsilon_{0}\left|P_{u}\right|}$ and we use $(9.258)$ to conclude.

This ends the proof of $(\mathrm{SR} 3)_{s}$ in this case.

The induction step for $(\mathrm{SR} 3)_{s}$ is complete. We have proved
Theorem 2. Assume that all parameter intervals which contain $\widetilde{I}$ are strongly regular. Then, all candidates but a proportion not larger than $|\widetilde{I}|^{\tau^{2}}$ satisfy (SR3)s.

### 9.13 The Induction Step for (SR3) ${ }_{u}$

We have already explained for $(\mathrm{SR} 3)_{u}$ the cases where $I_{\alpha}$ or $I_{\omega}$ contains $\widetilde{I}$, and the case where $I=I_{\alpha}=I_{\omega}$ and $x$ is large (cf. Proposition 31).

The case $I=I_{\alpha}=I_{\omega}, x$ small which has been treated for (SR3) $)_{s}$ in the Subsections 9.8-9.13 is completely similar for $(\mathrm{SR} 3)_{u}$. It is only not completely symmetric because we have assumed that $d_{s}^{0} \geqslant d_{u}^{0}$ and thus the formulas for the exponents are not symmetric. So, one has only to be careful with the inequalities involving the exponents. For instance, in (9.222) we had

$$
\sigma_{0}+\sigma_{1}-\frac{1}{2} \rho_{0}(1+\tau)>0 .
$$

As $\rho_{0}^{\prime}=\frac{d_{u}^{0}}{d_{s}^{0}} \leqslant \rho_{0}$, we still have

$$
\sigma_{0}+\sigma_{1}-\frac{1}{2} \rho_{0}^{\prime}(1+\tau)>0 .
$$

Checking everything in this way is rather tedious, and we leave this to the reader.
Remark. At several points, we have asked that the exponents $\rho_{0}, \rho_{1}, \rho_{0}^{\prime}, \rho_{1}^{\prime}, \sigma_{0}, \sigma_{1}$ of (SR3) ${ }_{s}$, (SR3) ${ }_{u}$ should satisfy some inequalities; one could worry whether these inequalities are compatible between themselves (it is easy to check that each is compatible with (9.21) through (9.24) and (9.28) through (9.29). But we are always bounding $\rho_{0}, \rho_{1}, \rho_{0}^{\prime}, \rho_{1}^{\prime}$ from below and $\sigma_{0}, \sigma_{1}$ from above, hence the compatibility is obvious.

## 10 The Well-Behaved Part of the Dynamics for Strongly Regular Parameters

### 10.1 Prime Elements and Prime Decomposition

In the last two sections, we fix a strongly regular parameter, i.e. the intersection of a decreasing sequence $\left(I_{m}\right)_{m \geqslant 0}$ of strongly regular parameter intervals.

The sequence $\mathcal{R}\left(I_{m}\right)$ is increasing and we set

$$
\begin{equation*}
\mathcal{R}=\bigcup_{m \geqslant 0} \mathcal{R}\left(I_{m}\right) \tag{10.1}
\end{equation*}
$$

Definition. An element $(P, Q, n) \in \mathcal{R}$ is prime if $n>0$ and it cannot be written as a simple composition of two shorter elements.

Obviously, for any $\left(a, a^{\prime}\right) \in \mathcal{B}$, the element $\left(P_{a a^{\prime}}, Q_{a a^{\prime}}, 1\right)$ is prime. Such elements are called trivial primes. Non trivial primes are those of length bigger than 1.

There are only finitely many trivial primes. On the other hand, there are typically countably many non trivial ones.

Proposition 36. Any element $(P, Q, n) \in \mathcal{R}$ with $n>0$ can be uniquely written as a simple composition of a finite sequence of prime elements.

Proof. The existence of such a decomposition is clear. We have to show it is unique. Assume on the opposite that we can write

$$
\begin{align*}
(P, Q, n) & =\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{r}, Q_{r}, n_{r}\right)  \tag{10.2}\\
& =\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) * \cdots *\left(P_{s}^{\prime}, Q_{s}^{\prime}, n_{s}^{\prime}\right)
\end{align*}
$$

It is sufficient to show that $\left(P_{1}, Q_{1}, n_{1}\right)=\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)$. This is true if $n_{1}=n_{1}^{\prime}$. Assume for instance that $n_{1}<n_{1}^{\prime}$. Then we have $P \subset P_{1}^{\prime} \subset P_{1}$ with $P_{1}^{\prime} \neq P_{1}$. By Proposition 8 in Subsection 6.5 , we can write

$$
\begin{equation*}
\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)=\left(P_{1}, Q_{1}, n_{1}\right) *(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \tag{10.3}
\end{equation*}
$$

which contradicts the fact that $\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)$ is prime.
Remark. In the prime decomposition

$$
\begin{equation*}
(P, Q, n)=\left(P_{1}, Q_{1}, n_{1}\right) *\left(P_{r}, Q_{r}, n_{r}\right) \tag{10.4}
\end{equation*}
$$

$P_{1}$ can be characterized as the thinnest prime rectangle containing $P$.

We will denote by $\mathcal{P}$ the set of prime elements of $\mathcal{R}$. We denote by $\mathcal{R}^{*}$ the set of elements of $\mathcal{R}$ of length $>0$.

Let $(P, Q, n)$ be an element of $\mathcal{R}^{*}$ and let

$$
\begin{equation*}
(P, Q, n)=\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{r}, Q_{r}, n_{r}\right) \tag{10.5}
\end{equation*}
$$

be its prime decomposition. We define

$$
\begin{align*}
& T^{+}((P, Q, n))=\left(P_{2}, Q_{2}, n_{2}\right) * \cdots *\left(P_{r}, Q_{r}, n_{r}\right)  \tag{10.6}\\
& T^{-}((P, Q, n))=\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{r-1}, Q_{r-1}, n_{r-1}\right)
\end{align*}
$$

if $r>1$. When $(P, Q, n)$ is prime, with $P \subset R_{a}$ and $Q \subset R_{a^{\prime}}$, we set

$$
\begin{align*}
T^{+}((P, Q, n)) & =\left(R_{a^{\prime}}, R_{a^{\prime}}, 0\right)  \tag{10.7}\\
T^{-}((P, Q, n)) & =\left(R_{a}, R_{a}, 0\right)
\end{align*}
$$

For $S=(P, Q, n) \in \mathcal{R}$, we write $S * \mathcal{R}$, resp. $\mathcal{R} * S$, for the set of elements which can be written as $(P, Q, n) *\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$, resp. $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) *(P, Q, n)$, for some $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in \mathcal{R}$. We have partitions

$$
\begin{equation*}
\mathcal{R}^{*}=\bigsqcup_{\mathcal{P}} S * \mathcal{R}=\bigsqcup_{\mathcal{P}} \mathcal{R} * S . \tag{10.8}
\end{equation*}
$$

Moreover, for any $S \in \mathcal{P}$, the restriction of $T^{+}$, resp. $T^{-}$, to $S * \mathcal{R}$, resp. $\mathcal{R} * S$, is a bijection onto $\mathcal{R}$, inverse of $S^{\prime} \mapsto S * S^{\prime}$, resp. $S^{\prime} \mapsto S^{\prime} * S$.

### 10.2 Number of Factors in a Prime Decomposition

We write $r(S)$ for the number of factors in the prime decomposition of an element $S$ of $\mathcal{R}$ (setting $r(S)=0$ if $S$ has length 0$)$. Let $(P, Q, n),\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ be elements of $\mathcal{R}$ such that $P^{\prime}$ is a child of $P$. When $P^{\prime}$ is a simple child, it is obtained by simple composition of $P$ with an element of length 1 and we have

$$
\begin{equation*}
r\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)=r(P, Q, n)+1 \tag{10.9}
\end{equation*}
$$

Proposition 37. If $P^{\prime}$ is a non-simple child of $P$, we have

$$
r\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \leqslant r(P, Q, n)
$$

Proof. Let

$$
\begin{equation*}
(P, Q, n)=\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{r}, Q_{r}, n_{r}\right) \tag{10.10}
\end{equation*}
$$

be the prime decomposition of $(P, Q, n)$.

Let $(\widehat{P}, \widehat{Q}, \widehat{n})$ be the element of $R$ such that

$$
\begin{equation*}
\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in(P, Q, n) \square(\widehat{P}, \widehat{Q}, \widehat{n}), \tag{10.11}
\end{equation*}
$$

(cf. Proposition 5 in Subsection 6.4). There exists $m \geqslant 0$ such that $Q$ and $\widehat{P}$ are $I_{m}$-transverse. Define

$$
\begin{align*}
\left(P^{i}, Q^{i}, n^{i}\right) & =\left(P_{i}, Q_{i}, n_{i}\right) * \cdots *\left(P_{r}, Q_{r}, n_{r}\right)  \tag{10.12}\\
& =\left(T^{+}\right)^{i-1}(P, Q, n), 1 \leqslant i \leqslant r .
\end{align*}
$$

We have an increasing sequence

$$
\begin{equation*}
Q=Q^{1} \subset Q^{2} \subset \cdots \subset Q^{r}=Q_{r} . \tag{10.13}
\end{equation*}
$$

Let $r^{\prime}$ be the largest integer in $\{1, \cdots, r\}$ such that $Q^{r^{\prime}}$ and $\widehat{P}$ are $I_{m}$-transverse for some $m \geqslant 0$ (and then for all large enough $m$ ). Define $(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}$ by the condition $Q^{\prime} \subset \widetilde{Q}$ and

$$
\begin{equation*}
(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in\left(P^{r^{\prime}}, Q^{r^{\prime}}, n^{r^{\prime}}\right) \square(\widehat{P}, \widehat{Q}, \widehat{n}) . \tag{10.14}
\end{equation*}
$$

We then have

$$
\begin{equation*}
\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)=\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{r^{\prime}-1}, Q_{r^{\prime}-1}, n_{r^{\prime}-1}\right) *(\widetilde{P}, \widetilde{Q}, \widetilde{n}) . \tag{10.15}
\end{equation*}
$$

The assertion of the proposition is, thus, a consequence of
Lemma 12. ( $\widetilde{P}, \widetilde{Q}, \widetilde{n})$ is prime.

Proof. Assume by contradiction that we can write

$$
\begin{equation*}
(\widetilde{P}, \widetilde{Q}, \widetilde{n})=\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) *\left(\widetilde{P}_{2}, \widetilde{Q}_{2}, \widetilde{n}_{2}\right) \tag{10.16}
\end{equation*}
$$

with $\widetilde{n}_{1}, \widetilde{n}_{2}>0$. Define

$$
\begin{equation*}
\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)=\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{r^{\prime}-1}, Q_{r^{\prime}-1}, n_{r^{\prime}-1}\right) *\left(\widetilde{P}_{1}, \widetilde{Q}_{1}, \widetilde{n}_{1}\right) . \tag{10.17}
\end{equation*}
$$

We have $P^{\prime} \subset P_{1}^{\prime}, P^{\prime} \neq P_{1}^{\prime}$, hence $P \subset P_{1}^{\prime}$. We also have

$$
\begin{equation*}
\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)=\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) *\left(\widetilde{P}_{2}, \widetilde{Q}_{2}, \widetilde{n}_{2}\right) . \tag{10.18}
\end{equation*}
$$

By Proposition 8 in Subsection 6.5, there exists $\left(P^{*}, Q^{*}, n^{*}\right)$ such that

$$
\begin{equation*}
(P, Q, n)=\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) *\left(P^{*}, Q^{*}, n^{*}\right) \tag{10.19}
\end{equation*}
$$

We claim that there exists $j \in\{1, \cdots, r\}$ such that

$$
\begin{equation*}
\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)=\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{j}, Q_{j}, n_{j}\right) . \tag{10.20}
\end{equation*}
$$

Indeed, if $j$ is the smallest integer such that

$$
\begin{equation*}
n_{1}^{\prime} \leqslant n_{1}+\cdots+n_{j} \tag{10.21}
\end{equation*}
$$

it follows from Proposition 8 that we can write

$$
\begin{equation*}
\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{j}, Q_{j}, n_{j}\right)=\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right) *(\bar{P}, \bar{Q}, \bar{n}) \tag{10.22}
\end{equation*}
$$

for some $(\bar{P}, \bar{Q}, \bar{n}) \in \mathcal{R}$. By the same proposition, we can also write

$$
\begin{equation*}
\left(P_{j}, Q_{j}, n_{j}\right)=\left(\bar{P}^{\prime}, \bar{Q}^{\prime}, \bar{n}^{\prime}\right) *(\bar{P}, \bar{Q}, \bar{n}) \tag{10.23}
\end{equation*}
$$

for some $\left(\bar{P}^{\prime}, \bar{Q}^{\prime}, \bar{n}^{\prime}\right) \in \mathcal{R}$. We have $\bar{n}<n_{j}$ by the definition of $j$, hence $\bar{n}^{\prime}>0$. As $\left(P_{j}, Q_{j}, n_{j}\right)$ is prime, we have $\bar{n}=0$ which implies our claim.

The integer $j$ satisfies $r^{\prime} \leqslant j \leqslant r$. We have

$$
\begin{equation*}
\left(\widetilde{P}_{2}, \widetilde{Q}_{2}, \widetilde{n}_{2}\right) \in\left(P^{j+1}, Q^{j+1}, n^{j+1}\right) \square(\widehat{P}, \widehat{Q}, \widehat{n}) \tag{10.24}
\end{equation*}
$$

which contradicts the definition of $r^{\prime}$.

This concludes the proof of the lemma and also of the proposition.

### 10.3 A Weighted Estimate on the Number of Children

We present in this subsection a variation over the estimates in Subsection 8.1, which will be important in the definition of a transfer operator.

We fix a constant $\kappa \in(0,1)$ close to 1 , but independent of $\varepsilon_{0}$. We set

$$
\begin{equation*}
d_{s}^{-}=d_{s}^{0}-C \varepsilon_{0} \tag{10.25}
\end{equation*}
$$

with a constant $C$ sufficiently large so $d_{s}^{-}$is smaller than the transverse Hausdorff dimension of the stable foliation $W^{s}(K)$ for the parameter that we are dealing with.

For $S=(P, Q, n)$, we set

$$
\begin{equation*}
\|P\|=|P|^{d_{s}^{-}} \kappa^{r(S)} \tag{10.26}
\end{equation*}
$$

(we will also write $r(P)$ instead of $r(S)$ ).
Proposition 38. For any $m \geqslant 1$, any $(P, Q, n) \in \mathcal{R}$, we have

$$
\sum_{P^{\prime}}\left\|P^{\prime}\right\| \leqslant C \kappa^{\frac{m}{2}}\|P\|
$$

where the sum in the left-hand side is over elements $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ such that $P^{\prime}$ is a descendent of the $m^{t h}$ generation of $P$.

We will first state a Lemma, then prove the proposition from the Lemma, and finally prove the Lemma.

Lemma 13. Let $\varepsilon_{1}>0$. If $\varepsilon_{0}$ is small enough, we have

$$
\sum_{P^{\prime}}\left\|P^{\prime}\right\| \leqslant \varepsilon_{1}\|P\|
$$

for all $(P, Q, n) \in \mathcal{R}$, where the sum in the left-hand side is over non-simple children of $P$.

Proof of the Proposition.

Let $m_{0} \geqslant 1$ be an integer to be determined later. Consider all chains

$$
\begin{equation*}
P=P^{0} \supset P^{1} \supset \cdots \supset P^{m_{0}}=P^{\prime} \tag{10.27}
\end{equation*}
$$

where $P$ is given and $P^{i+1}$ is a child of $P^{i}$. If $P^{i+1}$ is, for each $i$, a simple child of $P^{i}$, one has $r\left(P^{\prime}\right)=m_{0}+r(P)$ and the corresponding part of the sum in Proposition 38 satisfies

$$
\begin{equation*}
\sum\left\|P^{\prime}\right\| \leqslant C \kappa^{m_{0}}\|P\| \tag{10.28}
\end{equation*}
$$

(as long as $m_{0}=o\left(\varepsilon_{0}^{-1}\right)$ ). We choose $m_{0}$ such that in (10.28) we have From the lemma above, it follows that for every $\widetilde{P}$, we have

$$
\begin{gather*}
2 C \kappa^{m_{0}} \leqslant \kappa^{\frac{m_{0}}{2}}  \tag{10.29}\\
\sum\left\|\widetilde{P}^{\prime}\right\| \leqslant C\|\widetilde{P}\| \tag{10.30}
\end{gather*}
$$

where the sum is over all children of $\widetilde{P}$. Using the lemma again, when we sum over chains such that $P_{i+1}$ is a non-simple child of $P_{i}$ for some $i$, we obtain

$$
\begin{equation*}
\sum\left\|\widetilde{P}^{\prime}\right\| \leqslant m_{0} C^{m_{0}-1} \varepsilon_{1}\|P\| . \tag{10.31}
\end{equation*}
$$

Taking $\varepsilon_{1}$ small enough, we obtain

$$
\begin{equation*}
\sum\left\|P^{\prime}\right\| \leqslant \kappa^{\frac{m_{0}}{2}}\|P\| \tag{10.32}
\end{equation*}
$$

where the sum is now over all chains. The proposition follows immediately from (10.32) and (10.30).

Proof of Lemma 13. Let $(P, Q, n) \in \mathcal{R}$. Any non-simple child $P^{\prime}$ of $P$ is obtained as

$$
\begin{equation*}
\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in(P, Q, n) \square\left(P_{1}, Q_{1}, n_{1}\right) \tag{10.33}
\end{equation*}
$$

and we denote by $\widetilde{P}_{1}$, the parent of $P_{1}$. One has

$$
\begin{equation*}
\left|P^{\prime}\right| \leqslant C|P|\left|P_{1}\right| \delta\left(Q, P_{1}\right)^{-\frac{1}{2}} \tag{10.34}
\end{equation*}
$$

Therefore, we will have

$$
\begin{equation*}
\|P\|^{-1} \sum\left\|P^{\prime}\right\| \leqslant C \sum\left|P_{1}\right|^{d_{s}^{-}} \kappa^{r\left(P^{\prime}\right)-r(P)} \delta^{-\frac{1}{2} d_{s}^{-}} \tag{10.35}
\end{equation*}
$$

From the proof of Proposition 37, there is an increasing sequence

$$
\begin{equation*}
Q=Q^{1} \subset Q^{2} \subset \cdots \subset Q^{r(P)} \tag{10.36}
\end{equation*}
$$

such that $r\left(P^{\prime}\right)$ is the largest integer for which $Q^{r}$ and $P_{1}$ are $I_{m}$-transverse for large enough $m$. If

$$
\begin{equation*}
\left|Q^{r}\right| \leqslant \delta\left(Q, P_{1}\right)^{2} \tag{10.37}
\end{equation*}
$$

then, by Proposition $10, Q^{r}$ and $P_{1}$ are $I_{m}$-transverse for large $m$ and thus $r\left(P^{\prime}\right) \geqslant r$. On the other hand, there exists $\kappa^{*} \in(0,1)$ such that

$$
\begin{equation*}
\left|Q^{r}\right| \leqslant \kappa^{*}\left|Q^{r+1}\right| \tag{10.38}
\end{equation*}
$$

for $r<r(P)$. We infer that

$$
\begin{equation*}
r(P)-r\left(P^{\prime}\right) \leqslant C \log \left(\delta\left(Q, P_{1}\right)\right)^{-1} \tag{10.39}
\end{equation*}
$$

Therefore, if $\kappa$ is close enough to 1 , we have

$$
\begin{equation*}
\kappa^{r\left(P^{\prime}\right)-r(P)} \leqslant\left(\delta\left(Q, P_{1}\right)\right)^{-\frac{1}{6} d_{s}^{-}} \tag{10.40}
\end{equation*}
$$

and the right-hand side of (10.35) is bounded by

$$
\begin{equation*}
C \sum\left|P_{1}\right|^{d_{s}^{-}}\left(\delta\left(Q, P_{1}\right)\right)^{-\frac{2}{3} d_{s}^{-}} \tag{10.41}
\end{equation*}
$$

Using (R7), this is smaller than

$$
\begin{equation*}
C \sum\left|P_{1}\right|^{\frac{1}{3} d_{s}^{-}} . \tag{10.42}
\end{equation*}
$$

To estimate this sum, we first fix the parent $\widetilde{P}_{1}$ and sum over children $P_{1}$; it follows from Proposition 21 that the corresponding sum is bounded by $C\left|\widetilde{P}_{1}\right|^{\frac{1}{3} d_{s}^{-}}$and (10.42) is not greater than

$$
\begin{equation*}
C \sum\left|\widetilde{P}_{1}\right|^{\frac{1}{3} d_{s}^{-}} \tag{10.43}
\end{equation*}
$$

For each integer $m$, let us count now how many $\widetilde{P}_{1}$ may satisfy

$$
\begin{equation*}
2^{-m}\left|P_{s}\right| \geqslant\left|\widetilde{P}_{1}\right| \geqslant 2^{-m-1}\left|P_{s}\right| . \tag{10.44}
\end{equation*}
$$

As $Q$ is transverse to $P_{1}$ but not to $\widetilde{P}_{1}$, we must have, by Proposition 10:

$$
\begin{equation*}
\delta\left(Q, \widetilde{P}_{1}\right) \leqslant C 2^{-m(1-\eta)} \tag{10.45}
\end{equation*}
$$

which shows that there are no more than $C 2^{m \eta}$ such $\widetilde{P}_{1}$ 's. This implies that the sum (10.43) is at most of order $C \varepsilon_{0}^{\frac{1}{3} d_{s}^{-}}$, which yields the statement of the lemma.

## Remark.

1. In the lemma, the value $d_{s}^{-}=d_{s}^{0}-C \varepsilon_{0}$ that has been used to define $\|P\|$ is irrelevant. The assertion of the lemma is still true if we replace $d_{s}^{-}$by any positive number bounded away from 0 .
2. In the proposition, we can replace $d_{s}^{-}$by a slightly lower value (assuming, as usual, that $\varepsilon_{0}$ is arbitrarily small): if we take

$$
\begin{equation*}
\widehat{d}_{s}^{-}=d_{s}^{0}-o\left(\log \kappa^{-1}\right) \tag{10.46}
\end{equation*}
$$

the same argument works and the result of the proposition is still valid.
Corollary 14. Let $\varepsilon_{1}>0$. If $\varepsilon_{0}$ is small enough, we have

$$
\sum n|P|^{d_{s}^{-}}<\varepsilon_{1}
$$

where the sum in the left-hand side is over non-trivial primes $(P, Q, n)$.

Proof. Let $(P, Q, n)$ be a non trivial prime. We have, by Proposition 13 of Subsection 7.1

$$
\begin{equation*}
n \leqslant\left(\log \left(C|P|^{-1}\right)\right)^{\frac{\log 2}{\log 3 / 2}} \tag{10.47}
\end{equation*}
$$

Consequently, we have

$$
\begin{equation*}
n|P|^{d_{s}^{-}} \leqslant|P|^{\widehat{d_{s}^{-}}} \tag{10.48}
\end{equation*}
$$

with $\widehat{d}_{s}^{-}$as in (10.46) and $d_{s}^{0}-\widehat{d}_{s}^{-}$being independent of $\varepsilon_{0}$ (as $P$ is a non trivial prime, one has $\left.|P|<\varepsilon_{0}\right)$. Observe also that the thinnest $(\widetilde{P}, \widetilde{Q}, \widetilde{n}) \in \mathcal{R}\left(I_{0}\right)$ with $P \subset \widetilde{P}$ satisfies $\widetilde{Q} \subset Q_{u}$ hence $|\widetilde{P}| \leqslant \varepsilon_{0}^{\alpha}$ for some fixed positive $\alpha$.

We apply the proposition, taking the Remark 2 above into account and using $\widehat{d}_{s}^{-}$instead of $d_{s}^{-}$; we obtain

$$
\begin{align*}
\sum n|P|^{d_{s}^{-}} & \leqslant \sum|P|^{\widehat{d}_{s}^{-}}  \tag{10.49}\\
& =\kappa^{-1} \sum\|P\| \\
& \leqslant C \kappa^{-1}\left(1-\kappa^{\frac{1}{2}}\right)^{-1} \sum\|\widetilde{P}\|
\end{align*}
$$

where, in the last sum, $(\widetilde{P}, \widetilde{Q}, \widetilde{n})$ runs through the elements of $\mathcal{R}\left(I_{0}\right)$ with $|\widetilde{P}|$ of the order of $\varepsilon_{0}^{\alpha}$. We have

$$
\begin{align*}
\sum\|\widetilde{P}\| & \leqslant C \sum \varepsilon_{0}^{\alpha \widehat{d}_{s}^{-}} \kappa^{C^{-1} \log \varepsilon_{0}^{-1}}  \tag{10.50}\\
& \leqslant C \varepsilon_{0}^{\alpha\left(d_{s}^{-}-d_{s}^{0}\right)+C^{-1} \log \kappa^{-1}}
\end{align*}
$$

and the exponent is positive from (10.46). Putting this into (10.49) yields Corollary 14.

Remark. We do not know the range of $d$ for which

$$
\sum_{\mathcal{P}}|P|^{d}
$$

is convergent. The Corollary shows at least that there are relatively few primes: the sum

$$
\sum_{\mathcal{R}\left(I_{0}\right)}|P|^{d_{s}^{-}} \quad \text { is convergent. }
$$

### 10.4 Stable Curves

Let $\left(P_{k}, Q_{k}, n_{k}\right)_{k \geqslant 0}$ be a sequence of elements of $\mathcal{R}$ such that $P_{k+1}$ is strictly contained in $P_{k}$ for $k \geqslant 0$. Let $R_{a}$ be the rectangle of the Markov partition which contains $P_{0}$. The vertical part of the boundary of $P_{k}$ is the union of two graphs $\left\{x_{a}=\varphi_{k}^{ \pm}\left(y_{a}\right)\right\}$, and the $\varphi_{k}^{ \pm}$are uniformly bounded in the $C^{2}$ topology. Moreover, there exists $\kappa^{*} \in(0,1)$ such that $\left|P_{k+1}\right| \leqslant \kappa^{*}\left|P_{k}\right|$ for all $k \geqslant 0$. It follows that both sequences $\varphi_{k}^{ \pm}$converge in the $C^{1}$ topology to the same limit $\varphi_{\infty}$, which is of class $C^{1+\text { Lip }}$, where Lip stands for Lipschitz. We state this as

Proposition 39. The intersection $\bigcap_{k \geqslant 0} P_{k}$ is the graph $\left\{x_{a}=\varphi_{\infty}\left(y_{a}\right)\right\}$ of a $C^{1+\text { Lip }}$ function. Moreover, the $C^{1+\text { Lip }}$ norm of $\varphi_{\infty}$ is bounded independently of the sequence $\left(P_{k}\right)_{k \geqslant 0}$.

## Definition.

1. A stable curve is the intersection $\omega=\bigcap_{k \geqslant 0} P_{k}$ of a decreasing sequence of vertical-like rectangles as above. An unstable curve is the intersection $\omega^{\prime}=\bigcap_{k \geqslant 0} Q_{k}^{\prime}$ of a decreasing sequence of horizontal-like strips.
2. The set of stable curves, resp. unstable curves, is denoted by $\mathcal{R}_{+}^{\infty}$, resp. $\mathcal{R}_{-}^{\infty}$. The union of stable curves, resp. unstable curves, is denoted by $\widetilde{\mathcal{R}}_{+}^{\infty}$, resp. $\widetilde{\mathcal{R}}_{-}^{\infty}$.
3. Any stable curve $\omega \subset R_{a}$ has a canonical defining sequence characterized by the following conditions: $P_{0}=R_{a}$ and, for each $k, P_{k+1}$ is a child of $P_{k}$.
4. Two stable curves are equal or disjoint. Hence there is a canonical projection

$$
\pi: \widetilde{\mathcal{R}}_{+}^{\infty} \mapsto \mathcal{R}_{+}^{\infty}
$$

We will now define dynamics on a part of the sets $\mathcal{R}_{+}^{\infty}, \widetilde{\mathcal{R}}_{+}^{\infty}$.
Let $\mathcal{N}_{+}$be the set of stable curves $\omega$ which are contained in infinitely many prime elements and let $\mathcal{D}_{+}$be the complementary subset in $\mathcal{R}_{+}^{\infty}$. For $(P, Q, n) \in \mathcal{P}$, denote by $\mathcal{R}_{+}^{\infty}(P)$ the set of stable curves $\omega \in \mathcal{D}_{+}$such that $P$ is the thinnest prime containing $\omega$.

We, thus, have partitions

$$
\begin{align*}
\mathcal{R}_{+}^{\infty} & =\mathcal{N}_{+} \bigsqcup \mathcal{D}_{+},  \tag{10.51}\\
\mathcal{D}_{+} & =\bigsqcup_{\mathcal{P}} \mathcal{R}_{+}^{\infty}(P) . \tag{10.52}
\end{align*}
$$

We denote by $\widetilde{\mathcal{N}}_{+}, \widetilde{\mathcal{D}}_{+}, \widetilde{\mathcal{R}}_{+}^{\infty}(P)$ the respective pre-images by $\pi$.
Let $(P, Q, n) \in \mathcal{P}, \omega \in \mathcal{R}_{+}^{\infty}(P)$. For any $\left(P_{k}, Q_{k}, n_{k}\right)$ with $\omega \subset P_{k} \subset P$, we can write (cf. Remark after Proposition 36)

$$
\begin{equation*}
\left(P_{k}, Q_{k}, n_{k}\right)=(P, Q, n) *\left(P_{k}^{\prime}, Q_{k}^{\prime}, n_{k}^{\prime}\right) \tag{10.53}
\end{equation*}
$$

for some $\left(P_{k}^{\prime}, Q_{k}^{\prime}, n_{k}^{\prime}\right) \in \mathcal{R}$; we have

$$
\begin{equation*}
T^{+}\left(P_{k}, Q_{k}, n_{k}\right)=\left(P_{k}^{\prime}, Q_{k}^{\prime}, n_{k}^{\prime}\right) \tag{10.54}
\end{equation*}
$$

and we define $\omega^{\prime}=T^{+}(\omega)$ to be the stable curve obtained by the intersection of the $P_{k}^{\prime}$ when $P_{k}$ decrease to $\omega$. We have

$$
\begin{equation*}
g^{n}\left(P_{k}\right) \subset P_{k}^{\prime}, \quad g^{n}(\omega) \subset \omega^{\prime} \tag{10.55}
\end{equation*}
$$

and we also define

$$
\begin{equation*}
\widetilde{T}^{+} / \widetilde{\mathcal{R}}_{+}^{\infty}(P)=g^{n} / \widetilde{\mathcal{R}}_{+}^{\infty}(P) . \tag{10.56}
\end{equation*}
$$

We, thus, have a commutative diagram

$$
\begin{array}{ccc}
\widetilde{\mathcal{D}}_{+} \xrightarrow{\widetilde{T}^{+}} & \widetilde{\mathcal{R}}_{+}^{\infty} \\
\downarrow_{\pi} & & \downarrow^{\pi}  \tag{10.57}\\
\mathcal{D}_{+} \xrightarrow{T^{+}} & \mathcal{R}_{+}^{\infty}
\end{array}
$$

We observe that for $(P, Q, n) \in \mathcal{P}$ with $Q \subset R_{a}$, the image $T^{+}(\omega)$ of any $\omega \in \mathcal{R}_{+}^{\infty}(P)$ is contained in $R_{a}$.

Conversely, let $(P, Q, n) \in \mathcal{P}$ with $Q \subset R_{a}$ and let $\omega^{\prime} \in \mathcal{R}_{+}^{\infty}, \omega^{\prime} \subset R_{a}$. For any $\left(P_{k}^{\prime}, Q_{k}^{\prime}, n_{k}^{\prime}\right)$ with $\omega \subset P_{k}^{\prime}$, we define ( $P_{k}, Q_{k}, n_{k}$ ) by (10.53); the intersection $\omega$ of the $P_{k}^{\prime} \mathrm{s}$, when $P_{k}^{\prime}$ decrease to $\omega^{\prime}$, is the unique stable curve in $\mathcal{R}_{+}^{\infty}(P)$ such that $T^{+}(\omega)=\omega^{\prime}$.

Thus, $T^{+}$induces a bijection from $\mathcal{R}_{+}^{\infty}(P)$ on the set $\mathcal{R}_{+}^{\infty}(a)$ of stable curves contained in $R_{a}$. For $\omega \in \mathcal{R}_{+}^{\infty}(P)$, we have

$$
\begin{equation*}
\widetilde{T}^{+}(\omega)=\omega^{\prime} \cap Q . \tag{10.58}
\end{equation*}
$$

### 10.5 Topology and Geometry of $\mathcal{R}_{+}^{\infty}$ and $\widetilde{\mathcal{R}}_{+}^{\infty}$

Each stable curve is a compact subset of $R=\cup R_{a}$. Therefore, $\mathcal{R}_{+}^{\infty}$ may be viewed as a subset of the set of non empty compact subsets of $R$ endowed with the Hausdorff topology. The topology induced on $\mathcal{R}_{+}^{\infty}$ can also be viewed directly: for any $\omega=\cap P_{k}$ in $\mathcal{R}_{+}^{\infty}$, a basis of neighbourhoods of $\omega$ is obtained by considering for each $k$ the set $V_{k}$ of stable curves contained in $P_{k}$.

Equipped with this topology, $\mathcal{R}_{+}^{\infty}$ is a Cantor set. Each $\mathcal{R}_{+}^{\infty}(P), P \in \mathcal{P}$, is a closed subset, and also a Cantor set. The restriction of $T^{+}$to each $\mathcal{R}_{+}^{\infty}(P)$ is a homeomorphism onto $\mathcal{R}_{+}^{\infty}(a)$ (with $\left.Q \subset R_{a}\right)$.

However, the subset $\mathcal{N}_{+}$may be dense and the map $T^{+}$in general is not continuous on the whole of $\mathcal{D}_{+}$. We will see in the sequel that $\mathcal{N}_{+}$is, in some appropriate sense, negligible.

For each $\omega \in \mathcal{R}_{+}^{\infty}(a)$, we denote by $\varphi_{\omega}$ the $C^{1+\text { Lip }}$ map such that $\omega=\left\{x_{a}=\varphi_{\omega}\left(y_{a}\right)\right\}$; for each $a \in a$, each $y_{a}^{0} \in I_{a}^{u}$, the map

$$
\begin{align*}
\phi_{y_{a}^{0}}: & \mathcal{R}_{+}^{\infty}(a) \mapsto I_{a}^{s}  \tag{10.59}\\
& \omega \mapsto \varphi_{\omega}\left(y_{a}^{0}\right)
\end{align*}
$$

is a homeomorphism onto its image. Letting $y_{a}^{0}$ vary, we get an homeomorphism from $\mathcal{R}_{+}^{\infty}(a) \times I_{a}^{u}$ onto $\widetilde{\mathcal{R}}_{+}^{\infty}(a)$.

The regularity of the partial foliation $\widetilde{\mathcal{R}}_{+}^{\infty}(a)$ is given by
Proposition 40. There exists $C>0$ such that, for all $a \in a$, all distinct $\omega, \omega^{\prime} \in \mathcal{R}_{+}^{\infty}(a)$, all $y$, $y^{\prime} \in I_{a}^{u}$, (cf. Subsection 2.1) we have

$$
\begin{equation*}
\left|\log \frac{\varphi_{\omega}(y)-\varphi_{\omega^{\prime}}(y)}{\varphi_{\omega}\left(y^{\prime}\right)-\varphi_{\omega^{\prime}}\left(y^{\prime}\right)}\right| \leqslant C\left|y-y^{\prime}\right| \tag{10.60}
\end{equation*}
$$

In particular, the homeomorphisms $\phi_{y^{\prime}} \circ \phi_{y}^{-1}$ are bi-Lipschitzian, uniformly in $y, y^{\prime}$.

Proof. The calculations that support the proof will be found in Appendix B. Write $\omega=\cap P_{k}$, $\omega^{\prime}=\cap P_{k}^{\prime}$, where $\left(P_{k}\right)_{k \geqslant 0},\left(P_{k}^{\prime}\right)_{k \geqslant 0}$ are the canonical sequences associated with $\omega, \omega^{\prime}$. We write $\varphi_{k}$, $\varphi_{k}^{\prime}$ for $\varphi_{k}^{+}, \varphi_{k}^{\prime+}$. Let $\ell$ be the largest integer such that $P_{\ell}=P_{\ell}^{\prime}$. There are two cases.

Case 1. $P_{\ell+1}$ or $P_{\ell+1}^{\prime}$ is a simple child of $P_{\ell}$. Let $(A, B)$ be the implicit representation for $\left(P_{\ell}, Q_{\ell}, n_{\ell}\right)$. We have in this case, for all $y \in I_{a}^{u}$

$$
\begin{equation*}
C^{-1}\left|P_{\ell}\right| \leqslant\left|\varphi_{\omega}(y)-\varphi_{\omega^{\prime}}(y)\right| \leqslant C\left|P_{\ell}\right| \tag{10.61}
\end{equation*}
$$

To prove (10.60), we observe (cf. Appendix B) that we can write, for $k>\ell$,

$$
\begin{align*}
\varphi_{k}(y) & =A\left(y, \psi_{k}(y)\right)  \tag{10.62}\\
\varphi_{k}^{\prime}(y) & =A\left(y, \psi_{k}^{\prime}(y)\right)
\end{align*}
$$

with $\psi_{k}, \psi_{k}^{\prime}$ uniformly $C^{1}$-bounded and

$$
\begin{equation*}
\left|\psi_{k}(y)-\psi_{k}^{\prime}(y)\right| \geqslant C>0 \tag{10.63}
\end{equation*}
$$

We then have,

$$
\begin{equation*}
\varphi_{k}^{\prime}(y)-\varphi_{k}(y)=\int_{0}^{1} A_{x}\left(y,(1-t) \psi_{k}(y)+t \psi_{k}^{\prime}(y)\right) d t\left[\psi_{k}^{\prime}(y)-\psi_{k}(y)\right] \tag{10.64}
\end{equation*}
$$

From (10.64) we deduce (10.60) with $\varphi_{\omega}, \varphi_{\omega^{\prime}}$ replaced by $\varphi_{k}, \varphi_{k}^{\prime}$. We then let $k$ goes to $+\infty$ to obtain (10.60) for $\varphi_{\omega}, \varphi_{\omega^{\prime}}$.

Case 2. $P_{\ell+1}$ and $P_{\ell+1}^{\prime}$ are non-simple children of $P_{\ell}$. In this case (10.61) does not hold. The computation in Appendix B shows that

$$
\begin{equation*}
\left|\log \frac{\varphi_{\ell+1}(y)-\varphi_{\ell+1}^{\prime}(y)}{\varphi_{\ell+1}\left(y^{\prime}\right)-\varphi_{\ell+1}^{\prime}\left(y^{\prime}\right)}\right| \leqslant C\left|y-y^{\prime}\right| \tag{10.65}
\end{equation*}
$$

By the same proof as in Case 1, we have

$$
\begin{align*}
& \left|\log \frac{\varphi_{\ell+1}(y)-\varphi_{\omega}(y)}{\varphi_{\ell+1}\left(y^{\prime}\right)-\varphi_{\omega}\left(y^{\prime}\right)}\right| \leqslant C\left|y-y^{\prime}\right|  \tag{10.66}\\
& \left|\log \frac{\varphi_{\ell+1}^{\prime}(y)-\varphi_{\omega}^{\prime}(y)}{\varphi_{\ell+1}^{\prime}\left(y^{\prime}\right)-\varphi_{\omega}^{\prime}\left(y^{\prime}\right)}\right| \leqslant C\left|y-y^{\prime}\right| \tag{10.67}
\end{align*}
$$

As we have

$$
\begin{align*}
\left|\varphi_{\ell+1}(y)-\varphi_{\ell+1}^{\prime}(y)\right| & \leqslant C\left|\varphi_{\omega}(y)-\varphi_{\omega^{\prime}}(y)\right|  \tag{10.68}\\
\left|\varphi_{\ell+1}(y)-\varphi_{\omega}(y)\right| & \leqslant C\left|\varphi_{\omega}(y)-\varphi_{\omega^{\prime}}(y)\right|  \tag{10.69}\\
\left|\varphi_{\ell+1}^{\prime}(y)-\varphi_{\omega^{\prime}}(y)\right| & \leqslant C\left|\varphi_{\omega}(y)-\varphi_{\omega^{\prime}}(y)\right| \tag{10.70}
\end{align*}
$$

the inequality in the proposition follows.

The result of Proposition 40 implies that the transverse Hausdorff dimension $d_{s}=d_{s}(g)$ of $\widetilde{\mathcal{R}}_{+}^{\infty}$ is well-defined, being equal to the Hausdorff dimension of $\phi_{y}\left(\mathcal{R}_{+}^{\infty}(a)\right)$ for any $a \in a, y \in I_{a}^{u}$. We have just proved that it does not depend on $y$. That it does not depend on $a$ is seen as follows: for $\left(a, a^{\prime}\right) \in \mathcal{B}, g$ sends $\widetilde{R}_{+}^{\infty} \cap P_{a a^{\prime}}$ into $\widetilde{R}_{+}^{\infty} \cap R_{a^{\prime}}$; the transverse Hausdorff dimension of $\widetilde{R}_{+}^{\infty} \cap R_{a^{\prime}}$ is therefore not smaller than that of $\widetilde{R}_{+}^{\infty} \cap R_{a}$; as this is true for all $\left(a, a^{\prime}\right) \in \mathcal{B}$, the conclusion follows.

We will also identify below in this section the transverse Hausdorff dimension $d_{s}$ through a transfer operator in the classical manner of Bowen, Ruelle and Sinai.

Another partial control on the geometry of $\widetilde{R}_{+}^{\infty}$ is obtained via the usual estimate on the graph transform, in the setting of uniformly hyperbolic dynamics. Let $\omega, \widehat{\omega}$ be two stable curves and $j>0$. Assume that $\omega$ and $\widehat{\omega}$ belong to the domain of $\left(T^{+}\right)^{j}$. We say that $\omega$ and $\widehat{\omega}$ belong to the same component of the domain of $\left(T^{+}\right)^{j}$ if, for each $0 \leqslant i<j$, there exists a prime $P_{i}$ such that $\left(T^{+}\right)^{i}(\omega)$ and $\left(T^{+}\right)^{i}(\widehat{\omega})$ belong to $\mathcal{R}_{+}^{\infty}\left(P_{i}\right)$.

Proposition 41. There exists $\theta_{0} \in(0,1)$ such that, if $\omega$, $\widehat{\omega}$ belong to the same component of the domain of $\left(T^{+}\right)^{j}$, we have

$$
\left|D \varphi_{\omega}(y)-D \varphi_{\widehat{\omega}}(y)\right| \leqslant C \theta_{0}^{j}
$$

for all $y$.

Proof. We leave the proof of this very standard estimate to the reader.

### 10.6 Transverse Dilatation

This subsection is a preparation for the definition of a transfer operator in the next subsection. The weight function in this transfer operator is, up to a coboundary term, given by a transverse dilatation.

Let $(P, Q, n) \in \mathcal{P}, \omega=\left\{x=\varphi_{\omega}(y)\right\}$ a stable curve in $\mathcal{R}_{+}^{\infty}(P), \omega^{\prime}=T^{+}(\omega)$ its image. Let $(A, B)$ be the implicit representation of $(P, Q, n)$.

For $z=\left(\varphi_{\omega}(y), y\right) \in \omega$, let

$$
\begin{equation*}
v_{\omega}(z)=\frac{\partial}{\partial y}+D \varphi_{\omega}(y) \frac{\partial}{\partial x} \tag{10.71}
\end{equation*}
$$

be the normalized tangent vector to $\omega$ at $z$.
The matrix of $D \widetilde{T}_{+}$at $z$, computed in the bases $\left(\frac{\partial}{\partial x}, v_{\omega}(z)\right)$ at $z,\left(\frac{\partial}{\partial x}, v_{\omega^{\prime}}\left(z^{\prime}\right)\right)$ at $z^{\prime}=\widetilde{T}^{+}(z)$, is lower triangular; the first diagonal coefficient is

$$
\begin{equation*}
A_{x}^{-1}\left(y, x^{\prime}\right)\left(1-B_{x}\left(y, x^{\prime}\right) D \varphi_{\omega^{\prime}}\left(y^{\prime}\right)\right) \tag{10.72}
\end{equation*}
$$

We denote by $\widetilde{b}(z)$ the logarithm of the absolute value of this coefficient. As $\varphi_{\omega}$ is $C^{1+\text { Lip }}$ uniformly in $\omega$ and $g^{n}: P \mapsto Q$ has bounded distortion, we have, for all $z, z^{*} \in \omega$

$$
\begin{equation*}
\left|\widetilde{b}(z)-\widetilde{b}\left(z^{*}\right)\right| \leqslant C\left|z-z^{*}\right| \tag{10.73}
\end{equation*}
$$

Let $j>0$ and let $\omega, \widehat{\omega}$ be stable curves which belong to the same component of the domain of $\left(T^{+}\right)^{j}$. Let $z, \hat{z}$ be point of $\omega, \widehat{\omega}$, respectively, with the same $y$ coordinate. It follows from Proposition 41 that one has

$$
\begin{equation*}
|\widetilde{b}(z)-\widetilde{b}(\hat{z})| \leqslant C \theta_{0}^{j} \tag{10.74}
\end{equation*}
$$

We also have, from the definition of $\widetilde{b}$ :

$$
\begin{equation*}
|\widetilde{b}(z)+\log | P|\mid \leqslant C \tag{10.75}
\end{equation*}
$$

We want to get rid of the dependence of $\widetilde{b}$ on the $y$ coordinate along $\omega$ by adding a coboundary term. We define

$$
\begin{equation*}
\mathcal{D}_{+}^{\infty}=\bigcap_{j \geqslant 0} \operatorname{Dom}\left(T^{+}\right)^{j}=\bigcap_{j \geqslant 0}\left(T^{+}\right)^{-j}\left(\mathcal{D}_{+}\right) . \tag{10.76}
\end{equation*}
$$

For each $a$, fix some $y_{a}^{0} \in I_{a}^{u}$. Then, for $\omega \in \mathcal{D}_{+}^{\infty}, \omega \subset R_{a}, z \in \omega$, define

$$
\begin{equation*}
\Delta b(z)=\sum_{i \geqslant 0}\left(\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i}(z)\right)-\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i}\left(z^{0}\right)\right)\right) \tag{10.77}
\end{equation*}
$$

where $z^{0}$ is the point on $\omega$ with vertical coordinate equal to $y_{a}^{0}$.

From the cone condition, we have, for $i \geqslant 0$ :

$$
\begin{equation*}
\left\|\left(\widetilde{T}^{+}\right)^{i}(z)-\left(\widetilde{T}^{+}\right)^{i}\left(z^{0}\right)\right\| \leqslant C \lambda^{-i} \tag{10.78}
\end{equation*}
$$

The series defining $\Delta b$ is uniformly convergent from (10.73), (10.78), and $\Delta b$ is bounded on $\widetilde{\mathcal{D}}_{+}^{\infty}:=$ $\pi^{-1}\left(\mathcal{D}_{+}^{\infty}\right)$.

Write $z^{1}$ for the point on $T_{+}(\omega) \in R_{a^{\prime}}$ with vertical coordinate $y_{a^{\prime}}^{0}$. We have

$$
\begin{equation*}
\Delta b(z)-\Delta b\left(\widetilde{T}^{+}(z)\right)=\widetilde{b}(z)-b(\omega) \tag{10.79}
\end{equation*}
$$

with

$$
\begin{equation*}
b(\omega)=\widetilde{b}\left(z^{0}\right)+\sum_{i \geqslant 0}\left[\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i+1}\left(z^{0}\right)\right)-\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i}\left(z^{1}\right)\right)\right] . \tag{10.80}
\end{equation*}
$$

We call $b$ the (logarithmic) mean transverse dilatation.
Proposition 42. The mean transverse dilatation $b$, which differs from $\widetilde{b}$ on $\mathcal{D}_{+}^{\infty}$ by the coboundary of the bounded function $\Delta b$, satisfies

$$
\begin{equation*}
|b(\omega)-b(\widehat{\omega})| \leqslant C \theta_{1}^{j} \tag{10.81}
\end{equation*}
$$

if $\omega, \widehat{\omega}$ belong to the same component of the domain of $\left(T^{+}\right)^{j}$. Here $\theta_{1}$ is a fixed constant in $(0,1)$ larger than $\theta_{0}$.

Proof. We have only to prove (10.81). Let $z^{0} \in \omega, z^{1} \in T^{+}(\omega)$ as above and let $\hat{z}^{0} \in \widehat{\omega}, \hat{z}^{1} \in T^{+}(\widehat{\omega})$ be similarly defined. We have, for $i \geqslant 0$ :

$$
\begin{align*}
\left|\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i+1}\left(z^{0}\right)\right)-\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i}\left(z^{1}\right)\right)\right| \leqslant C \lambda^{-i}  \tag{10.82}\\
\left|\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i+1}\left(\hat{z}^{0}\right)\right)-\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i}\left(\hat{z}^{1}\right)\right)\right| \leqslant C \lambda^{-i} \tag{10.83}
\end{align*}
$$

From (10.74), we also have

$$
\begin{equation*}
\left|\widetilde{b}\left(z^{0}\right)-\widetilde{b}\left(\hat{z}^{0}\right)\right| \leqslant C \theta_{0}^{j} \tag{10.84}
\end{equation*}
$$

For $0 \leqslant i<j-1$, we compare $\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i+1}\left(z^{0}\right)\right)$ and $\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i+1}\left(\hat{z}^{0}\right)\right)$ as follows: let $\hat{z}^{i+1}$ be the point on $T_{+}^{i+1}(\widehat{\omega})$ with the same $y$-coordinate as $\left(\widetilde{T}^{+}\right)^{i+1}\left(z^{0}\right)$; one has

$$
\begin{align*}
\left|\widetilde{b}\left(\hat{z}^{i+1}\right)-\widetilde{b}\left(\left(\widetilde{T}^{+}\right)^{i+1}\left(z^{0}\right)\right)\right| & \leqslant C \theta_{0}^{j-i-1}  \tag{10.85}\\
\left\|\hat{z}^{i+1}-\left(\widetilde{T}^{+}\right)^{i+1}\left(\hat{z}^{0}\right)\right\| & \leqslant C \lambda^{i+1-j} \tag{10.86}
\end{align*}
$$

Dealing in the same way with the terms involving $z^{1}, \hat{z}^{1}$, we obtain (10.81) with

$$
\begin{equation*}
\theta_{1}=\max \left(\theta_{0}^{1 / 2}, \lambda^{-1 / 2}\right) \tag{10.87}
\end{equation*}
$$

proving the proposition.

### 10.7 Definition of a Transverse Operator

As $\Delta b$ is bounded, it follows from (10.75) that

$$
\begin{equation*}
|b(\omega)+\log | P|\mid \leqslant C \tag{10.88}
\end{equation*}
$$

for all $(P, Q, n) \in \mathcal{P}, \omega \in \mathcal{R}_{+}^{\infty}(P)$.

It is then a consequence of Corollary 14 in Subsection 10.3 that the series

$$
\begin{equation*}
\sum_{T^{+} \omega^{\prime}=\omega} \exp \left(-d b\left(\omega^{\prime}\right)\right) \tag{10.89}
\end{equation*}
$$

over pre-images $\omega^{\prime}$ of a given stable curve $\omega$ is converging, uniformly in $\omega$, for $d \geqslant d_{s}^{-}$. We will, therefore, define a transfer operator $L_{d}$ for $d \geqslant d_{s}^{-}$as follows: for a bounded function $h$ defined on $\mathcal{D}_{+}^{\infty}$, for $\omega \in \mathcal{D}_{+}^{\infty}$, we set

$$
\begin{equation*}
L_{d} h(\omega)=\sum_{T^{+} \omega^{\prime}=\omega} \exp \left(-d b\left(\omega^{\prime}\right)\right) h\left(\omega^{\prime}\right) \tag{10.90}
\end{equation*}
$$

We can also view this sum over pre-images as a sum over inverse branches of $T^{+}$, which are in one-to-one correspondence with the primes $(P, Q, n)$ such that $Q$ and $\omega$ belong to the same rectangle $R_{a}$. Accordingly, we split the series in two parts: a finite sum corresponding to the trivial primes, (cf. Subsection 10.1), which we denote by $L_{d}^{0}$ and which is defined for all values of $d$, and a perturbative term which we denote by $\Delta L_{d}$. The formula (10.90) defines a bounded operator from the space of bounded functions on $\mathcal{D}_{+}^{\infty}$ into itself, but to have nice spectral properties we need, as usual, to restrict to spaces of slightly more regular functions.

Let $\theta$ be a constant with

$$
\begin{equation*}
\theta_{1}<\theta<1 \tag{10.91}
\end{equation*}
$$

where $\theta_{1}$ comes from Proposition 42 and satisfies $\theta_{1}>\lambda^{-1}$ (cf. 10.87). Denote by $E$ the space of bounded functions $h$ on $\mathcal{D}_{+}^{\infty}$ which satisfy, for some constant $C>0$,

$$
\begin{equation*}
|h(\omega)-h(\widehat{\omega})| \leqslant C \theta^{j} \tag{10.92}
\end{equation*}
$$

whenever $\omega, \widehat{\omega}$ belong to the same component of the domain of $\left(T^{+}\right)^{j}$. We denote by $\|h\|_{\infty}$ the usual norm on bounded functions, by $|h|_{E}$ the best possible $C$ and set

$$
\begin{equation*}
\|h\|_{E}=\max \left(|h|_{E},\|h\|_{\infty}\right) . \tag{10.93}
\end{equation*}
$$

It is clear that $E$ is a Banach space.
Proposition 43. For $d \geqslant d_{s}^{-}, L_{d}$ restricts to a bounded operator on $E$. Moreover, the norm of the perturbative part $\Delta L_{d}$ is as small as we want if $\varepsilon_{0}$ is small enough.

Proof. Let $h \in E, \omega, \widehat{\omega} \in \mathcal{D}_{+}^{\infty}, j>0$. Assume that $\omega, \widehat{\omega}$ belong to the same component of the domain of $\left(T^{+}\right)^{j}$. Let $(P, Q, n)$ be a prime such that $Q, \omega, \widehat{\omega}$ belong to the same rectangle $R_{a}$, and let $\omega_{1}, \widehat{\omega}_{1}$ be the inverse images of $\omega, \widehat{\omega}$ by $T^{+}$corresponding to this inverse branch. By the definition of $\left|\left.\right|_{E}\right.$, we have

$$
\begin{equation*}
\left|h\left(\omega_{1}\right)-h\left(\widehat{\omega}_{1}\right)\right| \leqslant|h|_{E} \theta^{j+1} . \tag{10.94}
\end{equation*}
$$

From Proposition 42, we have

$$
\begin{equation*}
\left|b\left(\omega_{1}\right)-b\left(\widehat{\omega}_{1}\right)\right| \leqslant C \theta_{1}^{j+1} . \tag{10.95}
\end{equation*}
$$

It follows from (10.88) that

$$
\begin{equation*}
\left|\exp \left(-d b\left(\omega_{1}\right)\right)-\exp \left(-d b\left(\widehat{\omega}_{1}\right)\right)\right| \leqslant C(d)|P|^{d} \theta_{1}^{j+1} . \tag{10.96}
\end{equation*}
$$

Putting together (10.94) and (10.96), we have

$$
\begin{equation*}
\left|h\left(\omega_{1}\right) \exp \left(-d b\left(\omega_{1}\right)\right)-h\left(\widehat{\omega}_{1}\right) \exp \left(-d b\left(\widehat{\omega}_{1}\right)\right)\right| \leqslant C(d)|P|^{d}\left(\theta^{j+1}|h|_{E}+\theta_{1}^{j+1}| | h \|_{\infty}\right) . \tag{10.97}
\end{equation*}
$$

Summing over (non trivial) primes yields for $d \geqslant d_{s}^{-}$:

$$
\begin{align*}
\left|\Delta L_{d} h\right|_{E} & <\varepsilon_{1}\|h\|_{E},  \tag{10.98}\\
\left|L_{d} h\right|_{E} & <C\|h\|_{E}, \tag{10.99}
\end{align*}
$$

where $\varepsilon_{1}$ can be made arbitrarily small if $\varepsilon_{0}$ is small enough, according to Corollary 14. The same estimates (for $d \geqslant d_{s}^{-}$) for $\left\|\Delta L_{d} h\right\|_{\infty}$ and $\left\|L_{d} h\right\|_{\infty}$ are easier and can be seen directly. The proposition follows.

### 10.8 Spectral Properties of the Transfer Operator

Let us denote by $\mathcal{R}_{+}^{\infty}(K)$ the set of stable curves $\omega$ which are intersections of a sequence of rectangles belonging to $\mathcal{R}\left(I_{0}\right)$; these stable curves are precisely those which meet the initial horseshoe $K$.

Observe that $\mathcal{R}_{+}^{\infty}(K) \subset \mathcal{D}_{+}^{\infty}$. Denote by $E_{K}$ the space of bounded functions $h$ on $\mathcal{R}_{+}^{\infty}(K)$ which satisfy

$$
\begin{equation*}
|h(\omega)-h(\widehat{\omega})| \leqslant C \theta^{j}, \tag{10.100}
\end{equation*}
$$

whenever $\omega, \widehat{\omega}$ belong to the same component of the domain of $\left(T^{+}\right)^{j}$. Define $|h|_{E_{K}},\|h\|_{E_{K}}$ as above, which makes $E_{K}$ a Banach space.

Let $h \in E$; the restriction of $h$ to $\mathcal{R}_{+}^{\infty}(K)$ belongs to $E_{K}$ and we have

$$
\begin{equation*}
\left\|h / \mathcal{R}_{+}^{\infty}(K)\right\|_{E_{K}} \leqslant\|h\|_{E} . \tag{10.101}
\end{equation*}
$$

The formula for $L_{d}^{0}$ defines a bounded operator, still denoted by $L_{d}^{0}$, on $E_{K}$ and we have a commutative diagram

where $r: E \mapsto E_{K}$ is the restriction operator. The bounded operator $L_{d}^{0}: E_{K} \mapsto E_{K}$ is the subject of the classical theory by Bowen, Ruelle, Sinai for uniformly hyperbolic systems.

Let us recall some standard results of this theory.
a) There is a direct sum invariant decomposition

$$
\begin{equation*}
E_{K}=\mathbb{R} h_{d}^{\prime} \oplus H_{d}^{\prime} \tag{10.103}
\end{equation*}
$$

depending analytically on the parameter $d$, such that $h_{d}^{\prime}$ is a positive eigenfunction, with associated eigenvalue $\lambda_{d}^{\prime}>0$, and such that

$$
\begin{equation*}
s p\left(L_{d}^{0} / H_{d}^{\prime}\right) \subset\left\{|z|<\lambda_{d}^{\prime}\right\} . \tag{10.104}
\end{equation*}
$$

b) There exists a (unique) probability measure $\mu_{d}^{\prime}$ on $\mathcal{R}_{+}^{\infty}(K)$ such that

$$
\begin{equation*}
H_{d}^{\prime}=\left\{h \in E_{K}, \int h d \mu_{d}^{\prime}=0\right\} . \tag{10.105}
\end{equation*}
$$

One normalizes $h_{d}^{\prime}$ to have $\int h_{d}^{\prime} d \mu_{d}^{\prime}=1$. Then, the probability measure $\nu_{d}^{\prime}=h_{d}^{\prime} \mu_{d}^{\prime}$ is invariant under the restriction of $T^{+}$to $\mathcal{R}_{+}^{\infty}(K)$ (observe that $\widetilde{T}^{+}$on $\widetilde{R}_{+}^{\infty}(K)$ is just the restriction of $g$ ).

Let $E^{0}$ be the kernel of the restriction operator $r: E \mapsto E_{K}$. It is invariant under $L_{d}^{0}$.

Lemma 14. One has, for all $d \in \mathbb{R}$.

$$
s p\left(L_{d}^{0} / E^{0}\right) \subset\left\{|z| \leqslant \theta \lambda_{d}^{\prime}\right\} .
$$

Proof. Let $h \in E^{0}, j \geqslant 0$. We have

$$
\begin{equation*}
\left(L_{d}^{0}\right)^{j} h(\omega)=\sum_{\left(T^{+}\right)^{j}\left(\omega^{\prime}\right)=\omega}^{0} h\left(\omega^{\prime}\right) \exp \left(-d b^{(j)}\left(\omega^{\prime}\right)\right) \tag{10.106}
\end{equation*}
$$

where the symbol $\sum^{0}$ indicates that we only consider inverse branches of $T^{+}$associated with trivial primes. The notation $b^{(j)}$ denotes the Birkhoff sum

$$
\begin{equation*}
b^{(j)}\left(\omega^{\prime}\right)=\sum_{0 \leqslant i<j} b\left(\left(T^{+}\right)^{i}\left(\omega^{\prime}\right)\right) . \tag{10.107}
\end{equation*}
$$

We observe that in the sum in (10.106), each $\omega^{\prime}$ belongs to the same component of the domain of $\left(T^{+}\right)^{j}$ as a stable curve in $\mathcal{R}_{+}^{\infty}(K)$. As $h$ belongs to $E^{0}$, this implies that for such a $\omega^{\prime}$ we have

$$
\begin{equation*}
\left|h\left(\omega^{\prime}\right)\right| \leqslant|h|_{E} \theta^{j} \tag{10.108}
\end{equation*}
$$

On the other hand, we have

$$
\begin{equation*}
\sum^{0} \exp \left(-d b^{(j)}\left(\omega^{\prime}\right)\right) \leqslant C \lambda_{d}^{\prime j} \tag{10.109}
\end{equation*}
$$

and it follows that

$$
\begin{equation*}
\left\|\left(L_{d}^{0}\right)^{j} h\right\|_{\infty} \leqslant C \lambda_{d}^{j} \theta^{j}\|h\|_{E} \tag{10.110}
\end{equation*}
$$

Let $\widehat{\omega} \in \mathcal{R}_{+}^{\infty}$ belong to the same component of the domain of $\left(T^{+}\right)^{\ell}$ as $\omega$. Denote by $\widehat{\omega}^{\prime}$ the inverse image of $\widehat{\omega}$ associated to the same sequence of trivial primes as $\omega^{\prime}$. We have

$$
\begin{equation*}
\left|h\left(\omega^{\prime}\right)-h\left(\widehat{\omega}^{\prime}\right)\right| \leqslant|h|_{E} \theta^{j+\ell} \tag{10.111}
\end{equation*}
$$

and, from Proposition 42

$$
\begin{equation*}
\left|b^{(j)}\left(\omega^{\prime}\right)-b^{(j)}\left(\widehat{\omega}^{\prime}\right)\right| \leqslant C \theta_{1}^{\ell} \tag{10.112}
\end{equation*}
$$

Using also (10.108) and (10.109), we obtain

$$
\begin{equation*}
\left|\left(L_{d}^{0}\right)^{j} h(\omega)-\left(L_{d}^{0}\right)^{j} h(\widehat{\omega})\right| \leqslant C \theta^{j+\ell} \lambda_{d}^{j}\|h\|_{E} \tag{10.113}
\end{equation*}
$$

which implies the statement of the Lemma.

We deduce from Lemma 14 that there is a unique function in $E$, still denoted by $h_{d}^{\prime}$, which restricts to $h_{d}^{\prime}$ on $\mathcal{R}_{+}^{\infty}(K)$ and satisfies

$$
\begin{equation*}
L_{d}^{0}\left(h_{d}^{\prime}\right)=\lambda_{d}^{\prime} h_{d}^{\prime} \tag{10.114}
\end{equation*}
$$

Moreover, defining a supplementary hyperplane by

$$
\begin{equation*}
H_{d}^{\prime \prime}=r^{-1}\left(H_{d}^{\prime}\right) \oplus E^{0} \tag{10.115}
\end{equation*}
$$

we have that $H_{d}^{\prime \prime}$ is invariant under $L_{d}^{0}$ and

$$
\begin{equation*}
s p\left(L_{d}^{0} / H_{d}^{\prime \prime}\right) \subset\left\{|z| \leqslant \lambda_{d}^{\prime \prime}\right\} \tag{10.116}
\end{equation*}
$$

where $\lambda_{d}^{\prime \prime}<\lambda_{d}^{\prime}$ is independent of $\varepsilon_{0}$.
Using Proposition 43, we now consider $L_{d}$ itself, assuming that $\varepsilon_{0}$ is small enough and $d \geqslant d_{s}^{-}$.

As the norm of the perturbation part $\Delta L_{d}$ is arbitrarily small, we conclude that $L_{d}$ has a positive eigenfunction $h_{d}$, with associated eigenvalue $\lambda_{d}$ arbitrarily close to $\lambda_{d}^{\prime}$, and an invariant supplementary hyperplane $H_{d}$ satisfying

$$
\begin{equation*}
s p\left(L_{d} / H_{d}\right) \subset\left\{|z|<\lambda_{d}\right\} \tag{10.117}
\end{equation*}
$$

Moreover, $h_{d}, \lambda_{d}$ and $H_{d}$ depend analytically on $d$ for $d>d_{s}^{-}$because $L_{d}$ does. We check that

$$
\begin{equation*}
h_{d} \geqslant c^{-1}>0 \tag{10.118}
\end{equation*}
$$

Indeed, the sequence $h^{(n)}=\lambda_{d}^{-n} L_{d}^{n}(1)$ converge to a positive multiple of $h_{d}$. We have

$$
\begin{equation*}
h^{(n)}(\omega)=\sum_{\left(T^{+}\right)^{n}\left(\omega^{\prime}\right)=\omega} \exp \left(-d b^{(n)}\left(\omega^{\prime}\right)\right) \tag{10.119}
\end{equation*}
$$

Let $\omega, \widehat{\omega}$ be elements of $\mathcal{D}_{+}^{\infty}$ in the same rectangle $R_{a}$; let $\omega^{\prime}$, $\widehat{\omega}^{\prime}$ be pre-images of $\omega, \widehat{\omega}$ by $\left(T^{+}\right)^{n}$ associated with the same sequence of primes. We have (cf. 10.112)

$$
\begin{equation*}
\left|b^{(n)}\left(\omega^{\prime}\right)-b^{(n)}\left(\widehat{\omega}^{\prime}\right)\right| \leqslant C \tag{10.120}
\end{equation*}
$$

and it follows that

$$
\begin{equation*}
C^{-1} \leqslant\left(h^{(n)}\left(\omega^{\prime}\right)\right)^{-1} h^{(n)}(\omega) \leqslant C \tag{10.121}
\end{equation*}
$$

This implies (10.118). One normalizes $h_{d}$ in order to have

$$
\begin{equation*}
h_{d}=\lim _{n \rightarrow+\infty} \lambda_{d}^{-n} L_{d}^{n}(1) \tag{10.122}
\end{equation*}
$$

Denote then by $\mu_{d}$ the linear form on $E$ with kernel $H_{d}$ normalized by $\mu_{d}\left(h_{d}\right)=1$. We have, for all $h \in E$

$$
\begin{equation*}
\lim _{n \rightarrow \infty} \lambda_{d}^{-n} L_{d}^{n} h=\mu_{d}(h) h_{d} \tag{10.123}
\end{equation*}
$$

As $L_{d}$ is a positive operator, $\mu_{d}$ is positive. Observe also that for all $(P, Q, n) \in \mathcal{R}$, the characteristic function $\chi_{P}$ (equal to 1 if $\omega \subset P, 0$ otherwise) belongs to $E$ and satisfies $L^{n} \chi_{P}>0$ everywhere for some $n>0$. Therefore, there exists a unique probability measure on $\mathcal{R}_{+}^{\infty}$, still denoted by $\mu_{d}$, which coincides with $\mu_{d}$ on the intersection of $E$ with $C\left(\mathcal{R}_{+}^{\infty}\right)$.

### 10.9 The Gibbs Measure

From the defining property (10.123) of $\mu_{d}$, we have, for all $h \in E$

$$
\begin{equation*}
\mu_{d}\left(L_{d} h\right)=\lambda_{d} \mu_{d}(h) . \tag{10.124}
\end{equation*}
$$

We will now check the classical Jacobian property for $\mu_{d}$.
Let $(P, Q, n) \in \mathcal{R}$. Let $P_{1} \supset P$ be the thinnest prime containing $P$. The application $T^{+}$is injective on the set $\mathcal{R}_{+}^{\infty}(P)$ formed by the $\omega \in \mathcal{R}_{+}^{\infty}\left(P_{1}\right)$ which are contained in $P$. The image of this set by $T^{+}$is exactly the set of stable curves contained in $P^{\prime}$, with $T^{+}(P, Q, n)=\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$.

Let $h$ be a function in $E$ which vanishes outside $\mathcal{R}_{+}^{\infty}(P)$. Then, $L_{d} h$ vanishes on any curve not contained in $P^{\prime}$, and satisfies

$$
\begin{equation*}
L_{d} h\left(T^{+} \omega\right)=h(\omega) \exp (-d b(\omega)) \tag{10.125}
\end{equation*}
$$

for $\omega \in \mathcal{R}_{+}^{\infty}(P)$. The relation (10.124) for $h$ is the Jacobian property.
Consider in particular the case where $h$ is the characteristic function of $\mathcal{R}_{+}^{\infty}(P)$. We then obtain

$$
\begin{equation*}
\lambda_{d} \mu_{d}\left(\mathcal{R}_{+}^{\infty}(P)\right)=\int_{P^{\prime}} \exp \left(-d b\left(\left(T^{+}\right)^{-1} \omega\right)\right) d \mu_{d}(\omega) . \tag{10.126}
\end{equation*}
$$

We now will specify the value of $d$ by asking that

$$
\begin{equation*}
\lambda_{d}=1 \tag{10.127}
\end{equation*}
$$

Indeed, we have, for $d \geqslant d_{s}^{-}$

$$
\begin{equation*}
\frac{\partial}{\partial d} \lambda_{d}<0 \tag{10.128}
\end{equation*}
$$

This follows from the formula for $L_{d}$ and the fact that $b^{(n)}$ increases at least linearly with $n$ (cf. (10.107)).

We also see easily that

$$
\begin{equation*}
\lim _{d \rightarrow+\infty} \lambda_{d}=0 \tag{10.129}
\end{equation*}
$$

Finally, we have

$$
\begin{equation*}
\lambda_{d_{s}^{-}}>1 . \tag{10.130}
\end{equation*}
$$

Indeed, $d_{s}^{-}$was chosen in order to be smaller than the transverse Hausdorff dimension of $W^{s}(K)$. This means that the eigenvalue $\lambda_{d_{s}^{-}}^{\prime}$ for $\mathrm{L}_{d_{s}^{-}}^{0}$ on $E_{K}$ satisfies $\lambda_{d_{s}^{-}}^{\prime}>1$. As $\Delta L_{d_{s}^{-}}$is also a nonnegative operator, we have $\lambda_{d_{s}^{-}} \geqslant \lambda_{d_{s}^{-}}^{\prime}$. Therefore, (10.130) holds, and it follows from (10.128)-(10.130) that
(10.127) holds for a unique value of $d$. We will denote this value by $d_{s}$. We shall indeed see that $d_{s}$ is the transverse Hausdorff dimension of $\widetilde{\mathcal{R}}_{+}^{\infty}$ which we were able to define in Subsection 10.5.

We just write $\mu$ for the measure $\mu_{d_{s}}$ and $h^{*}$ for the eigenfunction $h_{d_{s}}$.
Proposition 44. For any $(P, Q, n) \in \mathcal{R}$, we have

$$
C^{-1}|P|^{d_{s}} \leqslant \mu(\{\omega \subset P\}) \leqslant C|P|^{d_{s}} .
$$

Proof. Let

$$
\begin{equation*}
(P, Q, n)=\left(P_{1}, Q_{1}, n_{1}\right) * \cdots *\left(P_{r}, Q_{r}, n_{r}\right) \tag{10.131}
\end{equation*}
$$

be the prime decomposition of $(P, Q, n)$. If $\omega \in \mathcal{D}_{+}^{\infty}$ satisfies $\left(T^{+}\right)^{i}(\omega) \in \mathcal{R}_{+}^{\infty}\left(P_{i+1}\right)$ for $0 \leqslant i<r$, we have, from (10.88), (see the definition of $b^{(r)}$ in (10.107)):

$$
\begin{equation*}
C^{-1}|P|^{d_{s}} \leqslant \exp \left(-d_{s} b^{(r)}(\omega)\right) \leqslant C|P|^{d_{s}} \tag{10.132}
\end{equation*}
$$

It, then, follows from the Jacobian property that

$$
\begin{align*}
\mu(\{\omega \subset P\}) & \geqslant C^{-1}|P|^{d_{s}} \mu\left(\left\{\omega \in \mathcal{R}_{+}^{\infty}(a)\right\}\right)  \tag{10.133}\\
& \geqslant C^{-1}|P|^{d_{s}}
\end{align*}
$$

where $R_{a}$ is the rectangle containing $Q$.

For the opposite inequality, we have also to take into account the other inverse branches of $T_{+}^{r}$ when we estimate $L_{d_{s}}^{r}\left(\chi_{P}\right)$, where $\chi_{P}$ is the characteristic function of $\{\omega \subset P\}$. For $0 \leqslant i \leqslant r$, let

$$
\begin{equation*}
\left(P^{i}, Q^{i}, n^{i}\right)=\left(P_{i+1}, Q_{i+1}, n_{i+1}\right) * \cdots *\left(P_{r}, Q_{r}, n_{r}\right) \tag{10.134}
\end{equation*}
$$

(with $\left(P^{r}, Q^{r}, n^{r}\right)=\left(R_{a}, R_{a}, 0\right)$ ). We have

$$
\begin{equation*}
L_{d_{s}} \chi_{P}=\chi_{P}^{1}+\Delta \chi_{P}^{1} \tag{10.135}
\end{equation*}
$$

where

$$
\chi_{P}^{1}\left(\omega^{1}\right)= \begin{cases}0 & \text { if } \omega^{1} \not \subset P^{1}  \tag{10.136}\\ \exp \left(-d_{s} b\left(\omega^{0}\right)\right. & \text { if } \omega^{1}=T^{+}\left(\omega^{0}\right) \text { for some } \omega^{0} \in \mathcal{R}_{+}^{\infty}(P)\end{cases}
$$

and

$$
\begin{equation*}
\Delta \chi_{P}^{1} \leqslant C \sum\left|P_{1}^{*}\right|^{d_{s}}, \tag{10.137}
\end{equation*}
$$

where the sum runs over prime elements $\left(P_{1}^{*}, Q_{1}^{*}, n_{1}^{*}\right)$ with $P_{1}^{*}$ contained in $P$ and distinct from $P$. By Proposition 38 in Subsection 10.3, we obtain

$$
\begin{equation*}
\mu\left(\Delta \chi_{P}^{1}\right) \leqslant C|P|^{d_{s}} \kappa^{\frac{r-1}{2}} \tag{10.138}
\end{equation*}
$$

If $r>1$, we write similarly

$$
\begin{equation*}
L_{d} \chi_{P}^{1}=\chi_{P}^{2}+\Delta \chi_{P}^{2} \tag{10.139}
\end{equation*}
$$

where $\chi_{P}^{2}$ is associated with the inverse branch defined by the prime $P_{2}$ and vanishes outside $P^{2}$. The perturbative term satisfies

$$
\begin{equation*}
\Delta \chi_{P}^{2} \leqslant C\left|P_{1}\right|^{d_{s}} \sum\left|P_{2}^{*}\right|^{d_{s}} \tag{10.140}
\end{equation*}
$$

where the sum now is over primes $P_{2}^{*}$ contained in $P^{1}$ and distinct from $P^{1}$. Proposition 38 now gives

$$
\begin{align*}
\mu\left(\Delta \chi_{P}^{2}\right) & \leqslant C\left|P_{1}\right|^{d_{s}}\left|P^{1}\right|^{d_{s}} \kappa^{\frac{r-2}{2}}  \tag{10.141}\\
& \leqslant C|P|^{d_{s}} \kappa^{\frac{r-2}{2}}
\end{align*}
$$

We iterate this process. At the last step, we have from (10.132)

$$
\begin{equation*}
\mu\left(\chi_{P}^{r}\right) \leqslant C|P|^{d_{s}} \tag{10.142}
\end{equation*}
$$

The contribution of the perturbative terms is bounded by

$$
\begin{equation*}
\mu\left(\sum_{1}^{r} \chi_{P}^{i}\right) \leqslant C|P|^{d_{s}} \sum_{1}^{r} \kappa^{\frac{r-i}{2}} \leqslant C|P|^{d_{s}} \tag{10.143}
\end{equation*}
$$

Corollary 15. The transverse Hausdorff dimension of $\widetilde{\mathcal{R}}_{+}^{\infty}$ is $\leqslant d_{s}$. More precisely, for any curve $\gamma$ which is transverse to $\widetilde{\mathcal{R}}_{+}^{\infty}$, the Hausdorff measure in dimension $d_{s}$ of the intersection of $\gamma$ with $\widetilde{\mathcal{R}}_{+}^{\infty}$ is finite.

We will see below that the transverse Hausdorff dimension is equal to $d_{s}$.

Proof. Let $\delta>0$, choose a finite collection of disjoint rectangles $P_{i}$ with $\left|P_{i}\right| \leqslant \delta$ for each $i$ and $\widetilde{\mathcal{R}}_{+}^{\infty} \subset \cup P_{i}$. We have

$$
\begin{align*}
1=\sum \mu\left(P_{i}\right) & \geqslant C^{-1} \sum\left|P_{i}\right|^{d_{s}}  \tag{10.144}\\
& \geqslant C^{-1} \sum\left[\operatorname{diam}\left(\gamma \cap P_{i}\right)\right]^{d_{s}}
\end{align*}
$$

and the statement of the Corollary follows.

The following statement shows that the dynamics $T^{+}$is only undefined on a small set.
Proposition 45. The transverse Hausdorff dimension of the set $\widetilde{\mathcal{R}}_{+}^{\infty}-\widetilde{\mathcal{D}}_{+}^{\infty}$ is $\leqslant d_{s}^{-}<d_{s}$.

Proof. We have

$$
\begin{equation*}
\widetilde{\mathcal{R}}_{+}^{\infty}-\widetilde{\mathcal{D}}_{+}^{\infty}=\bigcup_{n \geqslant 0}\left(T^{+}\right)^{-n}\left(\mathcal{N}_{+}\right) \tag{10.145}
\end{equation*}
$$

As each $\left(T^{+}\right)^{n}$ has countably many inverse branches which are Lipschitzian, it is sufficient to prove that the transverse Hausdorff dimension of $\mathcal{N}_{+}$is $\leqslant d_{s}^{-}$. But by the definition of $\mathcal{N}_{+}$, for any $\delta>0$, the union of prime rectangles $P$ with $|P|<\delta$ contains $\mathcal{D}$. Proposition 45 then follows from Corollary 14.

### 10.10 Transverse Hausdorff Dimension of $\widetilde{\mathcal{R}}_{+}^{\infty}$

Proposition 46. The transverse Hausdorff dimension of $\widetilde{\mathcal{R}}_{+}^{\infty}$ is the number $d_{s}$ characterized by $\lambda_{d_{s}}=1$.

Remark. We have already seen that the Hausdorff measure in dimension $d_{s}$ of the intersection of $\widetilde{\mathcal{R}}_{+}^{\infty}$ with a transverse curve is always finite. We do not know whether it is positive or always zero.

Proof. Let $\gamma$ be a smooth horizontal-like curve in some $R_{a}$. We denote by $[\gamma]$ the set of stable curves which meet $\gamma$. We will show that

$$
\begin{equation*}
\mu([\gamma]) \leqslant C(\operatorname{diam} \gamma)^{d_{s}}\left(\log (\operatorname{diam} \gamma)^{-1}\right)^{C_{0}} . \tag{10.146}
\end{equation*}
$$

This, being true for all such $\gamma$, clearly implies that the transverse Hausdorff dimension of $\widetilde{\mathcal{R}}_{+}^{\infty}$ is $\geqslant d_{s}$, which is sufficient to prove the proposition in view of Corollary 15.

Clearly, we may assume that $\mu([\gamma])>0$. Define $\left(P_{0}, Q_{0}, n_{0}\right) \in \mathcal{R}$ to be the element such that $P_{0}$ is the thinnest rectangle containing any stable curve in $[\gamma]$. There are at least two children of $P_{0}$ which contain a stable curve in $[\gamma]$. If one of these children is simple, we must have

$$
\begin{equation*}
\operatorname{diam} \gamma \geqslant C^{-1}\left|P_{0}\right| \tag{10.147}
\end{equation*}
$$

and also

$$
\begin{equation*}
\mu([\gamma]) \leqslant \mu\left(\left\{\omega \subset P_{0}\right\}\right) \leqslant C\left|P_{0}\right|^{d_{s}} \tag{10.148}
\end{equation*}
$$

by Proposition 44, which gives the required estimate (and even better). This case is said to have complexity 0 . In the remaining case, denote by $P_{0, i}$ the (non-simple) children of $P_{0}$ which contain a stable curve in $[\gamma]$. Each $P_{0, i}$ is obtained by parabolic composition:

$$
\begin{equation*}
\left(P_{0, i}, Q_{0, i}, n_{0, i}\right) \in\left(P_{0}, Q_{0}, n_{0}\right) \square\left(P_{0, i}^{*}, Q_{0, i}^{*}, n_{0, i}^{*}\right) \tag{10.149}
\end{equation*}
$$

and the widths are related through

$$
\begin{equation*}
C^{-1} \leqslant\left|P_{0, i}\right|\left|P_{0}\right|^{-1}\left|P_{0, i}^{*}\right|^{-1} \delta\left(Q_{0}, P_{0, i}^{*}\right)^{\frac{1}{2}} \leqslant C . \tag{10.150}
\end{equation*}
$$

Let $\delta_{0}=\sup _{i} \delta\left(Q_{0}, P_{0, i}^{*}\right)$; let $\gamma_{1}$ be any horizontal-like curve with the following property: a stable curve meets $\gamma_{1}$ if and only if it is contained in some $P_{0, i}^{*}$. For $\ell \geq 0$, denote by $\gamma_{1, \ell}$ a piece of $\gamma_{1}$ with the property that a stable curve meets $\gamma_{1, \ell}$ iff it is contained in some $P_{0, i}^{*}$ with

$$
\begin{equation*}
\delta_{0} 2^{-\ell-1} \leqslant \delta\left(Q_{0}, P_{0, i}^{*}\right) \leqslant \delta_{0} 2^{-\ell} \tag{10.151}
\end{equation*}
$$

(if there is no such $P_{0, i}^{*}$, take $\gamma_{1, \ell}=\emptyset$ ). We can now write

$$
\begin{align*}
\mu([\gamma]) & \leqslant C \sum_{i}\left|P_{0, i}\right|^{d_{s}} \text { (from Proposition 44) }  \tag{10.152}\\
& \leqslant C\left|P_{0}\right|^{d_{s}} \sum_{i}\left|P_{0, i}^{*}\right|^{d_{s}} \delta\left(Q_{0}, P_{0, i}^{*}\right)^{-\frac{1}{2} d_{s}} \\
& \leqslant C\left|P_{0}\right|^{d_{s}} \delta_{0}^{-\frac{1}{2} d_{s}} \sum_{\ell \geqslant 0} 2^{\frac{\ell d s}{2}} \sum^{(\ell)}\left|P_{0, i}^{*}\right|^{d_{s}} \\
& \leqslant C\left|P_{0}\right|^{d_{s}} \delta_{0}^{-\frac{1}{2} d_{s}} \sum_{\ell \geqslant 0} 2^{\frac{\ell d s}{2}} \mu\left(\left[\gamma_{1, \ell])}\right.\right.
\end{align*}
$$

again by Proposition 44. We have written $\sum^{(\ell)}$ for the partial sum over those $P_{0, i}^{*}$ satisfying (10.151). Assume for some constant $A>0$, that we have, for each $\ell \geqslant 0$ :

$$
\begin{equation*}
\mu\left(\left[\gamma_{1, \ell}\right]\right) \leqslant A\left(\operatorname{diam} \gamma_{1, \ell}\right)^{d_{s}} . \tag{10.153}
\end{equation*}
$$

Observe that we have

$$
\begin{align*}
\operatorname{diam} \gamma_{1, \ell} & \leqslant C 2^{-\ell} \operatorname{diam} \gamma_{1}  \tag{10.154}\\
\operatorname{diam} \gamma & \geqslant C^{-1} \delta_{0}^{-\frac{1}{2}}\left|P_{0}\right| \operatorname{diam} \gamma_{1} \tag{10.155}
\end{align*}
$$

Making use of (10.153)-(10.155) in (10.152) yields

$$
\begin{equation*}
\mu([\gamma]) \leqslant A C(\operatorname{diam} \gamma)^{d_{s}}, \tag{10.156}
\end{equation*}
$$

which is of the same form as (10.153), but with a worse constant $A C$ instead of $A$.
To obtain (10.146), it is thus sufficient to define a complexity index $c(\gamma) \in \mathbb{N}$ which satisfies

$$
\begin{align*}
& c(\gamma) \leqslant c \log \log \left|P_{0}\right|^{-1},  \tag{10.157}\\
& c(\gamma)=1+\sup _{\ell} c\left(\gamma_{1, \ell}\right), \tag{10.158}
\end{align*}
$$

the case of complexity 0 having already been defined and dealt with.
We want to use (10.158) to give an inductive definition of $c(\gamma)$. This will work if the $\gamma_{1, \ell}$ are in some sense "simpler" than $\gamma$. If all $\gamma_{1, \ell}$ have complexity 0 , we just set $c(\gamma)=1$. Assume therefore that some $\gamma_{1, \ell}$ has complexity $>0$. This means that there exists an element ( $P_{1, \ell}, Q_{1, \ell}, n_{1, \ell}$ ) with the following properties:

- each $P_{0, i}^{*}$ related to $\gamma_{1, \ell}$ through (10.151) is contained in some non-simple child of $P_{1, \ell} ;$
- at least two non-simple children of $P_{1, \ell}$ contain some $P_{0, i}^{*}$.

For any parameter interval $I$ containing the given parameter value, $P_{1, \ell}$ is $I$-critical (cf. Proposition 5 in Subsection 6.4). We now distinguish two cases.

Case 1. $P_{0}$ is also $I$-critical (for any $I$ as above).
Let $I^{*}$ be the largest parameter interval for which we have

$$
\begin{equation*}
\left|P_{0}\right|>\left|I^{*}\right|^{\beta} . \tag{10.159}
\end{equation*}
$$

Then, $\left(P_{0}, Q_{0}, n_{0}\right)$ cannot be $I^{*}$-bicritical as $I^{*}$ is $\beta$-regular, hence $Q_{0}$ cannot be $I^{*}$-critical. This implies

$$
\begin{equation*}
\delta_{0} 2^{-\ell} \geqslant C^{-1}\left|I^{*}\right| \geqslant C^{-1}\left|P_{0}\right|^{\beta^{-1}(1+\tau)} . \tag{10.160}
\end{equation*}
$$

As $P_{1, \ell}$ is not transverse to $Q_{0}$ (because $P_{0, i}$ was a child of $P_{0}$ ), we must have from Proposition 10 in Subsection 6.6 that

$$
\begin{equation*}
\left|P_{1, \ell}\right| \geqslant C^{-1}\left(\delta_{0} 2^{-\ell}\right)^{(1-\eta)^{-1}} \tag{10.161}
\end{equation*}
$$

Comparing with (10.160), this guarantees that

$$
\begin{equation*}
\left|P_{0}\right| \ll\left|P_{1, \ell}\right|^{\frac{1}{2}(1+\beta)} . \tag{10.162}
\end{equation*}
$$

This means indeed that every $P_{1, \ell}$ (such that the complexity of $\gamma_{1, \ell}$ is $>0$ ) is indeed simpler than $P_{0}$ and allows us to use (10.158) to define inductively $c(\gamma)$. Observe that the hypothesis of case 1 is always satisfied by the $P_{1, \ell}$. The inequality (10.157) follows from (10.162).

Case 2. $P_{0}$ is $I$-transverse for $I$ small enough.
From Case 1, we have already defined the complexity indices $c\left(\gamma_{1, \ell}\right)$ using (10.158) and again we define $c(\gamma)$ by (10.158). We have to check (10.157) in this case. This will hold if we have, for each $\ell$,

$$
\begin{equation*}
\log \log \left|P_{1, \ell}\right|^{-1} \leqslant c \log \log \left|P_{0}\right|^{-1} \tag{10.163}
\end{equation*}
$$

But (10.161) still holds. We also have

$$
\begin{align*}
\left|Q_{0}\right| & \ll \delta_{0} 2^{-\ell}(\text { from (R7)) }  \tag{10.164}\\
\log \left|Q_{0}\right|^{-1} & \leqslant C n_{0}  \tag{10.165}\\
\log \log \left|P_{0}\right|^{-1} & \geqslant \frac{\log \frac{3}{2}}{\log 2} \log n_{0}-C \tag{10.166}
\end{align*}
$$

from Proposition 13 of Subsection 7.1. Putting this together, we obtain (10.163). The proof of the proposition is now complete.

### 10.11 Invariant Measures

From the Gibbs measure $\mu$, which is not invariant but has the Jacobian property, we define a $T^{+}$-invariant measure $\nu$ on $\mathcal{D}_{+}^{\infty}$ by

$$
\begin{equation*}
d \nu=h^{*} d \mu \tag{10.167}
\end{equation*}
$$

The measure $\nu$ is actually a probability measure on $\mathcal{R}_{+}^{\infty}$ by Proposition 44 and Corollary 14 (see proof of Proposition 45), which implies that

$$
\begin{equation*}
\mu\left(\mathcal{R}_{+}^{\infty}-\mathcal{D}_{+}^{\infty}\right)=0 \tag{10.168}
\end{equation*}
$$

As $h^{*}$ is bounded and bounded away from 0 (cf. (10.118)), the statement of Proposition 44 is also valid for $\nu$ instead of $\mu$. To check that $\nu$ is indeed $T_{+}$-invariant, we first observe that, if $h_{0}, h_{1} \in E$, the product $h_{0} h_{1}$ also belongs to $E$; indeed we have

$$
\begin{equation*}
\left|h_{0} h_{1}\right|_{E} \leqslant\left\|h_{0}\right\|_{\infty}\left|h_{1}\right|_{E}+\left|h_{0}\right|_{E}\left\|h_{1}\right\|_{\infty} \tag{10.169}
\end{equation*}
$$

In particular, for any $h \in E, h h^{*}$ also belongs to $E$. Let $h \in E$. We write

$$
\begin{align*}
\int h\left(T^{+} \omega\right) d \nu(\omega) & =\int h\left(T^{+} \omega\right) h^{*}(\omega) d \mu(\omega)  \tag{10.170}\\
& =\sum_{\mathcal{P}} \int h\left(T^{+} \omega\right) h^{*}(\omega) \chi_{P}^{*}(\omega) d \mu(\omega)
\end{align*}
$$

where $\chi_{P}^{*}$ is the characteristic function of $\mathcal{R}_{+}^{\infty}(P)$. The Jacobian property gives

$$
\begin{equation*}
\int h\left(T^{+} \omega\right) h^{*}(\omega) \chi_{P}^{*}(\omega) d \mu(\omega)=\int h(\omega) h^{*}\left(\omega^{\prime}\right) \exp \left(-d_{s} b\left(\omega^{\prime}\right)\right) d \mu(\omega) \tag{10.171}
\end{equation*}
$$

where $\omega^{\prime}$ is the image of $\omega$ under the inverse branch of $T_{+}$associated with $P$. Summing over $P$ and using that $h^{*}$ is $L_{d_{s}}$-invariant gives

$$
\begin{equation*}
\int h\left(T^{+} \omega\right) d \nu(\omega)=\int h(\omega) d \nu(\omega) \tag{10.172}
\end{equation*}
$$

But $E \cap C\left(\mathcal{R}_{+}^{\infty}\right)$ is dense in the space of continuous functions $C\left(\mathcal{R}_{+}^{\infty}\right)$; the invariance of $\nu$ follows.

Let us now check that the invariant measure $\nu$ is ergodic. Let $A \subset \mathcal{R}_{+}^{\infty}$ be a $T^{+}$-invariant Borel subset with $\nu(A)>0$ and $A^{c}$ its complement. Let $\varepsilon>0$. We will prove that there exists $a \in a$ such that

$$
\begin{equation*}
\nu\left(A \cap \mathcal{R}_{+}^{\infty}(a)\right) \geqslant(1-\varepsilon) \nu\left(\mathcal{R}_{+}^{\infty}(a)\right) \tag{10.173}
\end{equation*}
$$

As $\varepsilon>0$ is arbitrary, this easily implies $\nu(A)=1$.

As $\nu(A)>0$, we can find $(P, Q, n)$ such that

$$
\begin{equation*}
\nu\left(\{\omega \subset P\} \cap A^{c}\right) \leqslant \varepsilon^{\prime} \nu(\{\omega \subset P\}), \tag{10.174}
\end{equation*}
$$

where $\varepsilon^{\prime} \varepsilon^{-1}$ is small. Let $r$ be the number of factors in the prime decomposition of $(P, Q, n)$. Up to a set of measure 0 , we have

$$
\begin{equation*}
\{\omega \subset P\}=\bigcup_{0 \leqslant j \leqslant r} \bigcup_{P_{j}}\left(T^{+}\right)^{-j}\left(\mathcal{R}_{+}^{\infty}\left(P_{j}\right)\right) \bmod 0 \tag{10.175}
\end{equation*}
$$

where $P_{j}$ runs through prime elements satisfying $P_{j} \subset\left(T^{+}\right)^{j}(P)$ and $\left(T^{+}\right)^{-j}$ is the inverse branch of $\left(T^{+}\right)^{j}$ whose image contains $P$. From (10.174), there exists $0 \leqslant j \leqslant r$ and $P_{j}$ such that

$$
\begin{equation*}
\nu\left(A^{c} \cap\left(T^{+}\right)^{-j}\left(\mathcal{R}_{+}^{\infty}\left(P_{j}\right)\right) \leqslant \varepsilon^{\prime} \nu\left(\left(T^{+}\right)^{-j}\left(\mathcal{R}_{+}^{\infty}\left(P_{j}\right)\right)\right) .\right. \tag{10.176}
\end{equation*}
$$

We apply the Jacobian property, taking (10.120) into account to get (10.173) with $\varepsilon=C \varepsilon^{\prime}$. We have proved that $\nu$ is ergodic. We summarize:

Proposition 47. The measure $d \nu=h^{*} d \mu$ is $T^{+}$-invariant, ergodic. It satisfies, for all $(P, Q, n) \in$ $\mathcal{R}$ :

$$
C^{-1}|P|^{d_{s}} \leqslant \nu(\{\omega \subset P\}) \leqslant C|P|^{d_{s}}
$$

and $\nu\left(\mathcal{D}_{+}^{\infty}\right)=1$.

We will now lift $\nu$ to obtain a $\widetilde{T}^{+}$-invariant probability measure on $\widetilde{\mathcal{R}}_{+}^{\infty}$.
Proposition 48. There exists a unique probability measure $\widetilde{\nu}$ on $\widetilde{\mathcal{R}}_{+}^{\infty}$ which is $\widetilde{T}^{+}$-invariant and projects onto $\nu$ under $\pi$. It is ergodic.

Proof. The arguments are standard.

Existence. Denote by $\mathcal{M}(\nu)$ the set of probability measures on $\mathcal{R}_{+}^{\infty}$ which project onto $\nu$. This is a compact set for the weak topology, invariant under $\widetilde{T}^{+}$because $\nu$ is $T^{+}$-invariant. One obtains a $\widetilde{T}^{+}$-invariant measure in $\mathcal{M}(\nu)$ by taking any $\widetilde{\nu}_{0} \in \mathcal{M}(\nu)$ and choosing a weak limit of a subsequence of

$$
\begin{equation*}
\frac{1}{n} \sum_{0}^{n-1}\left[\left(\widetilde{T}^{+}\right)^{j}\right]^{*}\left(\widetilde{\nu}_{0}\right) . \tag{10.177}
\end{equation*}
$$

Uniqueness. The set of fixed points for the action of $\widetilde{T}^{+}$on $\mathcal{M}(\nu)$ is thus non-empty. It is also compact and convex. If it has more than one point, it has at least two distinct extremal points $\widetilde{\nu}_{0}$, $\widetilde{\nu}_{1}$. As $\nu$ is ergodic, $\widetilde{\nu}_{0}$ and $\widetilde{\nu}_{1}$ are also ergodic. Still by the ergodicity of $\nu$, some stable curve $\omega$ must meet the basins of both $\widetilde{\nu}_{0}$ and $\widetilde{\nu}_{1}$. But stable curves are contracted exponentially fast under positive iteration by $T^{+}$; we should thus have $\widetilde{\nu}_{0}=\widetilde{\nu}_{1}$, a contradiction.

We have already said that $\widetilde{\nu}$ is ergodic.

Finally, we want to "spread" the $\widetilde{T}^{+}$-invariant measure $\widetilde{\nu}$ in order to obtain a $g$-invariant measure $\sigma$. Let $\Lambda=\Lambda_{g}$ as in the Introduction (cf. Subsection 1.2).

We first observe that the support of $\widetilde{\nu}$ is contained into $\Lambda \cap \widetilde{\mathcal{R}}_{+}^{\infty}$ : if $N \subset \widetilde{\mathcal{R}}_{+}^{\infty}$ is compact and disjoint from $\Lambda$, then $N$ is disjoint from the image of $\left(\widetilde{T}^{+}\right)^{j}$ if $j$ is large enough, hence $\widetilde{\nu}(N)=0$.

Let now $h$ be a continuous, and thus bounded, function on $\Lambda$. For $x \in \Lambda \cap \mathcal{D}_{+}^{\infty}$, we write

$$
\begin{equation*}
\widetilde{T}^{+}(x)=g^{N(x)}(x) \tag{10.178}
\end{equation*}
$$

where $N(x)=n$ if $x \in \widetilde{\mathcal{R}}_{+}^{\infty}(P)$ with $(P, Q, n) \in \mathcal{P}$. We define:

$$
\begin{equation*}
S h(x)=\sum_{0 \leqslant j<N(x)} h\left(g^{j}(x)\right) \tag{10.179}
\end{equation*}
$$

The function $S h$ is defined $\widetilde{\nu}$-almost everywhere. It satisfies:

$$
\begin{equation*}
|S h(x)| \leqslant\|h\|_{\infty} N(x) \tag{10.180}
\end{equation*}
$$

By Proposition 47 and Corollary 14 in Subsection 10.3 , the function $N$ is $\widetilde{\nu}$-integrable. We have therefore defined an operator

$$
\begin{equation*}
S: C(\Lambda) \mapsto L^{1}(\widetilde{\nu}) \tag{10.181}
\end{equation*}
$$

where $C(\Lambda)$ stands for the space of continuous functions on $\Lambda=\Lambda_{g}$.

We define a finite measure $\sigma$ on $\Lambda$ by

$$
\begin{equation*}
\int h d \sigma=\int S h d \widetilde{\nu} \tag{10.182}
\end{equation*}
$$

for $h \in C(\Lambda)$. From the definition of $S h$, we have

$$
\begin{equation*}
S(h \circ g)=S h+h \circ \widetilde{T}^{+}-h \tag{10.183}
\end{equation*}
$$

Thus, the $\widetilde{T}^{+}$-invariance of $\widetilde{\nu}$ implies that $\sigma$ is $g$-invariant. It is ergodic. The Lyapunov exponents of $\widetilde{T}^{+}$for $\widetilde{\nu}$ are non-zero because $\widetilde{T}^{+}$is uniformly hyperbolic. To get the Lyapunov exponents of $g$ for $\sigma$ we have only to change time, which is possible since $N$ is $\widetilde{\nu}$-integrable.

In the next and last section, we will see that in some appropriate geometric sense, the measure $\sigma$ captures "most" of the dynamics on $\Lambda$, and therefore can be considered as a naturally defined geometric invariant measure on $\Lambda$.

We end this section by observing that everything that has been done for $T^{+}$and positive iteration in Section 10, can also be done for $T^{-}$and negative iteration, leading to another naturally defined geometric invariant measure $\sigma^{-}$on $\Lambda$.

## 11 Some Further Geometric Properties of the Invariant Set

In this final section we pursue the geometric study of the invariant set $\Lambda=\Lambda_{g}$ in two directions. First, we will describe in a rather precise way, both from a dynamical and a geometrical point of view, the intersection of an unstable curve in $\mathcal{R}_{-}^{\infty}$, as defined in Subsection 10.4, with the invariant set $\Lambda$. In the second part of the section, we prove that $\Lambda$ is a saddle-like invariant set in the measure-theoretical sense: both its stable and unstable sets have Lebesgue measure 0 ; thus, no attractors are present in $\Lambda$.

### 11.1 One-Dimensional Analysis of the Invariant Set

Let $\omega^{*} \in \mathcal{R}_{-}^{\infty}$ be an unstable curve as defined in Subsection 10.4. Let $\left(P_{k}^{*}, Q_{k}^{*}, n_{k}^{*}\right)_{k \geqslant 0}$ be the canonical sequence associated to $\omega^{*}$ (cf. definition also in Subsection 10.4). We have

$$
\begin{equation*}
\omega^{*}=\bigcap_{k \geqslant 0} Q_{k}^{*}, \tag{11.1}
\end{equation*}
$$

where $Q_{0}^{*}$ is a rectangle $R_{a}$ and $Q_{k+1}^{*}$ is a child of $Q_{k}^{*}$ for each $k \geqslant 0$. We want to analyze the intersection $\omega^{*} \cap \Lambda$. In Section 10, we have analyzed the set $\widetilde{\mathcal{R}}_{+}^{\infty}$ and we know, in particular, that $\omega^{*} \cap \Lambda$ contains the subset $\omega^{*} \cap \widetilde{\mathcal{R}}_{+}^{\infty}$; this last subset has Hausdorff dimension $d_{s}$ characterized in terms of the transfer operator studied in Section 10; in particular, this dimension is independent of $\omega^{*}$.

Let us summarize the results of our analysis in this section.
Theorem 3. The intersection $\omega^{*} \cap \Lambda$ is the disjoint union of

- a, at most countable, family of Cantor sets $\Lambda_{i}\left(\omega^{*}\right)$,
- a, at most countable, set $\operatorname{Cr}\left(\omega^{*}\right)$,
- an exceptional set $\mathcal{E}\left(\omega^{*}\right)$,
with the following properties
(i) For each $i$, there exists a piece $\omega^{*}(i)$ of $\omega^{*}$ containing $\Lambda_{i}\left(\omega^{*}\right)$, an unstable curve $\omega^{*}{ }_{i}$ and an integer $n_{i}$ such that

$$
\begin{align*}
g^{n_{i}}\left(\omega^{*}(i)\right) & =\omega_{i}^{*},  \tag{11.2}\\
g^{n_{i}}\left(\Lambda_{i}\left(\omega^{*}\right)\right) & =\omega_{i}^{*} \cap \widetilde{\mathcal{R}}_{+}^{\infty} . \tag{11.3}
\end{align*}
$$

In particular, there is a special index $i=0$ for which $n_{0}=0, \omega^{*}(0)=\omega_{0}^{*}=\omega^{*}, \Lambda_{0}\left(\omega^{*}\right)=\omega^{*} \cap \widetilde{\mathcal{R}}_{+}^{\infty}$.
(ii) For every point $c \in C r\left(\omega^{*}\right)$, there exists a stable curve $\omega_{+}(c) \in \mathcal{R}_{+}^{\infty}$, an unstable curve $\omega_{-}(c) \in \mathcal{R}_{-}^{\infty}$, a positive integer $n(c)$ such that $g^{n(c)}(c)$ is a quadratic tangency point between $\omega^{+}(c)$ and $g^{N_{0}}\left(\omega^{-}(c) \cap L_{u}\right)$.
(iii) The Hausdorff dimension of $\mathcal{E}\left(\omega^{*}\right)$ is not greater than

$$
\begin{equation*}
\left(d_{s}^{0}+d_{u}^{0}-1\right) \frac{2 d_{s}^{0}}{2 d_{u}^{0}+d_{s}^{0}}+o(1) \tag{11.4}
\end{equation*}
$$

where the o(1) term is small provided $\tau$ is small enough. Consequently, the Hausdorff dimension of $\omega^{*} \cap \Lambda$ is equal to $d_{s}$.
(iv) Every point $x \in \mathcal{E}\left(\omega^{*}\right)$ is the intersection of a decreasing sequence of pieces $\left(\omega^{*}\left(i_{n}(x)\right)\right)_{n \geqslant 0}$.

## Remark.

1. The structure will be made more precise in the next subsections. We have tried here to extract the most significant features of our analysis.
2. Even with $d_{s}^{0}+d_{u}^{0}>1$, it may happen that $\Lambda$ is a uniformly hyperbolic horseshoe; then, the family $\left(\Lambda_{i}\left(\omega^{*}\right)\right)_{i}$ is finite, $C r\left(\omega^{*}\right)$ and $\mathcal{E}\left(\omega^{*}\right)$ are empty. When $\Lambda$ is not uniformly hyperbolic, the family $\left(\Lambda_{i}\left(\omega^{*}\right)\right)_{i}$ is countable and $\mathcal{E}\left(\omega^{*}\right)$ is a Cantor set; it is not clear in this case if $C r\left(\omega^{*}\right)$ can be empty.

### 11.2 Parabolic Cores

Let $(P, Q, n) \in \mathcal{R}, \mathcal{R}$ as in Subsection 10.1.

Definition. The parabolic core of $P$, denoted by $c(P)$, is the set of points of $W^{s}(\Lambda, \widehat{R})$ which belong to $P$ but not to any child of $P$. The parabolic core of $Q$, denoted by $c(Q)$, is the set of points of $W^{u}(\Lambda, \widehat{R})$ which belong to $Q$ but not to any child of $Q$.

We have partitions

$$
\begin{align*}
R \cap W^{s}(\Lambda, \widehat{R}) & =\bigsqcup_{\mathcal{R}} c(P) \sqcup \widetilde{\mathcal{R}}_{+}^{\infty}  \tag{11.5}\\
R \cap W^{u}(\Lambda, \widehat{R}) & =\bigsqcup_{\mathcal{R}} c(Q) \sqcup \widetilde{\mathcal{R}}_{-}^{\infty} \tag{11.6}
\end{align*}
$$

If $R_{a}$ is the rectangle which contains $\omega^{*}$, we also have

$$
\begin{equation*}
\omega^{*} \cap \Lambda=\bigsqcup_{P \subset R_{a}}\left(\omega^{*} \cap c(P)\right) \sqcup\left(\omega^{*} \cap \widetilde{\mathcal{R}}_{+}^{\infty}\right) \tag{11.7}
\end{equation*}
$$

The parabolic core is empty if and only if $P$ is $I$-decomposable for a small enough parameter interval containing the given strongly regular parameter value. In particular, $c(P)$ is empty if $Q$ is $I$-transverse. Thus, the union in (11.5), (11.7) can be restricted to those $(P, Q, n) \in \mathcal{R}$ such that $Q$ is $I$-critical for all $I$.

We will denote by $C\left(\omega^{*}\right)$ the set of elements $(P, Q, n) \in \mathcal{R}$ such that $c(P) \cap \omega^{*}$ is not empty. For any $(P, Q, n) \in C\left(\omega^{*}\right), Q$ is $I$-critical for all $I$.

### 11.3 Decomposition of $c(P) \cap \omega^{*}$

Let $(P, Q, n) \in C\left(\omega^{*}\right)$. For $k \geqslant 0$, set

$$
\begin{align*}
\left(P_{k}, Q_{k}, n_{k}\right) & =\left(P_{k}^{*}, Q_{k}^{*}, n_{k}^{*}\right) *(P, Q, n),  \tag{11.8}\\
\omega_{P}^{*} & =\bigcap_{k \geqslant 0} Q_{k} . \tag{11.9}
\end{align*}
$$

The unstable curve $\omega_{P}^{*}$ is contained in $Q$ and we have

$$
\begin{equation*}
g^{n}\left(\omega^{*} \cap c(P)\right) \subset \omega_{P}^{*} \cap L_{u} . \tag{11.10}
\end{equation*}
$$

We define a tree $\mathcal{A}\left(\omega^{*}, P\right)$ as follows. The vertices are the rectangles $P^{\prime} \subset P_{s}$ with the following property: for any parameter interval $I$ (containing the given parameter value, say $t$ ), for any $Q_{k} \supset \omega_{P}^{*}, Q_{k}$ and $P^{\prime}$ are not $I$-separated, and $Q_{k}$ and the parent of $P^{\prime}$ are $I$-critically related.

We connect two vertices by an (oriented) edge if one is the parent of the other. We say that a vertex $P^{\prime}$ is critical if, for all $I$ and $Q_{k} \supset \omega_{P}^{*}, Q_{k}$ and $P^{\prime}$ are $I$-critically related. Otherwise, we say that $P^{\prime}$ is transverse. The parent of a vertex is always a critical vertex, except if this vertex is $P_{s}$, the root of the tree. When $P^{\prime}$ is a transverse vertex, the smallest integer $k$ such that $Q_{k}, P$ are $I$-transverse for $I$ small enough is called the level of $P^{\prime}$.

Let $P^{\prime}$ be a critical vertex; then, for every parameter interval $I \ni t, P^{\prime}$ is $I$-critical and, therefore, decomposable.

Let $P^{\prime}$ be a transverse vertex of level 0 . We have $Q_{0}=Q$. Therefore, the parabolic composition $(P, Q, n) \square\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ is well defined and produces two children of $P$.

Let $P^{\prime}$ be a transverse vertex of level $k>0$. For all $m \geqslant k$, the parabolic composition $\left(P_{m}, Q_{m}, n_{m}\right) \square\left(P^{\prime}, Q^{\prime}\right.$, is well-defined and produces two elements $\left(P_{m}^{ \pm}, Q_{m}^{ \pm}, n_{m}^{ \pm}\right)$. The formulas

$$
\begin{array}{ll}
\omega_{P, P^{\prime},+}^{*} & :=\cap Q_{m}^{+},  \tag{11.11}\\
\omega_{P, P^{\prime},-}^{*} & :=\cap Q_{m}^{-},
\end{array}
$$

define unstable curves $\omega_{P, P^{\prime}, \pm}^{*}$ contained in $Q^{\prime}$. We also define pieces $\omega^{*}\left(P, P^{\prime}, \pm\right)$ through

$$
\begin{align*}
g^{n_{P, P^{\prime}}}\left(\omega^{*}\left(P, P^{\prime}, \pm\right)\right) & =\omega_{P, P^{\prime}, \pm}^{*}  \tag{11.12}\\
n_{P, P^{\prime}} & :=n+n^{\prime}+N_{0} . \tag{11.13}
\end{align*}
$$

Lemma 15. Let $x$ be a point in $\omega^{*} \cap c(P), y=g^{n+N_{0}}(x)$. Either $y$ belong to a transverse vertex of level $>0$ or it belongs to an infinite decreasing sequence of critical vertices.

Proof. We have $g^{n}(x) \in L_{u}$ (cf. (11.10)), $y \in L_{s} \subset P_{s}$, and $P_{s}$ is the root and a critical vertex of the tree $\mathcal{A}\left(\omega^{*}, P\right)$. We assume that the first possibility in the statement of the lemma does not hold and construct, starting with $P_{s}$, a sequence of critical vertices containing $y$.

Assume that $y$ belongs to a critical vertex $P^{\prime}$. As $P^{\prime}$ is indecomposable and $y \in W^{s}(\Lambda), y$ belongs to some child $P_{1}^{\prime}$ of $P^{\prime}$. This rectangle is a vertex of the tree: otherwise, $Q_{k}$ and $P_{1}^{\prime}$ would be $I$-separated if $I$ and $Q_{k}$ are thin enough, and then $g^{N_{0}}\left(g^{n}\left(\omega^{*}\right) \cap L_{u}\right) \cap P_{1}^{\prime}$ (which contains $y$ ) would be empty. The vertex $P_{1}^{\prime}$ cannot be transverse of level 0 because, as remarked above, the parabolic composition of $(P, Q, n)$ and $\left(P_{1}^{\prime}, Q_{1}^{\prime}, n_{1}^{\prime}\right)$ would produce a child of $P$ containing $x$, contradicting the hypothesis that $x \in c(P)$. Finally $P_{1}^{\prime}$ cannot be transverse of level $>0$ by hypothesis. It must be a critical vertex, and the induction step is complete.

Proposition 49. There is at most one point $x \in \omega^{*} \cap c(P)$ such that $y=g^{n+N_{0}}(x)$ belongs to $a$ decreasing sequence of critical vertices. When such a point exists, the intersection of this decreasing sequence of vertices is a stable curve which intersects $g^{N_{0}}\left(L_{u} \cap \omega_{P}^{*}\right)$ at $y$ as a quadratic tangency point.

Proof. Let $x$ be a point in $\omega^{*} \cap c(P)$ such that $y=g^{n+N_{0}}(x)$ belongs to a decreasing sequence $\left(P_{\ell}^{\prime}\right)_{\ell \geqslant 0}$ of critical vertices. Denote by $\omega_{+}$the stable curve which is the intersection of these critical vertices. For all parameter intervals $I$, all $k \geqslant 0, \ell \geqslant 0, Q_{k}$ and $P_{\ell}^{\prime}$ are $I$-critically related. This implies that

$$
\begin{equation*}
\lim _{\substack{k \rightarrow+\infty \\ \ell \rightarrow+\infty}} \delta\left(Q_{k}, P_{\ell}^{\prime}\right)=0 \tag{11.14}
\end{equation*}
$$

For large $k$ and $\ell$, let $\gamma_{k}$ (resp. ( $\gamma_{\ell}^{\prime}$ ) be the image in $Q_{k}$ (resp. the inverse image in $P_{\ell}^{\prime}$ ) of the intersection of $P_{k}$ with an horizontal curve (resp. the intersection of $Q_{\ell}^{\prime}$ with a vertical curve). By (11.14), the distance between the vertical-like curve $\gamma_{\ell}^{\prime}$ and the tip of the parabolic-like curve $g^{N_{0}}\left(\gamma_{k}\right)$ goes to zero as $k, \ell$ go to $+\infty$. Passing to the limit, we see that $\omega_{+}$has a tangency with $g^{N_{0}}\left(\omega_{P}^{*} \cap L_{u}\right)$. This tangency is quadratic in the following sense (cf. also the remark after the end of the proof): First, $g^{N_{0}}\left(\omega_{P}^{*} \cap L_{u}\right)$ is contained, with the exception of the tangency point, in one of the components of $P_{s}-\omega_{+}$; moreover, the angle between the tangent lines to $\omega_{+}(x), g^{N_{0}}\left(L_{u} \cap \omega_{P}^{*}\right)$ at points on these curves at the same distance and on the same side of the tangency point is of the
same order as this distance to the tangency point. This is a consequence of the uniform estimates (3.21), (3.22) in Subsection 3.5.

As $\omega_{+}$and $g^{N_{0}}\left(L_{u} \cap \omega_{P}^{*}\right)$ meet at only one point, this point must be $y$. If $x^{\prime}$ is a point with the same property as $x$, and we construct $\omega_{+}^{\prime}$ in the same way as $\omega_{+}$, we must have $\omega_{+}=\omega_{+}^{\prime}$ because otherwise $g^{N_{0}}\left(L_{u} \cap \omega_{P}^{*}\right) \cap \omega_{+}$or $g^{N_{0}}\left(L_{u} \cap \omega_{P}^{*}\right) \cap \omega_{+}^{\prime}$ is empty. But, then, we have $y^{\prime}:=g^{n+N_{0}}\left(x^{\prime}\right)=y$ and $x^{\prime}=x$.

Remark. Calculations involving partial derivatives of higher order for the maps $(A, B)$, which implicitly represent elements of $\mathcal{R}$, show that stable curves and unstable curves are actually of class $C^{\infty}$, with uniform estimates in the $C^{k}$ topology for all $k$. Then, quadratic tangency can be taken in the usual sense. However, the calculations involved, especially when considering parabolic composition, are quite long and not very interesting; we decided to stick to the $C^{1+\text { Lip }}$ regularity class, where the notion of "quadratic" tangency, as explained in the proof of Proposition 49, still makes sense.

It is easy to see exactly when a point $x \in \omega^{*} \cap c(P)$ with the property specified in Proposition 49 does exist: a necessary and sufficient condition is that the tree $\mathcal{A}\left(\omega^{*}, P\right)$ is infinite. In this case, the point $x$ will be a point of the set $\operatorname{Cr}\left(\omega^{*}\right)$ in the statement of Theorem 3 and the point $y=g^{n+N_{0}}(x)$ is said to be critical.

Summarizing what we have established so far, two cases may happen:

1) The tree $\mathcal{A}\left(\omega^{*}, P\right)$ is finite. Then, the intersection $\omega^{*} \cap c(P)$ is the finite disjoint union of the sets

$$
\begin{equation*}
\omega^{*}\left(P, P^{\prime}, \pm\right) \cap \Lambda \tag{11.15}
\end{equation*}
$$

where $P^{\prime}$ runs through the vertices of the tree which are transverse of level $>0$. The image under $g^{n_{P, P^{\prime}}}$ of the set (11.15) is the intersection $\omega_{P, P^{\prime}, \pm}^{*} \cap \Lambda$.
2) The tree $\mathcal{A}\left(\omega^{*}, P\right)$ is infinite. Then, the intersection $\omega^{*} \cap c(P)$ is the countable disjoint union of the sets $\omega^{*}\left(P, P^{\prime}, \pm\right) \cap \Lambda$ as above and a single point $x \in \operatorname{Cr}\left(\omega^{*}\right)$. The point $x=x_{P}$ is the limit of the pieces $\omega^{*}\left(P, P^{\prime}, \pm\right)$ (whose diameters goes to 0 as $\left|P^{\prime}\right|$ goes to 0 ).

### 11.4 The Structure of $\omega^{*} \cap \Lambda$

We are now ready to prove all the statements in Theorem 3, stated above in Subsection 11.1, with the exception of (iii) (the estimate on the Hausdorff dimension of $\mathcal{E}\left(\omega^{*}\right)$ ).

The structure of $\omega^{*} \cap \Lambda$ that we are looking for, which is roughly described in Theorem 3, is obtained by iterating the partition (11.7) and the decomposition of $\omega^{*} \cap c(P)$ described in Subsection 11.3.

At the first step, we have partitioned $\omega^{*} \cap \Lambda$ into the following subsets:

- the intersection $\omega^{*} \cap \widetilde{\mathcal{R}}_{+}^{\infty}$; points in this set ar said of type I;
- for each $(P, Q, n) \in C\left(\omega^{*}\right)$ such that $\mathcal{A}\left(P, \omega^{*}\right)$ is infinite, a point $x_{P}$ such that $y_{P}=g^{n+N_{0}}\left(x_{P}\right)$ is critical; such points $x_{P}$ are said of type II;
- for each $(P, Q, n) \in C\left(\omega^{*}\right)$, each vertex $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ of $\mathcal{A}\left(\omega^{*}, P\right)$ which is transverse of level bigger than 0 , each $\varepsilon \in\{+,-\}$, the intersection $\omega^{*}\left(P, P^{\prime}, \varepsilon\right) \cap \Lambda$; the image of this set under $g^{n_{P, P^{\prime}}}$ is the intersection $\omega_{P, P^{\prime}, \varepsilon}^{*} \cap \Lambda$ of another unstable curve with $\Lambda$.

The intersection $\omega_{P, P^{\prime}, \varepsilon}^{*} \cap \Lambda$ will be analyzed in the same way that $\omega^{*} \cap \Lambda$.
Consider a point $z_{0} \in \omega^{*} \cap \Lambda$. If it is of type I, it belongs to the set $\Lambda_{0}\left(\omega^{*}\right):=\omega^{*} \cap \widetilde{\mathcal{R}}_{+}^{\infty}$ of the statement of Theorem 3. If it is of type II, it belongs to $\operatorname{Cr}\left(\omega^{*}\right)$. Assume now that it is of type III. Then, it belongs to some $\omega^{*}\left(P, P^{\prime}, \varepsilon\right) \cap \Lambda$ as above. Define

$$
\begin{equation*}
z_{1}=g^{n_{P, P^{\prime}}}\left(z_{0}\right) \tag{11.16}
\end{equation*}
$$

which belongs to $\omega_{P, P^{\prime}, \varepsilon}^{*} \cap \Lambda=: \omega_{1}^{*}$. This point may in turn be of type I, II, III with respect to $\omega_{1}^{*}$. The process stops if $z_{1}$ is of type I or II; if $z_{1}$ is of type III, it belongs to some piece $\omega_{1}^{*}\left(P_{1}, P_{1}^{\prime}, \varepsilon_{1}\right)$; we define

$$
\begin{equation*}
z_{2}=g^{n_{P_{1}, P_{1}^{\prime}}}\left(z_{1}\right) \tag{11.17}
\end{equation*}
$$

which belongs to $\omega_{2}^{*} \cap \Lambda$, with

$$
\begin{equation*}
\omega_{2}^{*}:=g^{n_{P_{1}}, P_{1}^{\prime}}\left(\omega_{1}^{*}\left(P_{1}, P_{1}^{\prime}, \varepsilon\right)\right) \tag{11.18}
\end{equation*}
$$

Iterating this process lead to one of three possible outcomes:

1) the $z_{k}$ 's are defined and of type III for all $k \geqslant 0$; the corresponding initial points $z_{0}$ form the set $\mathcal{E}\left(\omega^{*}\right)$.
2) the $z_{k}$ 's are defined for $0 \leqslant k \leqslant \ell$ and $z_{\ell}$ is of type I, i.e. it belongs to $\widetilde{\mathcal{R}}_{+}^{\infty}$; let $\left(P_{k}, P_{k}^{\prime}, \varepsilon_{k}\right)$ for $0 \leqslant k<\ell$ be the data involved in the definitions of the $z_{k}$ 's. We collect together the initial points $z_{0}$ 's with the same set of data; such a set form one of the Cantor sets $\Lambda_{i}\left(\omega^{*}\right)$ in Theorem 3.
3) the $z_{k}$ 's are defined for $0 \leqslant k \leqslant \ell$ and $z_{\ell}$ is of type II. Then $z_{0}$ belongs to the set $C r\left(\omega^{*}\right)$.

We have now completely defined the partition of $\omega^{*} \cap \Lambda$ described in Theorem 3. The properties (i), (ii), (iv) follow immediately from the definitions.

### 11.5 Hausdorff Dimension of the Exceptional Set $\mathcal{E}\left(\omega^{*}\right)$

The self-similar structure apparent in the definition of $\mathcal{E}\left(\omega^{*}\right)$ is the key to obtain an estimate of the dimension of this set. More specifically, we have

$$
\begin{equation*}
\mathcal{E}\left(\omega^{*}\right)=\bigsqcup_{\left(P, P^{\prime}, \varepsilon\right)} g^{-n_{P, P^{\prime}}}\left(\mathcal{E}\left(\omega^{*}{ }_{P, P^{\prime}, \varepsilon}\right)\right) \tag{11.19}
\end{equation*}
$$

where $\varepsilon \in\{+,-\}, P$ runs through $C\left(\omega^{*}\right)$ and $P^{\prime}$ through vertices of $\mathcal{A}\left(\omega^{*}, P\right)$ which are transverse of level > 0 .

Lemma 16. The maps

$$
g^{n_{P, P^{\prime}}}: \omega^{*}\left(P, P^{\prime}, \varepsilon\right) \rightarrow \omega_{P, P^{\prime}, \varepsilon}^{*}
$$

have uniformly bounded distortion.

Proof. Let $k$ be an integer larger than the level of the transverse vertex $P^{\prime}$. Then, the parabolic composition of $\left(P_{k}, Q_{k}, n_{k}\right)$ (cf. (11.8)) and $\left(P^{\prime}, Q^{\prime}, n^{\prime}\right)$ is defined and produces an element $\left(P_{k}^{\prime}, Q_{k}^{\prime}, n_{k}^{\prime}\right)$ such that $Q_{k}^{\prime}$ contains $\omega_{P, P^{\prime}, \varepsilon}^{*}$. Let $\gamma_{k}^{*}$ be an horizontal segment in $P_{k}^{*}, \gamma_{k}$ its image under $g^{n_{k}^{*}}, \gamma_{k}^{\prime}$ the image of $\gamma_{k}^{*} \cap P_{k}^{\prime}$ under $g^{n_{k}^{\prime}}$.

The affine-like maps

$$
\begin{equation*}
g^{n_{k}^{*}}: P_{k}^{*} \rightarrow Q_{k}^{*}, \quad g^{n_{k}^{\prime}}: P_{k}^{\prime} \rightarrow Q_{k}^{\prime} \tag{11.20}
\end{equation*}
$$

have bounded distortion, hence the one-dimensional map

$$
\begin{equation*}
g^{n_{k}^{*}} \circ\left(g^{n_{k}^{\prime}}\right)^{-1}: \gamma_{k}^{\prime} \rightarrow \gamma_{k} \tag{11.21}
\end{equation*}
$$

have also uniformly bounded distortion. Letting $k$ go to $+\infty, \gamma_{k}^{\prime}$ converge to $\omega_{P, P^{\prime}, \varepsilon}^{*}$ and $\gamma_{k}$ to $\omega^{*}$ in the $C^{2-\varepsilon}$-topology for all $\varepsilon>0$. The statement of the lemma follows.

Lemma 17. Let

$$
\delta\left(\omega_{P}^{*}, P^{\prime}\right)=\lim _{k \rightarrow+\infty} \delta\left(Q_{k}, P^{\prime}\right)
$$

We have

$$
C^{-1} \leqslant \frac{\operatorname{diam} \omega^{*}\left(P, P^{\prime}, \varepsilon\right)}{|P|\left|P^{\prime}\right|\left(\delta\left(\omega_{P}^{*}, P^{\prime}\right)\right)^{-\frac{1}{2}}} \leqslant C
$$

Proof. As in the proof of Lemma 16, we write

$$
\begin{equation*}
g^{n_{P, P^{\prime}}}=g^{n_{k}^{\prime}} \circ\left(g^{n_{k}^{*}}\right)^{-1} \tag{11.22}
\end{equation*}
$$

From the estimate (3.27) for parabolic composition in Subsection 3.5, we have

$$
\begin{equation*}
C^{-1} \leqslant \frac{\operatorname{diam} \gamma_{k}^{*}}{\left|P_{k}\right|,\left|P^{\prime}\right|\left(\delta\left(Q_{k}, P^{\prime}\right)\right)^{-\frac{1}{2}}} \leqslant C \tag{11.23}
\end{equation*}
$$

We also have, from the estimates on simple composition

$$
\begin{align*}
C^{-1} & \leqslant \frac{\operatorname{diam} \omega^{*}\left(P, P^{\prime}, \varepsilon\right)\left|P_{k}^{*}\right|}{\operatorname{diam} \gamma_{k}^{*}} \leqslant C,  \tag{11.24}\\
C^{-1} & \leqslant \frac{\left|P_{k}\right|}{|P|\left|P_{k}^{*}\right|} \leqslant C . \tag{11.25}
\end{align*}
$$

Multiplying these three inequalities yields the Lemma.

Let us introduce

$$
\begin{equation*}
\chi(d)=\sum_{\left(P, P^{\prime}, \varepsilon\right)}\left[\operatorname{diam} \omega^{*}\left(P, P^{\prime}, \varepsilon\right)\right]^{d} \tag{11.26}
\end{equation*}
$$

If we are able, for some value of $d$, to show that the series defining $\chi$ is convergent and $\chi(d)$ is small, then by (11.19) and Lemma 16, we will deduce that the Hausdorff dimension of $\mathcal{E}\left(\omega^{*}\right)$ is $\leqslant d$.

In order to study $\chi$, we will first fix $P$ in $C\left(\omega^{*}\right)$ and sum over $\left(P^{\prime}, \varepsilon\right)$. As $\varepsilon$ takes only two values, and in view of Lemma 17, we define, for $P \in C\left(\omega^{*}\right)$ :

$$
\begin{equation*}
\chi_{P}(d)=\sum_{P^{\prime}}\left|P^{\prime}\right|^{d} \delta\left(\omega_{P}^{*}, P^{\prime}\right)^{-\frac{1}{2} d} . \tag{11.27}
\end{equation*}
$$

We will then have

$$
\begin{equation*}
\chi(d) \leqslant C \sum_{P}|P|^{d} \chi_{P}(d) \tag{11.28}
\end{equation*}
$$

In the sum (11.27), $P^{\prime}$ is a transverse vertex of level $>0$, and we therefore must have

$$
\begin{equation*}
\delta\left(\omega_{P}^{*}, P^{\prime}\right) \leqslant \delta_{\max }:=\min \left(\varepsilon_{0}, C|Q|^{1-\eta}\right) . \tag{11.29}
\end{equation*}
$$

In the series (11.27), we first sum over those $P^{\prime}$ such that

$$
\begin{equation*}
2^{-\ell} \delta_{\max } \geqslant \delta\left(\omega_{P}^{*}, P^{\prime}\right) \geqslant 2^{-\ell-1} \delta_{\max } \tag{11.30}
\end{equation*}
$$

for some fixed $\ell \geqslant 0$. This allow us to write

$$
\begin{equation*}
\chi_{P}(d) \leqslant C \delta_{\max }^{-\frac{1}{2} d} \sum_{\ell \geqslant 0} 2^{\frac{\ell d}{2}}\left(\sum^{(\ell)}\left|P^{\prime}\right|^{d}\right), \tag{11.31}
\end{equation*}
$$

where $\sum^{(\ell)}$ means that $P^{\prime}$ is constrained by (11.30). We divide $\sum^{(\ell)}$ into two parts.
In the first part, denoted by $\sum_{1}^{(\ell)}$, we consider only those $P^{\prime}$ such that its parent $\widetilde{P}^{\prime}$ satisfies

$$
\begin{equation*}
\left|\widetilde{P}^{\prime}\right| \leqslant 2^{-\ell} \delta_{\max } \tag{11.32}
\end{equation*}
$$

To estimate $\sum_{1}^{(\ell)}\left|P^{\prime}\right|^{d}$, first observe that, with $d$ bounded away from 0 , it follows from Proposition 21 in Subsection 8.1 that the sum of $\left|P^{\prime}\right|^{d}$ over children of a fixed parent $\widetilde{P}^{\prime}$ is bounded by $C\left|\widetilde{P}^{\prime}\right|^{d}$. We must therefore bound $\sum_{1}^{(\ell)}\left|\widetilde{P}^{\prime}\right|^{d}$.

Also, as $\widetilde{P}^{\prime}$ is a critical vertex, $\widetilde{P}^{\prime}$ cannot be very thin: from Proposition 10 in Subsection 6.6 , we have

$$
\begin{equation*}
\left|\widetilde{P}^{\prime}\right| \geqslant C^{-1}\left[\delta\left(\omega_{P}^{*}, P^{\prime}\right)\right]^{(1-\eta)^{-1}} \geqslant C^{-1}\left(\delta_{\max } 2^{-\ell}\right)^{(1-\eta)^{-1}} \tag{11.33}
\end{equation*}
$$

Finally, the number of $\widetilde{P}^{\prime}$ with $\left|\widetilde{P}^{\prime}\right|$ of order $2^{-m-\ell} \delta_{\max }$ is at most $C 2^{m}$ and the integer $m$ here is restricted by (11.33) to the range

$$
\begin{equation*}
1 \leqslant 2^{m} \leqslant C\left(\delta_{\max } 2^{-\ell}\right)^{-\eta(1-\eta)^{-1}} \tag{11.34}
\end{equation*}
$$

We, therefore, obtain for $d$ bounded away from 0 and 1 ,

$$
\begin{align*}
\sum_{1}^{(\ell)}\left|P^{\prime}\right|^{d} & \leqslant C \sum_{1}^{(\ell)}\left|\widetilde{P}^{\prime}\right|^{d}  \tag{11.35}\\
& \leqslant C \delta_{\max }^{d} 2^{-\ell d} \sum_{m} 2^{m(1-d)} \\
& \leqslant C\left(\delta_{\max } 2^{-\ell}\right)^{d-\eta}
\end{align*}
$$

In the second part of $\sum^{(\ell)}$, denoted by $\sum_{2}^{(\ell)}$, we have on the opposite

$$
\begin{equation*}
\left|\widetilde{P}^{\prime}\right|>2^{-\ell} \delta_{\max } \tag{11.36}
\end{equation*}
$$

As $\left|\widetilde{P}^{\prime}\right|>\delta\left(\omega_{P}^{*}, P^{\prime}\right)$, the number of possibilities for $\widetilde{P}^{\prime}$ is now bounded. As each $P^{\prime}$ is a transverse vertex, we must have (by (R7))

$$
\begin{equation*}
\left|P^{\prime}\right| \leqslant C\left(2^{-\ell} \delta_{\max }\right)^{(1-\eta)^{-1}} \tag{11.37}
\end{equation*}
$$

In particular, from $(11.36),(11.37), P^{\prime}$ is a non-simple child of $\widetilde{P}^{\prime}$. From Proposition 21 in Subsection 8.1, the number of $P^{\prime}$ with $\left|P^{\prime}\right|$ of order $2^{-m} \varepsilon_{0}$ is at most $2^{C m \eta}$.

We have

$$
\begin{equation*}
\sum_{2}^{(\ell)}\left|P^{\prime}\right|^{d} \leqslant \varepsilon_{0}^{d} \sum_{m} 2^{-m(d-C \eta)} \leqslant C \varepsilon_{0}^{-C \eta}\left(2^{-\ell} \delta_{\max }\right)^{d-C \eta} \tag{11.38}
\end{equation*}
$$

Putting (11.35) and (11.38) together yields

$$
\begin{equation*}
\sum^{\ell}\left|P^{\prime}\right|^{d} \leqslant C\left(\delta_{\max } 2^{-\ell}\right)^{d-C^{\prime} \eta} \tag{11.39}
\end{equation*}
$$

and introducing this in (11.31) allow us to estimate $\chi_{P}$ :

$$
\begin{equation*}
\chi_{P}(d) \leqslant C \delta_{\max }^{\frac{1}{2} d-C^{\prime} \eta} \tag{11.40}
\end{equation*}
$$

Finally, we obtain

$$
\begin{equation*}
\chi(d) \leqslant C \sum_{C\left(\omega^{*}\right)}|P|^{d}\left[\min \left(\varepsilon_{0},|Q|\right)\right]^{\frac{1}{2} d-C \eta} \tag{11.41}
\end{equation*}
$$

We do not know exactly the set $C\left(\omega^{*}\right)$, but we know that if $(P, Q, n) \in C\left(\omega^{*}\right)$, the parabolic core $c(P)$ is non-empty and $Q$ must be $I$-critical for all parameter intervals $I$ containing the given parameter value.

We use Hölder's inequality to separate the $P$ and $Q$ in (11.41): for any $p, q>1$ such that

$$
\begin{equation*}
\frac{1}{p}+\frac{1}{q}=1 \tag{11.42}
\end{equation*}
$$

we have

$$
\begin{equation*}
\chi(d) \leqslant C \chi_{+}(d)^{\frac{1}{p}} \chi_{-}(d)^{\frac{1}{q}} \tag{11.43}
\end{equation*}
$$

where

$$
\begin{array}{r}
\chi_{+}(d)=\sum_{Q \text { critical }}|P|^{d p} \\
\chi_{-}(d)=\sum_{Q \text { critical }} \min \left(\varepsilon_{0},|Q|\right)^{\left(\frac{1}{2} d-C \eta\right) q} \tag{11.45}
\end{array}
$$

We will choose $d, p, q$ (satisfying (11.42) in order to have $\chi_{+}(d)$ bounded and $\chi_{-}(d)$ small (when $\varepsilon_{0}$ is small). For such a choice, we can conclude that the Hausdorff dimension of $\mathcal{E}\left(\omega^{*}\right)$ is $\leqslant d$.

We now use that the parameter value is strongly regular, more precisely that the eight estimates (SR1), (SR2) of Subsection 9.2 on the size of the critical locus are satisfied.

It is not difficult to deduce from $(\mathrm{SR} 1)_{\widehat{u}}$ that, if

$$
\begin{equation*}
\frac{1}{2} d q>d_{s}^{0}+d_{u}^{0}-1 \tag{11.46}
\end{equation*}
$$

then, $\chi_{-}(d)$ will be small.

From (SR2) ${ }_{\widehat{u}}$, one can also deduce that if

$$
\begin{equation*}
d p \frac{d_{u}^{0}}{d_{s}^{0}}>d_{s}^{0}+d_{u}^{0}-1 \tag{11.47}
\end{equation*}
$$

then $\chi_{+}$will be bounded. The relations (11.42), (11.46), (11.47) are compatible exactly when

$$
\begin{equation*}
d>\left(d_{s}^{0}+d_{u}^{0}-1\right) \frac{2 d_{s}^{0}}{2 d_{u}^{0}+d_{s}^{0}} \tag{11.48}
\end{equation*}
$$

We observe that the right hand size is always $<d_{s}^{0}$. This ends the proof of Theorem 3 .
Remark. The inequalities (11.46), (11.47) should be understood in the following sense: the difference between the left and right-hand sides is much larger than $\tau$ (which is itself much larger than $\eta)$.

### 11.6 The Stable and Unstable Sets of $\Lambda$

Our goal at the end of this final section is to prove that the invariant set $\Lambda$ is a saddle-like object in the following measure-theoretical sense:

Theorem 4. For a strongly regular parameter, both the stable set $W^{s}(\Lambda)$ and the unstable set $W^{u}(\Lambda)$ have Lebesgue measure 0.

The situation is symmetrical and we will deal with the stable set.

We have:

$$
\begin{equation*}
W^{s}(\Lambda)=\bigsqcup_{n \geqslant 0} g^{-n}\left(W^{s}(\Lambda, \widehat{R}) \cap R\right) \tag{11.49}
\end{equation*}
$$

Therefore, it is sufficient to show that $W^{s}(\Lambda, \widehat{R}) \cap R$ has Lebesgue measure 0 . We write

$$
\begin{equation*}
R \cap W^{s}(\Lambda, \widehat{R})=\bigsqcup_{n \geqslant 0}\left(W^{s}(\Lambda, \widehat{R}) \cap R \cap g^{-n}\left(\widetilde{\mathcal{R}}_{+}^{\infty}\right)\right) \sqcup \mathcal{E}^{+} \tag{11.50}
\end{equation*}
$$

with

$$
\begin{equation*}
\mathcal{E}^{+}=\left\{z \in W^{s}(\Lambda, \widehat{R}) \cap R, g^{n}(z) \notin \widetilde{\mathcal{R}}_{+}^{\infty} \text { for all } n \geqslant 0\right\} . \tag{11.51}
\end{equation*}
$$

We have seen in Section 10 that $\widetilde{\mathcal{R}}_{+}^{\infty}$ is Lipschitzian with transverse Hausdorff dimension $d_{s}$. Therefore, the Hausdorff dimension of $\widetilde{\mathcal{R}}_{+}^{\infty}$ is $1+d_{s}$ and its Lebesgue measure is 0 . The same is true of $g^{-n}\left(\widetilde{\mathcal{R}}_{+}^{\infty}\right)$. We have to prove that the Lebesgue measure of $\mathcal{E}^{+}$is equal to 0.

### 11.7 Decomposition of $\mathcal{E}^{+}$

By the definition of $\mathcal{E}^{+}$and of the parabolic cores, we can write

$$
\begin{equation*}
\mathcal{E}^{+}=\bigsqcup_{P_{0}} \mathcal{E}^{+}\left(P_{0}\right) \tag{11.52}
\end{equation*}
$$

where

$$
\begin{equation*}
\mathcal{E}^{+}\left(P_{0}\right)=\mathcal{E}^{+} \cap c\left(P_{0}\right) \tag{11.53}
\end{equation*}
$$

and $\left(P_{0}, Q_{0}, n_{0}\right)$ runs through the set $\mathcal{C}_{-}$of elements of $R$ with $c\left(P_{0}\right) \neq \emptyset$. In particular, $Q_{0}$ is $I$-critical for all $I$ containing the given parameter value.

For any such $P_{0}$, we have

$$
\begin{align*}
g^{n_{0}}\left(\mathcal{E}^{+}\left(P_{0}\right)\right) & \subset Q_{0} \cap L_{u} \cap \mathcal{E}^{+}  \tag{11.54}\\
g^{n_{0}+N_{0}}\left(\mathcal{E}^{+}\left(P_{0}\right)\right) & \subset L_{s} \cap \mathcal{E}^{+} . \tag{11.55}
\end{align*}
$$

For $P_{1} \in \mathcal{C}_{-}$, define

$$
\begin{equation*}
\mathcal{E}^{+}\left(P_{0}, P_{1}\right)=\left\{z \in \mathcal{E}^{+}\left(P_{0}\right), g^{n_{0}+N_{0}}(z) \in c\left(P_{1}\right)\right\} . \tag{11.56}
\end{equation*}
$$

We have a partition

$$
\begin{equation*}
\mathcal{E}^{+}\left(P_{0}\right)=\bigsqcup_{P_{1}} \mathcal{E}^{+}\left(P_{0}, P_{1}\right) \tag{11.57}
\end{equation*}
$$

At step $k$, we have a partition

$$
\begin{equation*}
\mathcal{E}^{+}=\bigsqcup_{P_{0}, \cdots, P_{k}} \mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right) \tag{11.58}
\end{equation*}
$$

where the $\left(P_{i}, Q_{i}, n_{i}\right)$ runs through $\mathcal{C}_{-}$. We write

$$
\begin{align*}
m_{0} & =n_{0}  \tag{11.59}\\
m_{1} & =n_{0}+N_{0}+n_{1} \\
& \vdots \\
m_{j} & =n_{0}+N_{0}+n_{1}+N_{0}+\cdots+n_{j-1}+N_{0}+n_{j} \\
& =m_{j-1}+N_{0}+n_{j} .
\end{align*}
$$

For $0 \leqslant j \leqslant k$, we have

$$
\begin{align*}
g^{m_{j}}\left(\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right)\right) & \subset \quad Q_{j} \cap L_{u} \cap \mathcal{E}^{+}  \tag{11.60}\\
g^{m_{j}+N_{0}}\left(\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right)\right) & \subset L_{s} \cap \mathcal{E}^{+} \tag{11.61}
\end{align*}
$$

We define, for $P_{k+1} \in \mathcal{C}_{-}$

$$
\begin{equation*}
\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}, P_{k+1}\right)=\left\{z \in \mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right), g^{m_{k}+N_{0}}(z) \in c\left(P_{k+1}\right)\right\} \tag{11.62}
\end{equation*}
$$

and we have

$$
\begin{equation*}
\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right)=\bigsqcup_{P_{k+1}} \mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}, P_{k+1}\right) \tag{11.63}
\end{equation*}
$$

However, in order to have $\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right) \neq \emptyset$ strong restrictions on the $P_{i}$ must take place. We have already mentioned that $\left(P_{i}, Q_{i}, n_{i}\right) \in \mathcal{C}_{-}$. This is the only restriction on $\left(P_{0}, Q_{0}, n_{0}\right)$. But, from (11.55), $P_{1}$ must meet $P_{s}$ and we also know that $Q_{1}$ is critical. As the parameter is regular, we must have

$$
\begin{equation*}
\max \left(\left|P_{1}\right|,\left|Q_{1}\right|\right) \leqslant \varepsilon_{0}^{\beta} \tag{11.64}
\end{equation*}
$$

Assume that $\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k+1}\right)$ is not-empty. We already know that $Q_{k+1}$ is critical. It is also true that, for any parameter interval $I$ containing the given parameter value, $Q_{k}$ and $P_{k+1}$ cannot be $I$-transverse: if they were, parabolic composition would produce children of $P_{k}$ whose union contains $\mathcal{E}_{+}\left(P_{0}, \cdots, P_{k+1}\right)$, in contradiction with the definition of the parabolic core of $P_{k}$.

Let $I$ be the smallest parameter interval satisfying

$$
\begin{equation*}
|I| \geqslant 2\left(\max \left|Q_{k}\right|,\left|P_{k+1}\right|\right)^{1-\eta} . \tag{11.65}
\end{equation*}
$$

As $Q_{k}, P_{k+1}$ are not $I$-transverse, it follows from Proposition 9 in Subsection 6.6 that $P_{k+1}$ is $I$-critical. Therefore, we must have

$$
\begin{equation*}
\max \left(\left|P_{k+1}\right|,\left|Q_{k+1}\right|\right)<|I|^{\beta} \tag{11.66}
\end{equation*}
$$

which implies

$$
\begin{equation*}
\max \left(\left|P_{k+1}\right|,\left|Q_{k+1}\right|\right) \leqslant C\left|Q_{k}\right|^{\widetilde{\beta}}, \tag{11.67}
\end{equation*}
$$

with $\widetilde{\beta}=\beta(1-\eta)(1+\tau)^{-1}$. Taking $\widehat{\beta}<\widetilde{\beta}$ but close to $\beta$ and $\varepsilon_{0}$ sufficiently small, (11.67) and (11.64) give

$$
\begin{equation*}
\max \left(\left|P_{j}\right|,\left|Q_{j}\right|\right) \leqslant \varepsilon_{0}^{\widehat{\beta}^{j}} \tag{11.68}
\end{equation*}
$$

### 11.8 Size and Area of Parabolic Cores

Proposition 50. Let $(P, Q, n) \in \mathcal{C}_{-}$. With Leb standing for Lebesgue measure, we have

$$
\begin{align*}
\operatorname{diam}\left(g^{n}(c(P))\right) & \leqslant C|Q|^{\frac{1}{2}(1-\eta)}  \tag{11.69}\\
\operatorname{Leb}\left(g^{n}(c(P))\right) & \leqslant C|Q|^{\frac{3}{2}-\frac{1}{2} \eta}  \tag{11.70}\\
\operatorname{Leb}(c(P)) & \leqslant C|P||Q|^{\frac{1}{2}(1-\eta)} . \tag{11.71}
\end{align*}
$$

Remark. A posteriori, $c(P)$, which is contained in $W^{s}(\Lambda, \widehat{R})$, will have zero Lebesgue measure. However, we estimate here the diameter and Lebesgue measure of a larger set, as will be apparent in the proof.

Proof. We start with a general observation on an affine-map with implicit representation $(A, B)$. The Jacobian of the map is the product $A_{x}^{-1} B_{y}$. The distortion of Lebesgue measure under the map, which is produced by the oscillation of the logarithm of the Jacobian, is, therefore, controlled by the distortion of the affine-like map in the sense of Subsection 3.2. In particular, the distortion of Lebesgue measure by the restriction of iterates corresponding to the elements of $\mathcal{R}$ is uniformly bounded.

Thus, the third inequality (11.71) in the proposition is a consequence of the second. On the other hand, as $g^{n}(c(P)) \subset Q$, the second inequality (11.70) is an obvious consequence of the first. We have, therefore, only to prove (11.69). Set

$$
\begin{equation*}
Z=g^{n+N_{0}}(c(P)) . \tag{11.72}
\end{equation*}
$$

This set is contained in $g^{N_{0}}\left(Q \cap L_{u}\right) \cap W^{s}(\Lambda, \widehat{R})$, and, a fortiori, in $P_{s}$. Let $R(Z)$ be the set of $\left(P^{*}, Q^{*}, n^{*}\right) \in \mathcal{R}$ with $P^{*} \cap Z \neq \emptyset, P^{*} \subset P_{s}$.

By the definition of the parabolic core, $Q$ and $P^{*}$ cannot be $I$-transverse (for any parameter interval $I$ ). Assume that $\left|P^{*}\right|>2|Q|$; it follows from Proposition 10 in Subsection 6.6 that $P^{*}$ is $I$-critical (for any $I$ ); this implies that $P^{*}$ is $I$-decomposable if $I$ is small enough.

We conclude that

$$
\begin{equation*}
Z \subset \bigcup_{\substack{\left(P^{*}, Q^{*}, n^{*}\right) \in R(Z) \\\left|P^{*}\right| \leqslant 2|Q|}}\left[P^{*} \cap g^{N_{0}}\left(Q \cap L_{u}\right)\right] . \tag{11.73}
\end{equation*}
$$

We replace $P^{*} \cap g^{N_{0}}\left(Q \cap L_{u}\right)$ by the larger Jordan domain $V\left(P^{*}, Q\right)$ defined as follows (see figure 9): the boundary of $V\left(P^{*}, Q\right)$ is made of one arc in the boundary of $P^{*}$ and one arc in the boundary of $g^{N_{0}}\left(Q \cap L_{u}\right)$ and the interior of $V\left(P^{*}, Q\right)$ meets both $P^{*}$ and $g^{N_{0}}\left(Q \cap L_{u}\right)$.


Figure 9

Observe that, if $\left(P_{1}^{*}, Q_{1}^{*}, n_{1}^{*}\right),\left(P_{2}^{*}, Q_{2}^{*}, n_{2}^{*}\right)$ belong to $R(Z)$, we have either $V\left(P_{1}^{*}, Q\right) \subset V\left(P_{2}^{*}, Q\right)$ or the opposite inclusion. Moreover, the parabolic geometry of the picture (cf. (3.22)) gives the
estimate

$$
\begin{equation*}
\operatorname{diam} V\left(P^{*}, Q\right) \leqslant C\left(\delta_{L R}\left(Q, P^{*}\right)\right)^{\frac{1}{2}} \tag{11.74}
\end{equation*}
$$

But, for $\left|P^{*}\right| \leqslant 2|Q|$, the fact that $Q$ and $P^{*}$ are not $I$-transverse (for any $I$ ) implies that

$$
\begin{equation*}
\left.\delta_{L R}\left(Q, P^{*}\right)\right) \leqslant C|Q|^{1-\eta} . \tag{11.75}
\end{equation*}
$$

We conclude that

$$
\begin{align*}
\operatorname{diam} Z & \leqslant \sup _{\substack{\left(P^{*}, Q^{*}, n^{*} * \in R(Z) \\
\left|P^{*}\right| \leqslant 2|Q|\right.}} \operatorname{diam} V\left(P^{*}, Q\right)  \tag{11.76}\\
& \leqslant C|Q|^{\frac{1}{2}(1-\eta)}
\end{align*}
$$

Taking the image by the fixed map $g^{-N_{0}}$ yields (11.69).

### 11.9 Proof of Theorem 4

We will estimate first the Lebesgue measure of each domain $\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right)$. We have

$$
\begin{equation*}
g^{m_{k-1}+N_{0}}\left(\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right)\right) \subset c\left(P_{k}\right) . \tag{11.77}
\end{equation*}
$$

We now use that both the fixed map $g^{N_{0}}$ and the affine-like iterates $g^{n_{j}}: P_{j} \rightarrow Q_{j}($ for $0 \leqslant j<k)$ have uniformly bounded distortion with respect to Lebesgue measure. We are, therefore, able to deduce from (11.70) in Proposition 50 that

$$
\begin{equation*}
\operatorname{Leb}\left(\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right)\right) \leqslant C^{k+1}\left|Q_{k}\right|^{\frac{3}{2}-\frac{1}{2} \eta} \prod_{0}^{k} \frac{\left|P_{j}\right|}{\left|Q_{j}\right|} \tag{11.78}
\end{equation*}
$$

By (11.67), we have $\left|P_{j+1}\right| \ll\left|Q_{j}\right|$ for $j>0$ and it is easy to check that this still holds for $j=0$ (using (11.64) if $\left|Q_{0}\right| \geqslant \varepsilon_{0}$, (11.67) otherwise). It then follows from (11.78) that we have (for $k>0$ )

$$
\begin{equation*}
\operatorname{Leb}\left(\mathcal{E}^{+}\left(P_{0}, \cdots, P_{k}\right)\right) \ll\left|P_{0}\right|\left|Q_{k}\right|^{\frac{1}{2}(1-\eta)} . \tag{11.79}
\end{equation*}
$$

To obtain the estimate for $\mathcal{E}^{+}$, we have to sum over sequences $\left(P_{0}, \cdots, P_{k}\right)$. We first estimate, when $\left(P_{0}, Q_{0}, n_{0}\right)$ and ( $P_{k}, Q_{k}, n_{k}$ ) are fixed, how many admissible sequences have these two extremities.

The element ( $P_{k-1}, Q_{k-1}, n_{k-1}$ ) must satisfy

$$
\begin{equation*}
\left|Q_{k-1}\right| \geqslant C^{-1} \max \left(\left|P_{k}\right|\left|Q_{k}\right|\right)^{1 / \widetilde{\beta}} \tag{11.80}
\end{equation*}
$$

and also that $Q_{k-1}, P_{k}$ are not $I$-transverse (for any $I$ ). This implies that there is at most

$$
\begin{equation*}
C\left|Q_{k}\right|^{-\eta / \widetilde{\beta}} \tag{11.81}
\end{equation*}
$$

possibilities for $\left(P_{k-1}, Q_{k-1}, n_{k-1}\right)$. Iterating leads to a total number of sequences (with $P_{0}, Q_{k}$ fixed) which is at most

$$
\begin{equation*}
C\left|Q_{k}\right|^{-C \eta} \tag{11.82}
\end{equation*}
$$

Therefore, we obtain

$$
\begin{equation*}
\operatorname{Leb}\left(\mathcal{E}^{+}\right) \leqslant \sum_{P_{0}, Q_{k}}\left|P_{0}\right|\left|Q_{k}\right|^{\frac{1}{2}-C \eta} \tag{11.83}
\end{equation*}
$$

The sum $\sum\left|P_{0}\right|$ is obviously bounded. The sum $\sum_{Q_{k}}\left|Q_{k}\right|^{\frac{1}{2}-C \eta}$ is arbitrarily small (when $k$ is large). Indeed, we have from (11.68)

$$
\begin{equation*}
\left|Q_{k}\right| \leqslant \varepsilon_{0}^{\widehat{\beta}^{k}} \tag{11.84}
\end{equation*}
$$

On the other hand, recall that $Q_{k}$ is $I$-critical (for any $I$ ). As the parameter is strongly regular, we know, for instance, from $(\mathrm{SR} 1)_{\widehat{u}}$, that the series

$$
\begin{equation*}
\sum_{Q \text { critical }}|Q|^{d} \tag{11.85}
\end{equation*}
$$

is convergent if $d>d_{s}^{0}+d_{u}^{0}-1+o(1)$. As the maximal value taken by $d_{s}^{0}+d_{u}^{0}-1$ under hypothesis $(\mathrm{H} 4)$ is $\frac{1}{5}$, we indeed have

$$
\begin{equation*}
\lim _{k \rightarrow+\infty} \sum_{Q_{k}}\left|Q_{k}\right|^{\frac{1}{2}-C \eta}=0 \tag{11.86}
\end{equation*}
$$

and this concludes the proof of Theorem 4 and, thus, of the Main Theorem in the paper.

We can sum up the results in Sections 10 and 11 by rephrasing our main result as follows:
Theorem 5. Assume (H1)-(H4). Then, for most $g \in \mathcal{U}_{+}, \Lambda_{g} \subset W^{s}\left(\Lambda_{g}\right)$ and $\Lambda_{g} \subset W^{u}\left(\Lambda_{g}\right)$ carry geometric invariant measures, à la Sinai-Ruelle-Bowen [Si, Ru, BR], with non-zero Lyapunov exponents. Both $W^{s}\left(\Lambda_{g}\right)$ and $W^{u}\left(\Lambda_{g}\right)$ have Lebesgue measure zero and thus $\Lambda_{g}$ carries no attractors nor repellors.

## Appendix A

## Composition Formulas for Affine-Like Maps

We mostly recall in this appendix the formulas for the simple and parabolic compositions of implicitly defined affine-like maps.

We follow closely [PY2]. The main difference with [PY2] is that we consider maps depending on a parameter $t$, and we are interested also in some partial derivatives with respect to the parameter.
A. 1 Simple Composition. Here we consider a map $F_{t}:\left(x_{0}, y_{0}\right) \mapsto\left(x_{1}, y_{1}\right)$ implicitly defined by

$$
\left\{\begin{array}{l}
x_{0}=A\left(y_{0}, x_{1}, t\right)  \tag{A.1}\\
y_{1}=B\left(y_{0}, x_{1}, t\right)
\end{array}\right.
$$

and a map $F_{t}^{\prime}:\left(x_{1}, y_{1}\right) \mapsto\left(x_{2}, y_{2}\right)$ implicitly defined by

$$
\left\{\begin{array}{l}
x_{1}=A^{\prime}\left(y_{1}, x_{2}, t\right)  \tag{A.2}\\
y_{2}=B^{\prime}\left(y_{1}, x_{2}, t\right)
\end{array}\right.
$$

The composition $F_{t}^{\prime \prime}=F_{t}^{\prime} \circ F_{t}$ is implicitly defined by

$$
\left\{\begin{array}{l}
x_{0}=A^{\prime \prime}\left(y_{0}, x_{2}, t\right)  \tag{A.3}\\
y_{2}=B^{\prime \prime}\left(y_{0}, x_{2}, t\right)
\end{array}\right.
$$

and we want to relate the partial derivatives of $A^{\prime \prime}, B^{\prime \prime}$ to those of $A, B, A^{\prime}, B^{\prime}$. Set

$$
\begin{equation*}
\Delta:=1-A_{y}^{\prime}\left(y_{1}, x_{2}, t\right) B_{x}\left(y_{0}, x_{1}, t\right) \tag{A.4}
\end{equation*}
$$

When we solve the system (A.1), (A.2) for $x_{1}, y_{1}$, we obtain

$$
\left\{\begin{array}{l}
x_{1}=X\left(y_{0}, x_{2}, t\right)  \tag{A.5}\\
y_{1}=Y\left(y_{0}, x_{2}, t\right)
\end{array}\right.
$$

where the partial derivatives of $X, Y$ are given by

$$
\left\{\begin{array}{l}
X_{x}=A_{x}^{\prime} \Delta^{-1}  \tag{A.6}\\
X_{y}=A_{y}^{\prime} B_{y} \Delta^{-1} \\
X_{t}=\left(A_{t}^{\prime}+A_{y}^{\prime} B_{t}\right) \Delta^{-1} \\
Y_{x}=A_{x}^{\prime} B_{x} \Delta^{-1} \\
Y_{y}=B_{y} \Delta^{-1} \\
Y_{t}=\left(B_{t}+A_{t}^{\prime} B_{x}\right) \Delta^{-1} .
\end{array}\right.
$$

We have

$$
\left\{\begin{array}{l}
A^{\prime \prime}\left(y_{0}, x_{2}, t\right)=A\left(y_{0}, X, t\right)  \tag{A.7}\\
B^{\prime \prime}\left(y_{0}, x_{2}, t\right)=B^{\prime}\left(Y, x_{2}, t\right),
\end{array}\right.
$$

which gives

$$
\left\{\begin{array}{l}
A_{x}^{\prime \prime}=A_{x} A_{x}^{\prime} \Delta^{-1}  \tag{A.8}\\
B_{y}^{\prime \prime}=B_{y}^{\prime} B_{y} \Delta^{-1},
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
A_{y}^{\prime \prime}=A_{y}+A_{x} X_{y}  \tag{A.9}\\
B_{x}^{\prime \prime}=B_{x}^{\prime}+B_{y}^{\prime} Y_{x},
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
A_{t}^{\prime \prime}=A_{t}+A_{x} X_{t}  \tag{A.10}\\
B_{t}^{\prime \prime}=B_{t}^{\prime}+B_{y}^{\prime} Y_{t}
\end{array}\right.
$$

Next, from (A.4), we have

$$
\left\{\begin{array}{l}
-\Delta_{x}=B_{x x} X_{x} A_{y}^{\prime}+B_{x} A_{x y}^{\prime}+B_{x} A_{y y}^{\prime} Y_{x}  \tag{A.11}\\
-\Delta_{y}=A_{y y}^{\prime} Y_{y} B_{x}+A_{y}^{\prime} B_{x y}+A_{y}^{\prime} B_{x x} X_{y} \\
-\Delta_{t}=B_{x t} A_{y}^{\prime}+B_{x x} X_{t} A_{y}^{\prime}+B_{x} A_{y t}^{\prime}+B_{x} A_{y y}^{\prime} Y_{t} .
\end{array}\right.
$$

Taking logarithmic derivatives in (A.8) gives

$$
\begin{align*}
& \partial_{x} \log \left|A_{x}^{\prime \prime}\right|=\partial_{x} \log \left|A_{x}^{\prime}\right|+Y_{x} \partial_{y} \log \left|A_{x}^{\prime}\right|+X_{x} \partial_{x} \log \left|A_{x}\right|-\Delta_{x} \Delta^{-1},  \tag{A.12}\\
& \partial_{y} \log \left|A_{x}^{\prime \prime}\right|=\partial_{y} \log \left|A_{x}\right|+X_{y} \partial_{x} \log \left|A_{x}\right|+Y_{y} \partial_{y} \log \left|A_{x}^{\prime}\right|-\Delta_{y} \Delta^{-1}
\end{align*}
$$

$$
\begin{gather*}
\partial_{t} \log \left|A_{x}^{\prime \prime}\right|=\partial_{t} \log \left|A_{x}\right|+\partial_{t} \log \left|A_{x}^{\prime}\right|+X_{t} \partial_{x} \log \left|A_{x}\right|+Y_{t} \partial_{y} \log \left|A_{x}^{\prime}\right|-\Delta_{t} \Delta^{-1}  \tag{A.14}\\
\partial_{y} \log \left|B_{y}^{\prime \prime}\right|=\partial_{y} \log \left|B_{y}\right|+X_{y} \partial_{x} \log \left|B_{y}\right|+Y_{y} \partial_{y} \log \left|B_{y}^{\prime}\right|-\Delta_{y} \Delta^{-1}  \tag{A.15}\\
\partial_{x} \log \left|B_{y}^{\prime \prime}\right|=\partial_{x} \log \left|B_{y}^{\prime}\right|+Y_{x} \partial_{y} \log \left|B_{y}^{\prime}\right|+X_{x} \partial_{x} \log \left|B_{y}\right|-\Delta_{x} \Delta^{-1} \tag{A.16}
\end{gather*}
$$

Taking derivatives in (A.9) gives

$$
\left\{\begin{array}{l}
A_{y y}^{\prime \prime}=A_{y y}+2 A_{x y} X_{y}+A_{x x} X_{y}^{2}+A_{x} X_{y y}  \tag{A.18}\\
B_{x x}^{\prime \prime}=B_{x x}^{\prime}+2 B_{x y}^{\prime} Y_{x}+B_{y y}^{\prime} Y_{x}^{2}+B_{y}^{\prime} Y_{x x},
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
A_{y t}^{\prime \prime}=A_{y t}+X_{t} A_{x y}+X_{y} A_{x t}+X_{t} X_{y} A_{x x}+A_{x} X_{y t}  \tag{A.19}\\
B_{x t}^{\prime \prime}=B_{x t}^{\prime}+Y_{t} B_{x y}^{\prime}+Y_{x} B_{y t}^{\prime}+Y_{t} Y_{x} B_{y y}^{\prime}+B_{y}^{\prime} Y_{x t}
\end{array}\right.
$$

where the partial derivatives of $X, Y$ are obtained from (A.6):

$$
\begin{equation*}
X_{y y}=B_{y} \Delta^{-1}\left(A_{y y}^{\prime} Y_{y}+A_{y}^{\prime} \partial_{y} \log \left|B_{y}\right|+A_{y}^{\prime} X_{y} \partial_{x} \log \left|B_{y}\right|-A_{y}^{\prime} \Delta_{y} \Delta^{-1}\right), \tag{A.20}
\end{equation*}
$$

$$
\begin{equation*}
Y_{x x}=A_{x}^{\prime} \Delta^{-1}\left(B_{x x} X_{x}+B_{x} \partial_{x} \log \left|A_{x}^{\prime}\right|+B_{x} Y_{x} \partial_{y} \log \left|A_{x}^{\prime}\right|-B_{x} \Delta_{x} \Delta^{-1}\right), \tag{A.21}
\end{equation*}
$$

$$
\begin{equation*}
X_{y t}=B_{y} \Delta^{-1}\left(A_{y y}^{\prime} Y_{t}+A_{y t}^{\prime}+A_{y}^{\prime} \partial_{t} \log \left|B_{y}\right|+A_{y}^{\prime} X_{t} \partial_{x} \log \left|B_{y}\right|-A_{y}^{\prime} \Delta_{t} \Delta^{-1}\right) \tag{A.22}
\end{equation*}
$$

$$
\begin{equation*}
Y_{x t}=A_{x}^{\prime} \Delta^{-1}\left(B_{x x} X_{t}+B_{x t}+B_{x} \partial_{t} \log \left|A_{x}^{\prime}\right|+B_{x} Y_{t} \partial_{y} \log \left|A_{x}^{\prime}\right|-B_{x} \Delta_{t} \Delta^{-1}\right) . \tag{A.23}
\end{equation*}
$$

A. 2 Parabolic Composition. We have now a fold map $G_{t}=G_{+} \circ G_{0} \circ G_{-}$:

$$
\left(x_{u}, y_{u}\right) \xrightarrow{G_{-}}\left(w, y_{u}\right) \xrightarrow{G_{0}}\left(x_{s}, w\right) \xrightarrow{G_{+}}\left(x_{s}, y_{s}\right),
$$

with

$$
\left\{\begin{array}{l}
y_{s}=Y_{s}\left(w, x_{s}, t\right)  \tag{A.24}\\
x_{u}=X_{u}\left(w, y_{u}, t\right)
\end{array}\right.
$$

We also have an affine like map $F_{0}:\left(x_{0}, y_{0}\right) \mapsto\left(x_{u}, y_{u}\right)$ implicitly defined by

$$
\left\{\begin{array}{l}
x_{0}=A_{0}\left(y_{0}, x_{u}, t\right)  \tag{A.26}\\
y_{u}=B_{0}\left(y_{0}, x_{u}, t\right),
\end{array}\right.
$$

and another affine-like map $F_{1}:\left(x_{1}, y_{1}\right) \mapsto\left(x_{1}, y_{1}\right)$ implicitly defined by

$$
\left\{\begin{array}{l}
x_{s}=A_{1}\left(y_{s}, x_{1}, t\right)  \tag{A.27}\\
y_{1}=B_{1}\left(y_{s}, x_{1}, t\right) .
\end{array}\right.
$$

We assume that (PC1), (PC2) in Subsection 3.5 are satisfied. As we have seen in [PY2] and Subsection 3.5, the first step is to write

$$
\left\{\begin{array}{l}
x_{u}=X\left(w, y_{0}, t\right)  \tag{A.28}\\
y_{s}=Y\left(w, x_{1}, t\right)
\end{array}\right.
$$

where the partial derivatives of $X, Y$ are given by

$$
\left\{\begin{array}{l}
X_{w}=X_{u, w} \Delta_{0}^{-1}  \tag{A.29}\\
X_{y}=X_{u, y} B_{0, y} \Delta_{0}^{-1} \\
X_{t}=\left(X_{u, t}+X_{u, y} B_{0, t}\right) \Delta_{0}^{-1} \\
Y_{w}=Y_{s, w} \Delta_{1}^{-1} \\
Y_{x}=Y_{s, x} A_{1, x} \Delta_{1}^{-1} \\
Y_{t}=\left(Y_{s, t}+Y_{s, x} A_{1, t}\right) \Delta_{1}^{-1},
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
\Delta_{0}=1-X_{u, y} B_{0, x}  \tag{A.30}\\
\Delta_{1}=1-Y_{s, x} A_{1, y} .
\end{array}\right.
$$

We set

$$
\left\{\begin{array}{l}
\bar{Y}\left(w, y_{0}, t\right):=B_{0}\left(y_{0}, X, t\right)  \tag{A.31}\\
\bar{X}\left(w, x_{1}, t\right):=A_{1}\left(Y, x_{1}, t\right),
\end{array}\right.
$$

$$
\begin{equation*}
C\left(w, y_{0}, x_{1}, t\right):=w^{2}-\theta(\bar{X}, \bar{Y}, t) . \tag{A.32}
\end{equation*}
$$

The partial derivatives are given by

$$
\begin{align*}
& \left\{\begin{array}{l}
\bar{Y}_{w}=B_{0, x} X_{w} \\
\bar{Y}_{y}=B_{0, y}+B_{0, x} X_{y}=B_{0, y} \Delta_{0}^{-1} \\
\bar{Y}_{t}=B_{0, t}+B_{0, x} X_{t}=\left(B_{0, t}+B_{0, x} X_{u, t}\right) \Delta_{0}^{-1},
\end{array}\right.  \tag{A.33}\\
& \left\{\begin{array}{l}
\bar{X}_{w}=A_{1, y} Y_{w} \\
\bar{X}_{x}=A_{1, x}+A_{1, y} Y_{x}=A_{1, x} \Delta_{1}^{-1} \\
\bar{X}_{t}=A_{1, t}+A_{1, y} Y_{t}=\left(A_{1, t}+A_{1, y} Y_{s, t}\right) \Delta_{1}^{-1},
\end{array}\right. \tag{A.34}
\end{align*}
$$

$$
\left\{\begin{array}{l}
-C_{w}=-2 w+\theta_{x} \bar{X}_{w}+\theta_{y} \bar{Y}_{w}  \tag{A.35}\\
-C_{x}=\theta_{x} \bar{X}_{x} \\
-C_{y}=\theta_{y} \bar{Y}_{y} \\
-C_{t}=\theta_{x} \bar{X}_{t}+\theta_{y} \bar{Y}_{t}+\theta_{t}
\end{array}\right.
$$

We solve

$$
\begin{equation*}
C\left(w, y_{0}, x_{1}, t\right)=0 \tag{A.36}
\end{equation*}
$$

to define

$$
\begin{equation*}
w=W\left(y_{0}, x_{1}, t\right) \tag{А.37}
\end{equation*}
$$

(there are two solutions $W^{+}$and $W^{-}$).

The corresponding branch of the parabolic composition is implicitly defined by

$$
\left\{\begin{array}{l}
x_{0}=A_{0}\left(y_{0}, X\left(W, y_{0}, t\right), t\right)=: A\left(y_{0}, x_{1}, t\right)  \tag{A.38}\\
y_{1}=B_{1}\left(Y\left(W, x_{1}, t\right), x_{1}, t\right)=: B\left(y_{0}, x_{1}, t\right)
\end{array}\right.
$$

The partial derivatives of $A, B, W$ are given by

$$
\left\{\begin{array}{l}
A_{x}=A_{0, x} X_{w} W_{x}  \tag{A.39}\\
A_{y}=A_{0, y}+A_{0, x}\left(X_{y}+X_{w} W_{y}\right) \\
A_{t}=A_{0, t}+A_{0, x}\left(X_{t}+X_{w} W_{t}\right)
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
B_{y}=B_{1, y} Y_{w} W_{y}  \tag{A.40}\\
B_{x}=B_{1, x}+B_{1, y}\left(Y_{x}+Y_{w} W_{x}\right) \\
B_{t}=B_{1, t}+B_{1, y}\left(Y_{t}+Y_{w} W_{t}\right)
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
W_{x}=-C_{x} C_{w}^{-1}  \tag{A.41}\\
W_{y}=-C_{y} C_{w}^{-1} \\
W_{t}=-C_{t} C_{w}^{-1}
\end{array}\right.
$$

Substituting (A.29), (A.41), (A.35), (A.34) in the formulas (A.39)-(A.40) leads to

$$
\left\{\begin{array}{l}
A_{x}=A_{0, x} A_{1, x} C_{w}^{-1} \theta_{x} X_{u, w} \Delta_{0}^{-1} \Delta_{1}^{-1}  \tag{A.42}\\
B_{y}=B_{1, y} B_{0, y} C_{w}^{-1} \theta_{y} Y_{s, w} \Delta_{0}^{-1} \Delta_{1}^{-1}
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
A_{y}=A_{0, y}+A_{0, x} B_{0, y} \Delta_{0}^{-1}\left(X_{u, y}+X_{u, w} \theta_{y} \Delta_{0}^{-1} C_{w}^{-1}\right)  \tag{A.43}\\
B_{x}=B_{1, x}+B_{1, y} A_{1, x} \Delta_{1}^{-1}\left(Y_{s, x}+Y_{s, w} \theta_{x} \Delta_{1}^{-1} C_{w}^{-1}\right)
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
A_{t}=A_{0, t}+A_{0, x} \Delta_{0}^{-1}\left[X_{u, t}+X_{u, y} B_{0, t}+X_{u, w} C_{w}^{-1}\left(\theta_{t}+\theta_{x} \bar{X}_{t}+\theta_{y} \bar{Y}_{t}\right)\right]  \tag{A.44}\\
B_{t}=B_{1, t}+B_{1, y} \Delta_{1}^{-1}\left[Y_{s, t}+Y_{s, x} A_{1, t}+Y_{s, w} C_{w}^{-1}\left(\theta_{t}+\theta_{x} \bar{X}_{t}+\theta_{y} \bar{Y}_{t}\right)\right]
\end{array}\right.
$$

Taking the logarithmic derivatives in the first formula of (A.39), we obtain

$$
\begin{equation*}
\partial_{x} \log \left|A_{x}\right|=W_{x} X_{w} \partial_{x} \log \left|A_{0, x}\right|+W_{x} \partial_{w} \log \left|X_{w}\right|+\partial_{x} \log \left|W_{x}\right| \tag{A.45}
\end{equation*}
$$

$$
\begin{align*}
\partial_{y} \log \left|A_{x}\right| & =\partial_{y} \log \left|A_{0, x}\right|+\partial_{x} \log \left|A_{0, x}\right|\left(X_{y}+X_{w} W_{y}\right)  \tag{A.46}\\
& +\partial_{y} \log \left|X_{w}\right|+W_{y} \partial_{w} \log \left|X_{w}\right|+\partial_{y} \log \left|W_{x}\right|
\end{align*}
$$

$$
\begin{align*}
\partial_{t} \log \left|A_{x}\right| & =\partial_{t} \log \left|A_{0, x}\right|+\partial_{x} \log \left|A_{0, x}\right|\left(X_{t}+X_{w} W_{t}\right)  \tag{A.47}\\
& +\partial_{t} \log \left|X_{w}\right|+W_{t} \partial_{w} \log \left|X_{w}\right|+\partial_{t} \log \left|W_{x}\right|
\end{align*}
$$

From the second formula in (A.39), one gets

$$
\begin{align*}
A_{y y} & =A_{0, y y}+2 A_{0, x y}\left(X_{y}+X_{w} W_{y}\right)+A_{0, x x}\left(X_{y}+X_{w} W_{y}\right)^{2}  \tag{A.48}\\
& +A_{0, x}\left(X_{y y}+2 X_{w y} W_{y}+X_{w w} W_{y}^{2}+X_{w} W_{y y}\right)
\end{align*}
$$

$$
\begin{align*}
A_{y t} & =A_{0, y t}+A_{0, x y}\left(X_{y}+X_{w} W_{t}\right)+A_{0, x t}\left(X_{y}+X_{w} W_{y}\right)  \tag{A.49}\\
& +A_{0, x x}\left(X_{t}+X_{w} W_{t}\right)\left(X_{y}+X_{w} W_{y}\right) \\
& +A_{0, x}\left(X_{y t}+X_{w y} W_{t}+X_{w t} W_{y}+X_{w w} W_{y} W_{t}+X_{w} W_{y t}\right)
\end{align*}
$$

The symmetric formulas for $B$ are

$$
\begin{equation*}
\partial_{y} \log \left|B_{y}\right|=W_{y} Y_{w} \partial_{y} \log \left|B_{1, y}\right|+\partial_{y} \log \left|W_{y}\right|+W_{y} \partial_{w} \log \left|Y_{w}\right|, \tag{A.50}
\end{equation*}
$$

$$
\begin{align*}
B_{x x} & =B_{1, x x}+2 B_{1, x y}\left(Y_{x}+Y_{w} W_{x}\right)+B_{1, y y}\left(Y_{x}+Y_{w} W_{x}\right)^{2}  \tag{A.53}\\
& +B_{1, y}\left(Y_{x x}+2 Y_{w x} W_{x}+Y_{w w} W_{x}^{2}+Y_{w} W_{x x}\right),
\end{align*}
$$

$$
\begin{align*}
B_{x t} & =B_{1, x t}+B_{1, x y}\left(Y_{t}+Y_{w} W_{t}\right)+B_{1, y t}\left(Y_{x}+Y_{w} W_{x}\right)  \tag{A.54}\\
& +B_{1, y y}\left(Y_{x}+Y_{w} W_{x}\right)\left(Y_{t}+Y_{w} W_{t}\right) \\
& +B_{1, y}\left(Y_{x t}+Y_{w x} W_{t}+Y_{w t} W_{x}+Y_{w w} W_{x} W_{t}+Y_{w} W_{x t}\right)
\end{align*}
$$

In formulas (A.45)-(A.54), the partial derivatives of order 2 of $W$ are obtained from (A.41):

$$
\left\{\begin{array}{l}
W_{x x}=-C_{w}^{-1}\left(C_{w w} W_{x}^{2}+2 C_{w x} W_{x}+C_{x x}\right)  \tag{A.55}\\
W_{x y}=-C_{w}^{-1}\left(C_{w w} W_{x} W_{y}+C_{w x} W_{y}+C_{w y} W_{x}+C_{x y}\right) \\
W_{y y}=-C_{w}^{-1}\left(C_{w w} W_{y}^{2}+2 C_{w y} W_{y}+C_{y y}\right) \\
W_{x t}=-C_{w}^{-1}\left(C_{w w} W_{x} W_{t}+C_{w x} W_{t}+C_{w t} W_{x}+C_{x t}\right) \\
W_{y t}=-C_{w}^{-1}\left(C_{w w} W_{y} W_{t}+C_{w y} W_{t}+C_{w t} W_{y}+C_{y t}\right)
\end{array}\right.
$$

The partial derivatives of order 2 of $C$ are obtained from (A.35):

$$
\begin{equation*}
-C_{w t}=\theta_{x t} \bar{X}_{w}+\theta_{x x} \bar{X}_{w} \bar{X}_{t}+\theta_{x y}\left(\bar{X}_{w} \bar{Y}_{t}+\bar{Y}_{w} \bar{X}_{t}\right)+\theta_{y y} \bar{Y}_{w} \bar{Y}_{t}+\theta_{y t} \bar{Y}_{w}+\theta_{x} \bar{X}_{w t}+\theta_{y} \bar{Y}_{w t} \tag{A.58}
\end{equation*}
$$

$$
\left\{\begin{array}{l}
-C_{x x}=\theta_{x x} \bar{X}_{x}^{2}+\theta_{x} \bar{X}_{x x} \\
-C_{x y}=\theta_{x y} \bar{X}_{x} \bar{Y}_{y} \\
-C_{y y}=\theta_{y y} \bar{Y}_{y}^{2}+\theta_{y} \bar{Y}_{y y}
\end{array}\right.
$$

$$
\left\{\begin{array}{l}
-C_{x t}=\theta_{x t} \bar{X}_{x}+\theta_{x x} \bar{X}_{x} \bar{X}_{t}+\theta_{x y} \bar{X}_{x} \bar{Y}_{t}+\theta_{x} \bar{X}_{x t}  \tag{A.60}\\
-C_{y t}=\theta_{y t} \bar{Y}_{y}+\theta_{y y} \bar{Y}_{y} \bar{Y}_{t}+\theta_{x y} \bar{Y}_{y} \bar{X}_{t}+\theta_{y} \bar{Y}_{y t}
\end{array}\right.
$$

The partial derivatives of order 2 of $\bar{X}, \bar{Y}$ are obtained from (A.34):

$$
\begin{align*}
& \left\{\begin{array}{l}
\bar{X}_{w w}=A_{1, y y} Y_{w}^{2}+A_{1, y} Y_{w w} \\
\bar{X}_{w x}=A_{1, x y} Y_{w}+A_{1, y y} Y_{w} Y_{x}+A_{1, y} Y_{w x} \\
\bar{X}_{w t}=A_{1, y t} Y_{w}+A_{1, y y} Y_{w} Y_{t}+A_{1, y} Y_{w t} \\
\bar{X}_{x x}=A_{1, x x}+2 A_{1, x y} Y_{x}+A_{1, y y} Y_{x}^{2}+A_{1, y} Y_{x x} \\
\bar{X}_{x t}=A_{1, x t}+A_{1, x y} Y_{t}+A_{1, y t} Y_{x}+A_{1, y y} Y_{x} Y_{t}+A_{1, y} Y_{x t} .
\end{array}\right.  \tag{A.61}\\
& \left\{\begin{array}{l}
\bar{Y}_{w w}=B_{0, x x} X_{w}^{2}+B_{0, x} X_{w w} \\
\bar{Y}_{w y}=B_{0, x y} X_{w}+B_{0, x x} X_{w} X_{y}+B_{0, x} X_{w y} \\
\bar{Y}_{w t}=B_{0, x t} X_{w}+B_{0, x x} X_{w} X_{t}+B_{0, x} X_{w t} \\
\bar{Y}_{y y}=B_{0, y y}+2 B_{0, x y} X_{y}+B_{0, x x} X_{y}^{2}+B_{0, x} X_{y y} \\
\bar{Y}_{y t}=B_{0, y t}+B_{0, x y} X_{t}+B_{0, x t} X_{y}+B_{0, x x} X_{y} X_{t}+B_{0, x} X_{y t} .
\end{array}\right.
\end{align*}
$$

Finally, from (A.29), we obtain

$$
\left\{\begin{align*}
X_{w w} & =\Delta_{0}^{-1}\left(X_{u, w w}+2 X_{u, w y} \bar{Y}_{w}+X_{u, y y} \bar{Y}_{w}^{2}+X_{u, y} B_{0, x x} X_{w}^{2}\right),  \tag{A.63}\\
X_{w y} & =\Delta_{0}^{-1}\left(X_{u, w y} \bar{Y}_{y}+X_{u, y y} \bar{Y}_{w} \bar{Y}_{y}+X_{u, y} X_{w}\left(B_{0, x y}+B_{0, x x} X_{y}\right)\right), \\
X_{y y} & =\Delta_{0}^{-1}\left(X_{u, y y} \bar{Y}_{y}^{2}+X_{u, y}\left(B_{0, y y}+2 B_{0, x y} X_{y}+B_{0, x x} X_{y}^{2}\right)\right) \\
X_{w t} & =\Delta_{0}^{-1}\left[X_{u, w t}+X_{u, w y} \bar{Y}_{t}+X_{u, y y} \bar{Y}_{w} \bar{Y}_{t}+X_{u, y t} \bar{Y}_{w}\right. \\
& \left.+X_{u, y} X_{w}\left(B_{0, x t}+B_{0, x x} X_{t}\right)\right] \\
X_{y t} & =\Delta_{0}^{-1}\left[X_{u, y t} \bar{Y}_{y}+X_{u, y y} \bar{Y}_{y} \bar{Y}_{t}\right. \\
& \left.+X_{u, y}\left(B_{0, y t}+B_{0, x y} X_{t}+B_{0, x t} X_{y}+B_{0, x x} X_{y} X_{t}\right)\right],
\end{align*}\right.
$$

(A.64)

$$
\left\{\begin{aligned}
Y_{w w} & =\Delta_{1}^{-1}\left(Y_{s, w w}+2 Y_{s, w x} \bar{X}_{w}+Y_{s, x x} \bar{X}_{w}^{2}+Y_{s, x} A_{1, y y} Y_{w}^{2}\right) \\
Y_{w x} & =\Delta_{1}^{-1}\left(Y_{s, w x} \bar{X}_{x}+Y_{s, x x} \bar{X}_{w} \bar{X}_{x}+Y_{s, x} Y_{w}\left(A_{1, x y}+A_{1, y y} Y_{x}\right)\right), \\
Y_{x x} & =\Delta_{1}^{-1}\left(Y_{s, x x} \bar{X}_{x}^{2}+Y_{s, x}\left(A_{1, x x}+2 A_{1, x y} Y_{x}+A_{1, y y} Y_{x}^{2}\right)\right) \\
Y_{w t} & =\Delta_{1}^{-1}\left[Y_{s, w t}+Y_{s, w x} \bar{X}_{t}+Y_{s, x x} \bar{X}_{w} \bar{X}_{t}+Y_{s, x t} \bar{X}_{w}\right. \\
& \left.+Y_{s, x} Y_{w}\left(A_{1, y t}+A_{1, y y} Y_{t}\right)\right] \\
Y_{x t} & =\Delta_{1}^{-1}\left[Y_{s, x t} \bar{X}_{x}+Y_{s, x x} \bar{X}_{x} \bar{X}_{t}\right. \\
& \left.+Y_{s, x}\left(A_{1, x t}+A_{1, x y} Y_{t}+A_{1, y t} Y_{x}+A_{1, y y} Y_{x} Y_{t}\right)\right] .
\end{aligned}\right.
$$

## Appendix B

## On the Lipschitz Regularity of $\widetilde{R}_{+}^{\infty}$

B. 1 In this appendix, we will perform some calculations and obtain some estimates that uphold the proof of Proposition 40 in Subsection 10.5, showing that the holonomy maps of the partial foliation $\widetilde{\mathcal{R}}_{+}^{\infty}$ are uniformly Lipschitz.

As always, we have to consider two cases, dealing separately with affine-like maps and parabolic composition, the later case being more involved than the first.
B. 2 The Case of Affine-Like Maps. We consider here, with the notations of Subsection 3.1, an affine-like map $F$ with implicit representation $(A, B)$ : the domain of $F$ is a vertical strip $P$ in a rectangle $I_{0}^{s} \times I_{0}^{u}$, its image is a horizontal strip in a rectangle $I_{1}^{s} \times I_{1}^{u}$, the respective coordinates being $x_{0}, y_{0}, x_{1}, y_{1}$.

In $I_{1}^{s} \times I_{1}^{s}$, we are given a one-parameter family of vertical-like curves which are graphs

$$
\begin{equation*}
\omega(s)=\left\{x_{1}=\varphi\left(y_{1}, s\right)\right\}, \text { with }\left|\frac{\partial \varphi}{\partial y_{1}}\right| \text { not too large. } \tag{B.1}
\end{equation*}
$$

We assume that, for all $y_{1}, s$

$$
\begin{equation*}
\frac{\partial \varphi}{\partial s}\left(y_{1}, s\right)>0 \tag{B.2}
\end{equation*}
$$

and that we have a uniform bound

$$
\begin{equation*}
\left|\frac{\partial}{\partial y_{1}} \log \frac{\partial \varphi}{\partial s}\left(y_{1}, s\right)\right| \leqslant T \tag{B.3}
\end{equation*}
$$

We define the vertical-like curves in $P$ by

$$
\begin{equation*}
\Omega(s)=F^{-1}(\omega(s) \cap Q) \tag{B.4}
\end{equation*}
$$

We assume that $F$ satisfies a cone condition and we have a good control on the distortion of $F$. Under these hypotheses, $\Omega(s)$ will be a graph

$$
\begin{equation*}
\Omega(s)=\left\{x_{0}=\Phi\left(y_{0}, s\right)\right\} \tag{B.5}
\end{equation*}
$$

and we want to obtain the analogue of (B.3):

$$
\begin{equation*}
\left|\frac{\partial}{\partial y_{0}} \log \right| \frac{\partial \Phi}{\partial s}\left(y_{0}, s\right)\left|\mid \leqslant T^{\prime}\right. \tag{B.6}
\end{equation*}
$$

The relation between $\varphi$ and $\Phi$ is as follows in the equation

$$
\begin{equation*}
y_{1}=B\left(y_{0}, \varphi\left(y_{1}, s\right)\right) \tag{B.7}
\end{equation*}
$$

we solve for $y_{1}$ to define

$$
\begin{equation*}
y_{1}=\psi\left(y_{0}, s\right) . \tag{B.8}
\end{equation*}
$$

We then have

$$
\begin{equation*}
\Phi\left(y_{0}, s\right)=A\left(y_{0}, \varphi\left(\psi\left(y_{0}, s\right), s\right)\right) . \tag{B.9}
\end{equation*}
$$

From (B.7), (B.8), we get

$$
\begin{equation*}
\frac{\partial \psi}{\partial s}=B_{x} \frac{\partial \varphi}{\partial s}\left(1-B_{x} \frac{\partial \varphi}{\partial y_{1}}\right)^{-1} \tag{B.10}
\end{equation*}
$$

(where, as always, the functions must be taken at the appropriate arguments); then, from (B.9) we get

$$
\begin{equation*}
\frac{\partial \Phi}{\partial s}=A_{x} \frac{\partial \varphi}{\partial s}\left(1-B_{x} \frac{\partial \varphi}{\partial y_{1}}\right)^{-1} . \tag{B.11}
\end{equation*}
$$

We must now take the logarithmic derivative of the right-hand term with respect to $y_{0}$. It is the sum of three terms:

$$
\begin{equation*}
Z_{1}=\partial_{y} \log \left|A_{x}\right|+\partial_{x} \log \left|A_{x}\right| \frac{\partial \varphi}{\partial y_{1}} \frac{\partial \psi}{\partial y_{0}}, \tag{B.12}
\end{equation*}
$$

$$
\begin{equation*}
Z_{2}=\frac{\partial}{\partial y_{1}}\left(\log \frac{\partial \varphi}{\partial s}\right) \frac{\partial \psi}{\partial y_{0}}, \tag{B.13}
\end{equation*}
$$

$$
\begin{equation*}
Z_{3}=\left(1-B_{x} \frac{\partial \varphi}{\partial y_{1}}\right)^{-1}\left\{\frac{\partial^{2} \varphi}{\partial y_{1}^{2}} \frac{\partial \psi}{\partial y_{0}} B_{x}+\frac{\partial \varphi}{\partial y_{1}}\left(B_{y} \partial_{x} \log \left|B_{y}\right|+B_{x x} \frac{\partial \varphi}{\partial y_{1}} \frac{\partial \psi}{\partial y_{0}}\right)\right\} . \tag{B.14}
\end{equation*}
$$

In these formulas, we have from (B.7), (B.8)

$$
\begin{equation*}
\frac{\partial \psi}{\partial y_{0}}=B_{y}\left(1-B_{x} \frac{\partial \varphi}{\partial y_{1}}\right)^{-1} . \tag{B.15}
\end{equation*}
$$

We see that we obtain (B.6) with

$$
\begin{equation*}
T^{\prime} \leqslant a T+C \tag{B.16}
\end{equation*}
$$

Here, the constant $C$ is bounded in terms of $\left\|\left(1-B_{x} \frac{\partial \varphi}{\partial y_{1}}\right)^{-1}\right\|_{\infty}$, the distortion of $F$, the cone condition, $\left\|\frac{\partial \varphi}{\partial y_{1}}\right\|_{\infty}$ and $\left\|\frac{\partial^{2} \varphi}{\partial y_{1}^{2}}\right\|_{\infty}$. The constant $a$ is given by:

$$
\begin{equation*}
a=\left\|\frac{\partial \psi}{\partial y_{0}}\right\|_{\infty} \leqslant C\left\|B_{y}\right\|_{\infty} \tag{B.17}
\end{equation*}
$$

and, therefore, $a$ will be $<\frac{1}{2}$ if $|Q|=\left\|B_{y}\right\|_{\infty}$ is sufficiently small.
Remark. The control of $\frac{\partial \Phi}{\partial y_{0}}$ is guaranteed by the cone condition and the control of $\frac{\partial^{2} \Phi}{\partial y_{0}^{2}}$ by the distortion of $F$.

Thus, the mixed second derivative does not "explode" by taking pre-images by affine-like maps.
B. 3 The Parabolic Case. We now use the setting and notations of Subsection 3.5. We have intervals $I_{0}^{s}, I_{0}^{u}, I_{u}^{s}, I_{u}^{u}, I_{s}^{s}, I_{s}^{u}$ with respective coordinates $x_{0}, y_{0}, x_{u}, y_{u}, x_{s}, y_{s}$.

We have an affine-like map $F$ with domain $P \subset I_{0}^{s} \times I_{0}^{u}$, image $Q \subset I_{u}^{s} \times I_{u}^{u}$, and implicit representation $(A, B)$. We also have a folding map $G$ with domain $L_{u} \subset I_{u}^{s} \times I_{u}^{u}$, image $L_{s} \subset I_{s}^{s} \times I_{s}^{u}$. The map $G$ is implicitly defined by the system

$$
\left\{\begin{array}{l}
y_{s}=Y_{s}\left(w, x_{s}\right)  \tag{B.18}\\
x_{u}=X_{u}\left(w, y_{u}\right) \\
w^{2}=\theta\left(y_{u}, x_{s}\right)
\end{array}\right.
$$

with $Y_{s}, X_{u}, \theta$ as in Subsections 2.3 and 3.5. The affine-like map $F$ should satisfy (cf. (PC1) in Subsection 3.5, (R4) in Subsection 5.3):

$$
\begin{equation*}
\left|B_{x}\right|<b, \quad\left|B_{x x}\right|<b \tag{B.19}
\end{equation*}
$$

with $b \ll 1$. As above we are given a one-parameter family of vertical-like curves which are graphs

$$
\begin{equation*}
\omega(s)=\left\{x_{s}=\varphi\left(y_{s}, s\right)\right\} \tag{B.20}
\end{equation*}
$$

but these curves are now very close to being exactly vertical

$$
\begin{equation*}
\left|\frac{\partial \varphi}{\partial y_{s}}\right|<b, \quad\left|\frac{\partial^{2} \varphi}{\partial y_{s}^{2}}\right|<b \tag{B.21}
\end{equation*}
$$

We also assume as above that

$$
\begin{equation*}
\frac{\partial \varphi}{\partial s}\left(y_{s}, s\right)>0 \tag{B.22}
\end{equation*}
$$

and that we have some uniform bound

$$
\begin{equation*}
\left|\frac{\partial}{\partial s} \log \frac{\partial \varphi}{\partial s}\right| \leqslant T \tag{B.23}
\end{equation*}
$$

We then define the curves $\Omega^{ \pm}(s)$ as the connected components of $F^{-1}\left(Q \cap G^{-1}\left(\omega(s) \cap L_{s}\right)\right)$.

They should be graphs

$$
\begin{equation*}
\Omega^{ \pm}(s)=\left\{x_{0}=\Phi^{ \pm}\left(y_{0}, s\right)\right\} \tag{B.24}
\end{equation*}
$$

and we want some uniform control

$$
\begin{equation*}
\left|\frac{\partial}{\partial y_{0}} \log \right| \frac{\partial \Phi}{\partial s}\left|\mid \leqslant T^{\prime}\right. \tag{B.25}
\end{equation*}
$$

As in Subsection 3.5 (formulas (3.14), (3.15) we eliminate $y_{u}$ to write

$$
\begin{equation*}
x_{u}=X\left(w, y_{0}\right) \tag{B.26}
\end{equation*}
$$

Similarly, in the equation

$$
\begin{equation*}
y_{s}=Y_{s}\left(w, \varphi\left(y_{s}, s\right)\right), \tag{B.27}
\end{equation*}
$$

we solve for $y_{s}$ to write

$$
\begin{equation*}
y_{s}=Y(w, s) . \tag{B.28}
\end{equation*}
$$

We then set

$$
\begin{align*}
& \bar{X}(w, s) \quad:=\varphi(Y(w, s), s)  \tag{B.29}\\
& \bar{Y}\left(w, y_{0}\right) \quad:=B\left(y_{0}, X\left(w, y_{0}\right)\right), \\
& C\left(w, y_{0}, s\right):=w^{2}-\theta\left(\bar{Y}\left(w, y_{0}\right), \bar{X}(w, s)\right) .
\end{align*}
$$

Then, one should solve $C=0$ to get

$$
\begin{equation*}
w=W^{ \pm}\left(y_{0}, s\right), \tag{B.32}
\end{equation*}
$$

$$
\begin{equation*}
\Phi\left(y_{0}, s\right)=A\left(y_{0}, X\left(W\left(y_{0}, s\right), y_{0}\right)\right), \tag{B.33}
\end{equation*}
$$

with $W=W^{ \pm}, \Phi=\Phi^{ \pm}$.
The relations (B.26) through (B.33) are the equations which allow us to calculate $\frac{\partial}{\partial y_{0}} \log \left|\frac{\partial \Phi}{\partial s}\right|$.
From (B.33), we have

$$
\begin{equation*}
\frac{\partial \Phi}{\partial s}=A_{x} X_{w} \frac{\partial W}{\partial s} . \tag{B.34}
\end{equation*}
$$

From the implicit definition of $X$

$$
\begin{equation*}
X\left(w, y_{0}\right)=X_{u}\left(w, B_{0}\left(y_{0}, X\left(w, y_{0}\right)\right)\right) \tag{B.35}
\end{equation*}
$$

we get

$$
\begin{equation*}
X_{w}=\left(1-X_{u, y} B_{x}\right)^{-1} X_{u, w} . \tag{B.36}
\end{equation*}
$$

One has also

$$
\begin{align*}
& \frac{\partial W}{\partial s}=-\frac{\partial C}{\partial s} C_{w}^{-1},  \tag{B.37}\\
& \frac{\partial C}{\partial s}=-\theta_{x}\left(\frac{\partial \varphi}{\partial s}+\frac{\partial \varphi}{\partial y} \frac{\partial Y}{\partial s}\right) . \tag{B.38}
\end{align*}
$$

From (B.27), (B.28), one obtains

$$
\begin{equation*}
\frac{\partial Y}{\partial s}=Y_{s, x} \frac{\partial \varphi}{\partial s}\left(1-Y_{s, x} \frac{\partial \varphi}{\partial y}\right)^{-1}, \tag{B.39}
\end{equation*}
$$

and putting this in (B.38) leads to

$$
\begin{equation*}
\frac{\partial C}{\partial s}=-\theta_{x} \frac{\partial \varphi}{\partial s}\left(1-Y_{s, x} \frac{\partial \varphi}{\partial y}\right)^{-1} \tag{B.40}
\end{equation*}
$$

Therefore, introducing (B.36), (B.37), (B.40) into (B.34), we have

$$
\begin{equation*}
\frac{\partial \Phi}{\partial s}=A_{x} X_{u, w}\left(1-X_{u, y} B_{x}\right)^{-1} C_{w}^{-1} \theta_{x} \frac{\partial \varphi}{\partial s}\left(1-Y_{s, x} \frac{\partial \varphi}{\partial y}\right)^{-1} . \tag{B.41}
\end{equation*}
$$

We now take the logarithmic derivative of this product of seven terms with relation to $y_{0}$. We obtain a sum of seven terms:
(B.42) $\quad Z_{1}=\partial_{y} \log \left|A_{x}\right|+\partial_{x} \log \left|A_{x}\right|\left(X_{y}+X_{w} W_{y}\right)$,
(B.43) $\quad Z_{2}=\left(\partial_{w} \log \left|X_{u, w}\right|\right) W_{y}+\partial_{y} \log \left|X_{u, w}\right|\left(\bar{Y}_{y}+\bar{Y}_{w} W_{y}\right)$,

$$
\begin{equation*}
Z_{4}=C_{w}^{-1}\left[\theta_{x} \bar{X}_{w} W_{y}+\theta_{y}\left(\bar{Y}_{y}+\bar{Y}_{w} W_{y}\right)\right] \tag{B.45}
\end{equation*}
$$

(B.46) $\quad Z_{5}=\partial_{y} \log \left|\theta_{x}\right|\left(\bar{Y}_{y}+\bar{Y}_{w} W_{y}\right)+\partial_{x} \log \left|\theta_{x}\right| \bar{X}_{w} W_{y}$,

$$
\begin{align*}
Z_{3} & =\left(1-X_{u, y} B_{x}\right)^{-1}\left\{X_{u y}\left(B_{x y}+B_{x x}\left(X_{y}+X_{w} W_{y}\right)\right)\right.  \tag{B.44}\\
& \left.+B_{x}\left(X_{u, y w} W_{y}+X_{u, y y}\left(\bar{Y}_{y}+\bar{Y}_{w} W_{y}\right)\right)\right\}
\end{align*}
$$

$$
\begin{equation*}
Z_{6}=\left(\frac{\partial}{\partial y_{s}} \log \frac{\partial \varphi}{\partial s}\right) Y_{w} W_{y} \tag{B.47}
\end{equation*}
$$

The terms in these expressions that have not yet been introduced are:

$$
\begin{align*}
X_{y} & =X_{u, y} B_{y}\left(1-X_{u, y} B_{x}\right)^{-1},  \tag{B.49}\\
W_{y} & =-C_{y} C_{w}^{-1}=\theta_{y} \bar{Y}_{y} C_{w}^{-1}, \\
\bar{Y}_{y} & =B_{y}+B_{x} X_{y}, \\
\bar{Y}_{w} & =B_{x} X_{w}, \\
\bar{X}_{w} & =\frac{\partial \varphi}{\partial y_{s}} Y_{w}, \\
Y_{w} & =Y_{s, w}\left(1-Y_{s, x} \frac{\partial \varphi}{\partial y_{s}}\right)^{-1} .
\end{align*}
$$

We obtain the estimates
$\left|X_{w}\right| \leqslant C,\left|Y_{w}\right| \leqslant C$,
$\left|\bar{X}_{w}\right| \leqslant C b,\left|\bar{Y}_{w}\right| \leqslant C b$,
$\left|X_{y}\right| \leqslant C|Q|,\left|\bar{Y}_{y}\right| \leqslant C|Q|$,
$\left|W_{y}\right| \leqslant C|Q|\left|C_{w}^{-1}\right|$,

$$
\begin{equation*}
\left|Z_{i}\right| \leqslant C+C|Q|\left|C_{w}^{-1}\right| \text { for } i=1,2,3,5,7 ; \tag{B.58}
\end{equation*}
$$

$$
\begin{equation*}
\left|Z_{4}\right| \leqslant C|Q|\left|C_{w}^{-1}\right|+C b|Q|\left|C_{w}^{-1}\right|^{2} ; \tag{B.59}
\end{equation*}
$$

$$
\begin{equation*}
\left|Z_{6}\right| \leqslant C|Q|\left|C_{w}^{-1}\right| T \tag{B.60}
\end{equation*}
$$

Therefore, we obtain (B.25) with

$$
\begin{equation*}
T^{\prime}=C|Q|\left\|C_{w}^{-1}\right\|_{\infty} T+C+C|Q|\left\|C_{w}^{-1}\right\|_{\infty}+C b|Q|\left\|C_{w}^{-1}\right\|_{\infty}^{2} . \tag{B.62}
\end{equation*}
$$

Regarding $C_{w}^{-1}$, the same considerations as in Subsection 3.5 apply. One checks that $C_{w w}$ is close to two, hence the value of $\left|C_{w}^{-1}\right|$ at the argument where $C$ vanishes is of the order of $|\bar{C}|^{-\frac{1}{2}}$ where

$$
\begin{equation*}
\bar{C}\left(y_{0}, s\right)=\min _{w} C\left(w, y_{0}, s\right) . \tag{B.63}
\end{equation*}
$$

In the context of Section 10, we have

$$
\begin{equation*}
|Q| \ll|\bar{C}| \ll 1, \tag{B.64}
\end{equation*}
$$

and, therefore,

$$
\begin{equation*}
T^{\prime} \leqslant C|Q|^{\frac{1}{2}} T+C . \tag{B.65}
\end{equation*}
$$

This shows that $T^{\prime}$ does not "blow up".
B. 4 In the context of Proposition 40 in Subsection 10.5, Case 2 in the proof, we had an element $\left(P_{\ell}, Q_{\ell}, n_{\ell}\right) \in \mathcal{R}$ and two non-simple children $P_{\ell+1}, P_{\ell+1}^{\prime}$ of $P_{\ell \ell}$. Let $\omega_{\ell+1}, \omega_{\ell+1}^{\prime}$ be components of the vertical part of the boundary of $P_{\ell+1}, P_{\ell+1}^{\prime}$ respectively. We wanted to have (cf. (10.65))

$$
\begin{equation*}
\left|\log \frac{\varphi_{\ell+1}(y)-\varphi_{\ell+1}^{\prime}(y)}{\varphi_{\ell+1}\left(y^{\prime}\right)-\varphi_{\ell+1}^{\prime}\left(y^{\prime}\right)}\right| \leqslant C\left|y-y^{\prime}\right| \tag{B.66}
\end{equation*}
$$

where $\omega_{\ell+1}=\left\{x=\varphi_{\ell+1}(y)\right\}, \omega_{\ell+1}^{\prime}=\left\{x=\varphi_{\ell+1}^{\prime}(y)\right\}$.
To prove (10.66), we will imbed both $\omega_{\ell+1}$ and $\omega_{\ell+1}^{\prime}$ in a one-parameter family of curves

$$
\begin{equation*}
\Omega(s)=\{x=\Phi(y, s)\} \tag{B.67}
\end{equation*}
$$

with

$$
\begin{equation*}
\Omega\left(s_{0}\right)=\omega_{\ell+1}, \quad \Omega\left(s_{1}\right)=\omega_{\ell+1}^{\prime} \tag{B.68}
\end{equation*}
$$

$$
\frac{\partial}{\partial s} \Phi>0
$$

$$
\begin{equation*}
\left|\frac{\partial}{\partial y} \log \frac{\partial \Phi}{\partial s}\right| \leqslant C . \tag{B.70}
\end{equation*}
$$

Indeed, in this case, we can write

$$
\begin{equation*}
\frac{\varphi_{\ell+1}(y)-\varphi_{\ell+1}^{\prime}(y)}{\varphi_{\ell+1}\left(y^{\prime}\right)-\varphi_{\ell+1}^{\prime}\left(y^{\prime}\right)}=\frac{\int_{s_{0}}^{s_{1}} \frac{\partial \Phi}{\partial s}(y, s) d s}{\int_{s_{0}}^{s_{1}} \frac{\partial \Phi}{\partial s}\left(y^{\prime}, s\right) d s}, \tag{B.71}
\end{equation*}
$$

with, from (B.70), for any $s$

$$
\begin{equation*}
e^{-C\left|y-y^{\prime}\right|} \leqslant\left(\frac{\partial \Phi}{\partial s}(y, s)\right)^{-1} \frac{\partial \Phi}{\partial s}\left(y^{\prime}, s\right) \leqslant e^{C\left|y-y^{\prime}\right|}, \tag{B.72}
\end{equation*}
$$

which yields (B.66).

## Appendix C

## A Toy Model for the Transversality Relation

C. 1 Our goal in this appendix is to explain why the complicated definition of the transversality relation in Subsection 5.4, is in some way "natural", if we require some useful properties for the proof of our Main Theorem. The toy model that we are considering is an abstract one. It is much simpler than the real situation of Section 5 because the sets in which the relation takes place are well defined to begin with: in Section 5 , we need to know the transversality relation in order to construct the classes $\mathcal{R}(I)$.
C. 2 A partially ordered set $X$ is a forest if, for any $x_{0} \in X$, the set $\left\{x \geqslant x_{0}\right\}$ is finite and totally ordered. A tree is a forest with a single maximal element. Let $X_{1}, \cdots, X_{n}$ be forests and let $A$ be a subset of $X=X_{1} \times \cdots \times X_{n}$ (one should think of $A$ as the graph of an $n$-ary relation). We say that $A$ is hereditary if whenever $x=\left(x_{1}, \cdots, x_{n}\right), y=\left(y_{1}, \cdots, y_{n}\right)$ are such that and $y_{i} \leqslant x_{i}$ for all $1 \leqslant i \leqslant n$ (abbreviated as $y \leqslant x$ ), then $y \in A$ if $x \in A$.

Two points $x=\left(x_{1}, \cdots, x_{n}\right), y=\left(y_{1}, \cdots, y_{n}\right)$ of $X$ are coordinate-wise comparable (c-comparable for short) if for each $i \in\{1, \cdots, n\}$, we have $x_{i} \geqslant y_{i}$ or $x_{i} \leqslant y_{i}$. In this case, we set

$$
\begin{equation*}
x \vee y=\left(\max \left(x_{i}, y_{i}\right)\right)_{1 \leqslant i \leqslant n} \tag{C.1}
\end{equation*}
$$

The set $A$ is concave if, whenever $x, y \in A$ are $c$-comparable, then the point $x \vee y$ also belongs to A.

The intersection of hereditary, resp. concave, subsets of $X$ is hereditary, resp. concave. It follows that any subset $A \subset X$ is contained in a smallest concave hereditary subset, called the $c$.h-envelope and denoted by $\widehat{A}$.

Example. When the number of factors $n=1$, any subset is concave: the $c$.h-envelope of $A \subset X_{1}$ is the set of $x \in X_{1}$ such that $x \leqslant y$ for some $y \in A$.
C. 3 We construct the $c . h$-envelope when $n=2$.

Proposition. Let $X_{1}, X_{2}$ be forests and $A$ be a subset of $X=X_{1} \times X_{2}$. Let $A_{1}$ be the set of $x \in X$ such that $x=y \vee z$ for some $c$-comparable $y, z \in A$. The c.h-envelope of $A$ is equal to the set $A_{2}$ of $t=\left(t_{1}, t_{2}\right)$ such that $t_{1} \leqslant x_{1}, t_{2} \leqslant x_{2}$ for some $x=\left(x_{1}, x_{2}\right) \in A_{1}$.

Proof. It is clear that the set $A_{2}$ defined in the proposition is hereditary and it is contained in the $c . h$-envelope $\widehat{A}$ of $A$. We have to prove that $A_{2}$ is concave. We first prove the

Lemma 18. If $y, z \in A_{1}$ are $c$-comparable, $y \vee z$ also belongs to $A_{1}$.

Proof. By the definition of $A_{1}$, we can write $y=y^{\prime} \vee y^{\prime \prime}, z=z^{\prime} \vee z^{\prime \prime}$ with $y^{\prime}, y^{\prime \prime}, z^{\prime}, z^{\prime \prime}$ in $A, y^{\prime}$ and $y^{\prime \prime} c$-comparable, $z^{\prime}$ and $z^{\prime \prime} c$-comparable. We may assume that $y_{1} \leqslant z_{1}$ and $y_{2} \geqslant z_{2}$, and also $z_{1}=z_{1}^{\prime}, y_{2}=y_{2}^{\prime}$; then, we have $z_{1}^{\prime}=z_{1} \geqslant y_{1} \geqslant y_{1}^{\prime}$ and $y_{2}^{\prime}=y_{2} \geqslant z_{2} \geqslant z_{2}^{\prime}$, hence $y^{\prime}, z^{\prime}$ are $c$-comparable with $y^{\prime} \vee z^{\prime}=y \vee z$.

End of the Proof of the Proposition. Let $t^{\prime}, t^{\prime \prime} \in A_{2}$ be $c$-comparable and let $x^{\prime}, x^{\prime \prime} \in A_{1}$ be such that $x_{i}^{\prime} \geqslant t_{i}^{\prime}, x_{i}^{\prime \prime} \geqslant t_{i}^{\prime \prime}$ for $i=1,2$. As $X_{1}$ and $X_{2}$ are forests, $x^{\prime}$ and $x^{\prime \prime}$ are $c$-comparable. From the lemma, $x^{\prime} \vee x^{\prime \prime}$ belongs to $A_{1}$; then $t^{\prime} \vee t^{\prime \prime}$ belongs to $A_{2}$.
C. 4 For $n \geqslant 3$, the situation is more complicated, as the two examples below indicate.

Example 1. Let $X_{1}, X_{2}, X_{3}$ be forests and let $x, y, z$ be three points of $X=X_{1} \times X_{2} \times X_{3}$ such that

$$
\begin{align*}
& x_{1} \geqslant y_{1} \geqslant z_{1}  \tag{C.2}\\
& y_{2} \geqslant x_{2}, \quad y_{2} \geqslant z_{2}  \tag{C.3}\\
& z_{3} \geqslant x_{3}, \quad z_{3} \geqslant y_{3} \tag{C.4}
\end{align*}
$$

Let $A=\{x, y, z\} \subset X$. If we define, as in the proposition above,

$$
\begin{equation*}
A_{1}=\{u \vee v, u, v \in A, u, v \quad c \text {-comparable }\} \tag{C.5}
\end{equation*}
$$

and if we assume that $x_{2}, z_{2}$ are not comparable and $x_{3}, y_{3}$ are not comparable, then we have

$$
\begin{equation*}
A_{1}=\left\{x, y, z, \quad y \vee z=\left(y_{1}, y_{2}, z_{3}\right)\right\} \tag{C.6}
\end{equation*}
$$

On the other hand, the point $w=\left(x_{1}, y_{2}, z_{3}\right)=x \vee(y \vee z)$ certainly belongs to the $c$. $h$-envelope of $A$, but does not satisfy $w_{i} \leqslant u_{i}(i=1,2,3)$ for any $u \in A_{1}$. This example shows that the analogue of the proposition above is false for $n=3$.

Example 2. Let $X_{1}, X_{2}, X_{3}$ forests and let $x, y, z \in X=X_{1} \times X_{2} \times X_{3}$ such that

$$
\begin{array}{ll}
x_{1} \geqslant y_{1}, & x_{1} \geqslant z_{1} \\
y_{2} \geqslant x_{2}, & y_{2} \geqslant z_{2} \\
z_{3} \geqslant x_{3}, & z_{3} \geqslant y_{3} \tag{C.9}
\end{array}
$$

but none of the pairs $\left(y_{1}, z_{1}\right),\left(x_{2}, z_{2}\right),\left(x_{3}, y_{3}\right)$ is made of comparable elements. Let $A=\{x, y, z\}$. The sets $\{u \leqslant x\},\{v \leqslant y\},\{w \leqslant z\}$ are disjoint and their union is the $c . h$-envelope of $A$ : any $u \leqslant x, v \leqslant y$ cannot be $c$-comparable; otherwise, as $X_{3}$ is a forest and $x, y$ are larger than $u \wedge v=\left(\min \left(u_{i}, v_{i}\right)\right), x_{3}$ and $y_{3}$ would be comparable. On the other hand, if $u^{\prime} \leqslant x, u^{\prime \prime} \leqslant x$ and $u^{\prime}, u^{\prime \prime} c$-comparable, then $u^{\prime} \vee u^{\prime \prime} \leqslant x$.
C. 5 We have the following partial result:

Proposition. Let $X_{1}, \cdots, X_{n}$ be forests and let $A$ be a subset of $X=X_{1} \times \cdots \times X_{n}$. Let $A_{1}$ be the set of elements $x \in X$ for which there exists $x^{1}, x^{2}, \cdots, x^{n}$ in $A$ with $x_{j}=x_{j}^{j} \geqslant x_{j}^{i}$ for all $1 \leqslant i, j \leqslant n$. Let $A_{2}$ be the set of elements $y \in X$ such that $y \leqslant x$ for some $x$ in $A_{1}$. Then $A_{2}$ contains the $c$.h-envelope of $A$.

Remark. Example 2 above shows that $A_{2}$ can be strictly larger than the c.h-envelope. Example 1 shows that the straightforward generalization of the case $n=2$ does not work.

Proof of the Proposition. It is very similar to the proof of the proposition in C. 3 above and left to the reader.
C. 6 We will now see how the definition of the transversality relation in Subsection 5.4 is a natural consequence of the proposition above. As observed earlier, an essential difference with the toy model is that the transversality relation is used to construct the classes $\mathcal{R}(I)$. So, let us just try to define the relation for the starting class $\mathcal{R}\left(I_{0}\right)$ associated to the initial horseshoe $K$. We would have:

$$
\begin{align*}
& X_{1}=\left\{(P, Q, n) \in \mathcal{R}\left(I_{0}\right), Q \subset Q_{u}\right\},  \tag{C.10}\\
& X_{2}=\left\{\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in \mathcal{R}\left(I_{0}\right), \quad P \subset P_{s}\right\}, \tag{C.11}
\end{align*}
$$

and $X_{3}$ is the set of parameter intervals. All sets are partially ordered by inclusion (of the $Q$ 's for $X_{1}$, of the $P$ 's for $X_{2}$ ), and are obviously trees, with respective roots $\left(P_{u}, Q_{u}, n_{u}\right),\left(P_{s}, Q_{s}, n_{s}\right), I_{0}$.

We start from an intuitive definition of transversality: for $(P, Q, n) \in X_{1},\left(P^{\prime}, Q^{\prime}, n^{\prime}\right) \in X_{2}, I \in X_{3}$, we write

$$
Q \widehat{\pitchfork}_{I} P^{\prime}
$$

if for all $t \in I$ we have

$$
\begin{equation*}
\delta\left(Q, P^{\prime}\right) \geqslant 2 \max \left(I,|Q|^{1-\eta},\left|P^{\prime}\right|^{1-\eta}\right) . \tag{C.12}
\end{equation*}
$$

(The number $\eta$ in the exponent is necessary in order to keep the distortions under control.)
The corresponding subset of $X_{1} \times X_{2} \times X_{3}$ is

$$
\begin{equation*}
A=\left\{\left(Q, P^{\prime}, I\right), \quad Q \widehat{币}_{I} P^{\prime}\right\} . \tag{C.13}
\end{equation*}
$$

This set is hereditary but it is not, a priori, concave. The concavity property (Propositions 4 and 7 in Section 6) is very useful in many places. So, we wish to replace $A$ by a larger set which is hereditary and concave. If we apply the recipe of the proposition in C.5, we are led first to define
a set $A_{1}$ and then a set $A_{2}$. According to the proposition, $A_{1}$ should be the set of $\left(Q, P^{\prime}, I\right)$ for which there exist $Q_{2}, Q_{3} \subset Q, P_{1}^{\prime}, P_{3}^{\prime} \subset P^{\prime}, I_{1}, I_{2} \subset I$ satisfying

$$
\begin{equation*}
Q \widehat{\pitchfork}_{I_{1}} P_{1}^{\prime}, \quad Q_{2} \widehat{\pitchfork}_{I_{2}} P^{\prime}, \quad Q_{3} \widehat{\pitchfork}_{I} P_{3}^{\prime} \tag{C.14}
\end{equation*}
$$

As $I_{1}, I_{2}$ can be chosen arbitrarily small, we can replace them with single values $t_{1}, t_{2} \in I$ of the parameter; the three conditions in (C.14) become:

- there exists $P_{1}^{\prime} \subset P^{\prime}, t_{1} \in I$ such that

$$
\begin{equation*}
\delta\left(Q, P_{1}^{\prime}\right) \geqslant 2 \max \left(|Q|^{1-\eta},\left|P_{1}^{\prime}\right|^{1-\eta}\right) \tag{C.15}
\end{equation*}
$$

- there exists $Q_{2} \subset Q, t_{2} \in I$ such that

$$
\begin{equation*}
\delta\left(Q_{2}, P^{\prime}\right) \geqslant 2 \max \left(\left|Q_{2}\right|^{1-\eta},\left|P^{\prime}\right|^{1-\eta}\right) \tag{C.16}
\end{equation*}
$$

- there exists $Q_{3} \subset Q, P_{3}^{\prime} \subset P^{\prime}$ such that

$$
\begin{equation*}
\delta\left(Q_{3}, P_{3}^{\prime}\right) \geqslant 2 \max \left(|I|,\left|Q_{3}\right|^{1-\eta},\left|P_{3}^{\prime}\right|^{1-\eta}\right) \tag{C.17}
\end{equation*}
$$

for all $t \in I$.
As $P_{1}^{\prime}, Q_{2}, Q_{3}, P_{3}^{\prime}$ may be chosen arbitrarily thin, it is natural to replace (C.15), (C.16), (C.17) by

$$
\begin{align*}
& \delta\left(Q, P_{1}^{\prime}\right) \geqslant 2|Q|^{1-\eta} \text { for } t=t_{1} ;  \tag{C.15}\\
& \delta\left(Q_{2}, P^{\prime}\right) \geqslant 2\left|P^{\prime}\right|^{1-\eta} \text { for } t=t_{2} ; \\
& \delta\left(Q_{3}, P_{3}^{\prime}\right) \geqslant 2|I| \text { for all } t \in I .
\end{align*}
$$

Finally, the largest $t$ value of $\delta\left(Q, P_{1}^{\prime}\right)$ that one can hope for (by choosing $P_{1}^{\prime} \subset P^{\prime}$ appropriately) is $\delta_{R}\left(Q, P^{\prime}\right)$; similarly, the largest value of $\delta\left(Q_{2}, P^{\prime}\right)$ that one can hope for is $\delta_{L}\left(Q, P^{\prime}\right)$ and the largest value of $\delta\left(Q_{3}, P_{3}^{\prime}\right)$ that one can hope for is $\delta_{L R}\left(Q, P^{\prime}\right)$. Notice that we need anyway to eliminate $P_{1}^{\prime}, Q_{2}, P_{3}^{\prime}, Q_{3}$ from the definition because in $\mathcal{R}(I)$ (instead of $\mathcal{R}\left(I_{0}\right)$ ), elements are constructed inductively and thinner rectangles are constructed at the end. Replacing $\delta\left(Q, P_{1}^{\prime}\right)$ by $\delta_{R}\left(Q, P^{\prime}\right)$, $\delta\left(Q_{2}, P^{\prime}\right)$ by $\delta_{L}\left(Q, P^{\prime}\right)$ and $\delta\left(Q_{3}, P_{3}^{\prime}\right)$ by $\delta_{L R}\left(Q, P^{\prime}\right)$, we obtain the three conditions, (T1), (T2),(T3) in Subsection 5.4. This defines $\pitchfork$.

The last step is to go from $A_{1}$ to $A_{2}$, taking the hereditary envelope of $A_{1}$, which corresponds exactly to the transition from $\bar{\hbar}$ to $\pitchfork$ in Subsection 5.4.

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