# KANTOROVICH POTENTIALS AND CONTINUITY OF TOTAL COST FOR RELATIVISTIC COST FUNCTIONS 

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#### Abstract

In this paper we consider the optimal mass transport problem for relativistic cost functions, introduced in [12] as a generalization of the relativistic heat cost. A typical example of such a cost function is $c_{t}(x, y)=h\left(\frac{y-x}{t}\right), h$ being a strictly convex function when the variable lies on a given ball, and infinite otherwise. It has been already proved that, for every $t$ larger than some critical time $T>0$, existence and uniqueness of optimal maps hold; nonetheless, the existence of a Kantorovich potential is known only under quite restrictive assumptions. Moreover, the total cost corresponding to time $t$ has been only proved to be a decreasing rightcontinuous function of $t$. In this paper, we extend the existence of Kantorovich potentials to a much broader setting, and we show that the total cost is a continuous function. To obtain both results the two main crucial steps are a refined "chain lemma" and the result that, for $t>T$, the points moving at maximal distance are negligible for the optimal plan.


## 1. Introduction

In this paper, we consider the classical mass transport problem with a particular class of cost functions $c$. While the case when $c$ is strictly convex and real-valued is well understood (we refer to [34] for a survey), our goal is to continue the study of a wide class of non real-valued cost functions, called "relativistic costs", which were introduced in [12] as a generalization to the "relativistic heat cost".

More precisely, the relativistic heat cost, introduced by Brenier in [14], is defined as $c(x, y)=$ $h(y-x)$, where $h$ is given by the formula

$$
h(z)= \begin{cases}1-\sqrt{1-|z|^{2}} & |z| \leq 1  \tag{1.1}\\ +\infty & |z|>1\end{cases}
$$

As explained by Brenier in the aforementioned paper, this cost function can be used in order to study a relativistic heat equation, where the region where the cost is infinite comes from the fact that the heat has finite speed propagation. Afterwards, the corresponding relativistic heat equation has also been studied by McCann and Puel in [28] by means of the Jordan-KinderlehrerOtto approach introduced in [25], and by Andreu, Caselles and Mazón via PDE methods in a series of papers $[3,5,4,6,7,8,9,16]$. We refer to $[14,12]$ and the references therein for more on this topic.

Notice that, from the mathematical point of view, a quite interesting feature of this cost function is that it is strictly convex and bounded on its domain, hence in particular it is not continuous on $\mathbb{R}^{n} \times \mathbb{R}^{n}$ (while the non real-valued cost functions have been often considered as continuous, see for instance $[23,22,29,30,33,11])$.

More in general, one can consider costs "of relativistic type", meaning that they are strictly convex and bounded in a strictly convex domain $\mathscr{C}$ and $+\infty$ outside. The corresponding optimal mass transport problem has been studied in [12], see also [24].

In the classical case (strictly convex and real-valued cost function), a lot can be said; in particular, an optimal transport map exists, is unique, and it is induced by a Kantorovich potential $\varphi$ (for instance the optimal transport map is simply the gradient of $\varphi$ when the cost is the Euclidean norm squared, see $[27,13]$ ). This potential is also related to the so-called Kantorovich's dual problem. For non-negative and lower semi-continuous cost functions, the following duality result is well known (see [21]).

Theorem 1.1 (Kantorovich duality). Let $\mu$ and $\nu$ be two probability measures with finite second order moment on $\mathbb{R}^{n}$. Then, the following equality holds:

$$
\sup _{\mathcal{A}}\left\{\int_{\mathbb{R}^{n}} \varphi(x) d \mu(x)+\int_{\mathbb{R}^{n}} \psi(y) d \nu(y)\right\}=\min _{\gamma \in \Pi(\mu, \nu)} \int_{\mathbb{R}^{n} \times \mathbb{R}^{n}} c(x, y) d \gamma(x, y)<+\infty .
$$

Here and in the following, $\Pi(\mu, \nu)$ denotes the set of positive measures on $\mathbb{R}^{n} \times \mathbb{R}^{n}$ whose first (resp., second) marginal is $\mu$ (resp., $\nu$ ), while $\mathcal{A}$ is the set of pairs $(\varphi, \psi)$ of Lipschitz functions defined on $\mathbb{R}^{n}$ that satisfy $\varphi(x)+\psi(y) \leq c(x, y)$ for all $x, y \in \mathbb{R}^{n}$.

However, it should be emphasized that in general, Kantorovich's dual problem (on the left hand side) has no solutions, counterexamples can be found for instance in [10]. A sufficient condition for maximisers to exist is that the cost function is a bounded and uniformly continuous function, see [34]; a more general criterion for real-valued cost functions is given in [2]. In the case of non real-valued cost functions, the existence of maximisers is proved for reflector-type problems [23, 22] and for Alexandrov's Gauss curvature prescription problem [29, 11]; in both cases the cost function is a continuous non real-valued function on the unit sphere in Euclidean space. Existence of maximisers for the dual problem has also been proved in the case of the Wiener space for the quadratic cost [20].

One of the main results of the present paper is the existence of a Kantorovich potential for the mass transport problem relative to a relativistic cost. Beside a usual hypothesis on the initial measure, the main assumption is that of a "supercritical regime" which means, loosely speaking, that the mass transport problem is already well-posed even if the set of admissible motions is shrunk by a homothety of ratio $1-\varepsilon$. The precise statement is in Theorem B.

To conclude this introductory part, we recall that for real-valued cost functions, the enhanced formula involving a potential is the starting point of the study of the optimal map regularity. This topic, initiated by Caffarelli [15], has gone under tremendous development, we refer to [34, Chapter 12] for more on the subject. The question about the regularity of the optimal map for the relativistic heat cost in a very general setting (where only part of the mass can be moved at finite cost) is highlighted in [14].
1.1. Previous results. We briefly recall here some results already known, to enlighted the novelty of the results of the present paper. To do so, we need to introduce a parameter $t$, called the "speed", which takes into account the relativistic behaviour of the cost functions we
consider. To be precise, whenever $h: \mathbb{R}^{n} \rightarrow[0,+\infty]$ is a strictly convex function, bounded in some strictly convex set $\mathscr{C}$ and $+\infty$ outside, then for every $t>0$ we call

$$
c_{t}(x, y)=h\left(\frac{y-x}{t}\right)
$$

a relativistic cost function (see the formal Definition 3.1). Now, given two probability measures $\mu$ and $\nu$ with compact support, we can see what happens when the speed $t$ varies: to understand it, let us define the total cost as

$$
\begin{equation*}
\mathcal{C}(t)=\min _{\pi \in \Pi(\mu, \nu)} \mathcal{C}_{t}(\gamma)=\min _{\pi \in \Pi(\mu, \nu)} \iint_{\mathbb{R}^{2 n}} c_{t}(x, y) d \gamma(x, y) . \tag{1.2}
\end{equation*}
$$

It is easy to guess that the total cost is infinite when $t$ is small (at least if $\mu \neq \nu$ ), while it becomes finite when $t$ is larger, and finally it does not depend on $t$ if it is sufficiently big. The following properties are proved in [12].

Proposition 1.2. Given two probability measures $\mu \neq \nu$ with compact support and a relativistic cost function, the following holds.
a) The function $t \mapsto \mathcal{C}(t)$ is decreasing.
b) The speed $T:=\inf \{t ; \mathcal{C}(t)<+\infty\}$ is positive and finite ( $T$ is called the critical speed).
c) The total $\operatorname{cost} \mathcal{C}(T)$ is finite.
d) The function $\mathcal{C}(t)$ is right continuous.

Moreover, one can prove the existence of an optimal transport map for supercritical speed.
Theorem 1.3. Let $\mu$ and $\nu$ be two probability measures with compact support on $\mathbb{R}^{n}$ and $c_{t}$ be a relativistic cost. Assume that $\mu$ is absolutely continuous with respect to the Lebesgue measure and that $t \geq T$. Then, there exists a unique optimal transport plan $\gamma_{t}$ for the cost $c_{t}$, and this plan is induced by a map $F_{t}$.
1.2. Statement of the main results. Let us briefly describe the main achievements of this paper. First of all, we can observe that the total cost is a continuous function of $t$ when the cost is "highly relativistic": basically, a relativistic cost function is called "highly relativistic" when the slope of $h$ explodes at the boundary $\partial \mathscr{C}$, as it happens for the original relativistic heat cost (1.1) and most of the important examples (see the formal Definition 3.2).

Theorem A (Continuity of the total cost). Let $\mu$ and $\nu$ be two probability measures with compact support on $\mathbb{R}^{n}$, $c_{t}$ be a highly relativistic cost function, and assume that $\mu \in L^{\infty}\left(\mathbb{R}^{n}\right)$. Then, the total cost function $t \mapsto \mathcal{C}(t)$ is continuous on $[T,+\infty)$. In particular, for every $\rho \ll 1$ and every $t$, there exists $\bar{C}=\bar{C}(n, \mu, \nu, t, \rho)$ such that, if $\delta \ll 1$,

$$
\begin{equation*}
\mathcal{C}(t) \leq \mathcal{C}(t-\delta t) \leq \mathcal{C}(t)(1+\operatorname{exc}(\delta))+\frac{\bar{C}}{\kappa_{\delta}(\rho)} \tag{1.3}
\end{equation*}
$$

where $\operatorname{exc}(\delta)$ and $\kappa_{\delta}(\rho)$ converge to 0 and $+\infty$ respectively when $\delta \rightarrow 0$.
The precise values of the constants $\operatorname{exc}(\delta)$ and $\kappa_{\delta}(\rho)$ are introduced in Definition 4.1 and in (3.11) respectively. However, knowing that $\operatorname{exc}(\delta) \rightarrow 0$ and $\kappa_{\delta}(\rho) \rightarrow \infty$ when $\delta \rightarrow 0$ is enough
to infer the continuity of the total cost from (1.3). Of course, Theorem A extends the plain right continuity stated in Proposition 1.2.

Our second main result is the existence of a Kantorovich potential.
Theorem B (Existence of a Kantorovich potential). Let $\mu$ and $\nu$ be two probability measures with compact support on $\mathbb{R}^{n}$, $c_{t}$ be a highly relativistic cost function, and assume that $\mu$ is absolutely continuous with respect to the Lebesgue measure and has a connected support. Then, for any supercritical speed $t>T$ there exists a Kantorovich potential $\varphi_{t}$. In particular, there is a unique optimal transport plan, and it corresponds to an optimal transport map $F_{t}$, defined $\mu$-almost everywhere as

$$
F_{t}(x)=x+t \nabla h^{*}\left(-t \widetilde{\nabla} \varphi_{t}(x)\right),
$$

where $h^{*}$ denotes the Legendre transform of $h$ and $\widetilde{\nabla} \varphi_{t}$ is the approximate gradient of $\varphi_{t}$.
The classical notions of Kantorovich potential and of approximate gradient are given in Definitions 5.3 and 5.4. Let us recall that the existence of a Kantorovich potential was already proved in [12], but only under a regularity assumption on $\mu$, and only for almost every $t>T$.

We underline that all our proofs are self-contained, in particular they do not rely on Theorem 1.3.

There are two ingredients which are crucial for the proof of both our main theorems. The first one is the "finite chain Lemma" 2.10, which generalizes similar results already used in the literature, and which basically uses the cyclical monotonicity in order to modify a transport plan to get a competitor. We will also use a discrete version of this finite chain result, given in Lemma 5.5.

The second ingredient can be stated as an independent interesting result. It says that, for any supercritical time $t>T$, the optimal transport plan moves no point at "maximal distance", that is, the pairs $(x, y)$ such that $y-x$ belongs to the boundary of $t \mathscr{C}$ are $\gamma$-negligible.

Theorem C (No points move at maximal distance). Let $\mu$ and $\nu$ be two probability measures, $c_{t}$ a highly relativistic cost, $\mu \ll \mathscr{L}^{n}, t>T$ and $\gamma$ an optimal plan with respect to $c_{t}$. Then

$$
\gamma\left(\left\{(x, y) \in \mathbb{R}^{n} \times \mathbb{R}^{n}: y-x \in t \partial \mathscr{C}\right\}\right)=0
$$

In all our results above we used the assumption that the cost is highly relativistic: actually, we can provide a counterexample showing that for most of our claims the sole relativistic assumption is not enough.

The paper is organised as follows. In Section 2 we prove the finite chain Lemma 2.10. In Section 3 we prove Theorem C. In Section 4 we prove Theorem A, and we present an example which works as counterexample for most of our claims without the highly relativistic assumption. Finally, in Section 5 we prove Theorem B.

## 2. Preliminaries and the finite chain lemma

The goal of this section is to show the "finite chain Lemma" 2.10. We start with a section containing most of our technical definitions and preliminary results.
2.1. General definitions and introductory lemmas. Throughout this paper, $X$ and $Y$ will be arbitrary Polish spaces. We will denote by $\mathcal{M}^{+}(X)$ the space of positive Borel measures, and by $\mathcal{P}(X)$ the subspace of probability measures; usually, $\mu$ and $\nu$ will be two measures in $\mathcal{M}^{+}(X)$ and $\mathcal{M}^{+}(Y)$ respectively, having the same total mass, i.e. $\|\mu\|_{\mathcal{M}^{+}(X)}=\|\nu\|_{\mathcal{M}^{+}(Y)}$. We will denote by $\Pi(\mu, \nu)$ the space of transport plans between $\mu$ and $\nu$, that is, an element of $\Pi(\mu, \nu)$ is some measure $\gamma \in \mathcal{M}^{+}(X \times Y)$ whose marginals are $\mu$ and $\nu$. Given a map $g: X \rightarrow Y$, we will denote by $g_{\#}: \mathcal{M}^{+}(X) \rightarrow \mathcal{M}^{+}(Y)$ the push-forward operator, defined by

$$
\int_{Y} \varphi(y) d g_{\#} \mu(y):=\int_{X} \varphi(g(x)) d \mu(x) \quad \forall \varphi \in \mathrm{C}_{b}(Y)
$$

The projections of $X \times Y$ onto $X$ and $Y$ will be denoted as $\pi_{1}$ and $\pi_{2}$; for simplicity of notations, the corresponding projections from $\mathcal{M}^{+}(X \times Y)$ onto $\mathcal{M}^{+}(X)$ and $\mathcal{M}^{+}(Y)$ are also denoted by $\pi_{1}$ and $\pi_{2}$, instead of $\pi_{1 \#}$ and $\pi_{2 \#}$. In particular, $\gamma \in \Pi(\mu, \nu)$ means $\pi_{1} \gamma=\mu, \pi_{2} \gamma=\nu$. When $\alpha$ and $\mu$ are two measures such that $\alpha \ll \mu$, we denote by $\mathrm{d}_{\alpha}$ the density of $\alpha$ with respect to $\mu$, that is, $\alpha=\mathrm{d}_{\alpha} \mu$.

Definition 2.1. For any two measures $\mu_{1}, \mu_{2} \in \mathcal{M}^{+}(X)$, we define $\mu_{1} \wedge \mu_{2}$ as

$$
\mu_{1} \wedge \mu_{2}(B):=\inf \left\{\mu_{1}\left(B_{1}\right)+\mu_{2}\left(B_{2}\right): B_{1} \cap B_{2}=\emptyset, B_{1} \cup B_{2}=B\right\}
$$

where $B, B_{1}$ and $B_{2}$ are Borel subsets of $X$. Notice that, if $\mu_{1} \leq \mu$ and $\mu_{2} \leq \mu$ for some $\mu \in \mathcal{M}^{+}(X)$, then also $\mu_{1} \wedge \mu_{2} \leq \mu$. In particular, one has

$$
\begin{equation*}
\mu_{1} \wedge \mu_{2}=\min \left\{\mathrm{d}_{1}, \mathrm{~d}_{2}\right\} \mu \tag{2.1}
\end{equation*}
$$

Definition 2.2 ( $\alpha \leq \mu$ and its relative transport plan). Given $\alpha \ll \mu \in \mathcal{M}^{+}(X)$, we write that $\alpha \leq \mu$ if $\mathrm{d}_{\alpha} \leq 1$, and we will denote by $\mathcal{M}_{\mu}^{+}(X)$ the set of the Borel measures $\alpha \in \mathcal{M}^{+}(X)$ such that $\alpha \leq \mu$. If $\alpha \in \mathcal{M}_{\mu}^{+}(X)$, for any transport plan $\gamma \in \Pi(\mu, \nu)$ we denote by $\mathrm{d}_{\alpha} \gamma$ the transport plan defined by

$$
\int_{X \times Y} \varphi(x, y) d \mathrm{~d}_{\alpha} \gamma(x, y)=\iint_{X \times Y} \varphi(x, y) \mathrm{d}_{\alpha}(x) d \gamma(x, y) \quad \forall \varphi \in \mathrm{C}_{b}(X \times Y)
$$

Notice that the first marginal of $\mathrm{d}_{\alpha} \gamma$ is precisely $\alpha$. Similarly, we define $\mathrm{d}^{\beta} \gamma$ if $\beta \leq \nu$.
Roughly speaking, when $\gamma$ is a transport plan between $\mu$ and $\nu$, for any $\alpha \leq \mu$ the relative transport plan $\mathrm{d}_{\alpha} \gamma$ is the "portion" of the transport plan $\gamma$ relative to $\alpha$, that is, a smaller transport plan which transports the portion $\alpha$ of $\mu$ onto some portion of $\nu$. With the next definition, we "keep track" of what is precisely the latter portion of $\nu$.

Definition 2.3. For any transport plan $\gamma \in \Pi(\mu, \nu)$ we define $\vec{\Phi}: \mathcal{M}_{\mu}^{+}(X) \rightarrow \mathcal{M}_{\nu}^{+}(Y)$ and $\overleftarrow{\Phi}: \mathcal{M}_{\nu}^{+}(Y) \rightarrow \mathcal{M}_{\mu}^{+}(X)$ as

$$
\vec{\Phi}(\alpha):=\pi_{2}\left(\mathrm{~d}_{\alpha} \gamma\right), \quad \overleftarrow{\Phi}(\beta):=\pi_{1}\left(\mathrm{~d}^{\beta} \gamma\right)
$$

Through the paper, we will make extensive use of the disintegration Theorem (see $[18,1]$ for a proof).

Theorem 2.4 (Disintegration Theorem). Let $X, Y$ be two Polish spaces and $\gamma \in \mathcal{P}(X \times Y)$. Then there exists a measurable family $\left(\gamma_{x}\right)_{x \in X}$ of probability measures on $Y$ such that $\gamma=$ $\pi_{1} \gamma \otimes \gamma_{x}$, namely,

$$
\int_{X \times Y} \varphi(x, y) d \gamma(x, y)=\int_{X}\left(\int_{Y} \varphi(x, y) d \gamma_{x}(y)\right) d \pi_{1} \gamma(x)
$$

for all $\varphi \in \mathrm{C}_{b}(X \times Y)$.
Thanks to the Disintegration Theorem, we can easily compose two transport plans.
Definition 2.5 (Composition of plans). Let $X, Y$, and $Z$ be three Polish spaces, and let $\gamma \in$ $\mathcal{M}(X \times Z)$ and $\theta \in \mathcal{M}(Z \times Y)$ be such that

$$
\pi_{1} \gamma=\mu, \quad \pi_{2} \gamma=\pi_{1} \theta=\omega, \quad \pi_{2} \theta=\nu
$$

Then, disintegrating $\gamma=\gamma_{z} \otimes \omega$ and $\theta=\omega \otimes \theta_{z}$, we define the plan $\theta \circ \gamma \in \Pi(\mu, \nu)$ as

$$
\iint_{X \times Y} \varphi(x, y) d \gamma \circ \theta(x, y):=\int_{Z}\left(\int_{X \times Y} \varphi(x, y) d \gamma_{z}(x) d \theta_{z}(y)\right) d \omega(z)
$$

for every $\varphi \in \mathrm{C}_{b}(X \times Y)$.
We can now show the first basic properties concerning the above definitions.
Lemma 2.6. If $\gamma \in \Pi(\mu, \nu)$ and $\theta \ll \gamma$, then $\theta \ll \mathrm{d}_{\pi_{1} \theta \gamma}$ and, analogously, $\theta \ll \mathrm{d}^{\pi_{2} \theta} \gamma$.
Proof. Since characteristic functions of rectangles form a basis for the Borel sets, it is enough to show that, for any rectangle $A \times B \subseteq X \times Y$ such that $\mathrm{d}_{\pi_{1} \theta} \gamma(A \times B)=0$, there holds $\theta(A \times B)=0$.

Let us take such a rectangle and, for any $j \in \mathbb{N}$, set $A_{j}:=\left\{x \in A: \mathrm{d}_{\pi_{1} \theta}(x) \geq 1 / j\right\}$. Then,

$$
0=\mathrm{d}_{\pi_{1} \theta} \gamma(A \times B)=\iint_{A \times B} \mathrm{~d}_{\pi_{1} \theta}(x) d \gamma(x, y) \geq \frac{1}{j} \iint_{A_{j} \times B} d \gamma(x, y)=\frac{\gamma\left(A_{j} \times B\right)}{j}
$$

which letting $j \in \mathbb{N}$ vary implies that $\gamma\left(A^{+} \times B\right)=0$, being $A^{+}=\cup_{j \in \mathbb{N}} A_{j}=A \backslash A_{0}$, where $A_{0}=\left\{x \in A: \mathrm{d}_{\pi_{1} \theta}(x)=0\right\}$. Since $\theta \ll \gamma$, this implies that also $\theta\left(A^{+} \times B\right)=0$, so that

$$
\theta(A \times B)=\theta\left(A_{0} \times B\right) \leq \theta\left(A_{0} \times Y\right)=\pi_{1} \theta\left(A_{0}\right)=\mathrm{d}_{\pi_{1} \theta} \mu\left(A_{0}\right)=\int_{A_{0}} \mathrm{~d}_{\pi_{1} \theta}(x) d \mu(x)=0
$$

and the proof is concluded.
In the above lemma, it is not possible to replace "<" by " $\leq$ " both in the assumptions and in the conclusion, as the next example shows.

Example 2.7. Let $\gamma=\mathcal{L}([0,1] \times[0,1])$ and $\theta=\mathcal{L}([0, a] \times[0, a])$ for some $a \in(0,1)$, so that $\theta \leq \gamma$. Then, $\pi_{1} \theta=a \mathcal{L}([0, a])$, so that $\mathrm{d}_{\pi_{1} \theta}(x)$ equals a in $[0, a]$ and 0 in $(a, 1]$; as a consequence, $\mathrm{d}_{\pi_{1} \theta} \gamma=a \mathcal{L}([0, a] \times[0,1])$, hence we have $\theta \ll \mathrm{d}_{\pi_{1} \theta} \gamma$, according to Lemma 2.6, but not $\theta \leq \mathrm{d}_{\pi_{1} \theta} \gamma$. Actually, we can observe that in general, in the assumptions of Lemma 2.6, one has

$$
\left\|\mathrm{d}_{\pi_{1} \theta} \gamma\right\|_{X \times Y}=\left\|\pi_{1}\left(\mathrm{~d}_{\pi_{1} \theta} \gamma\right)\right\|_{X}=\left\|\pi_{1} \theta\right\|_{X}=\|\theta\|_{X \times Y}
$$

hence one never has $\theta \leq \mathrm{d}_{\pi_{1} \theta} \gamma$, unless the two measures coincide.

It is easy to note that, in general, $\vec{\Phi}(\overleftarrow{\Phi}(\beta)) \neq \beta$ and $\overleftarrow{\Phi}(\vec{\Phi}(\alpha)) \neq \alpha$. Nevertheless, the following holds.

Lemma 2.8. With the notations of Definition 2.3, if $0 \neq \alpha \in \mathcal{M}_{\mu}^{+}(X)$, then $\alpha \ll \overleftarrow{\Phi}(\vec{\Phi}(\alpha))$. In particular, $\alpha \wedge \overleftarrow{\Phi}(\vec{\Phi}(\alpha))>0$.

Proof. The second property follows by the first one - just recall (2.1)- and in turn the first property is a direct consequence of Lemma 2.6. Indeed, setting $\theta=\mathrm{d}_{\alpha} \gamma \ll \gamma$ and recalling Definition 2.3, we get by Lemma 2.6 that

$$
\mathrm{d}_{\alpha} \gamma \ll \mathrm{d}^{\pi_{2}\left(\mathrm{~d}_{\alpha} \gamma\right)} \gamma=\mathrm{d}^{\vec{\Phi}(\alpha)} \gamma,
$$

and the result follows because

$$
\pi_{1}\left(\mathrm{~d}_{\alpha} \gamma\right)=\alpha, \quad \pi_{1}\left(\mathrm{~d}^{\vec{\Phi}(\alpha)} \gamma\right)=\overleftarrow{\Phi}(\vec{\Phi}(\alpha))
$$

The last property that we list here will be quite useful in the following.
Lemma 2.9. Given $\alpha \leq \mu$ and $\beta \leq \nu$, one has

$$
\alpha \wedge \overleftarrow{\Phi}(\beta)>0 \Longleftrightarrow \beta \wedge \vec{\Phi}(\alpha)>0
$$

Proof. By symmetry, the conclusion is obtained as soon as we show that

$$
\begin{equation*}
\alpha \wedge \overleftarrow{\Phi}(\beta)=0 \Longleftrightarrow \gamma(A \times B)=0 \tag{2.2}
\end{equation*}
$$

where the sets $A \subseteq X$ and $B \subseteq Y$ are defined as $A=\left\{x: \mathrm{d}_{\alpha}(x)>0\right\}$ and $B=\left\{y: \mathrm{d}^{\beta}(y)>0\right\}$.
First of all, let us assume that $\gamma(A \times B)=0$, and observe that by construction $\alpha \leq \mu\llcorner A$ and $\beta \leq \nu\llcorner B$, so that

$$
\overleftarrow{\Phi}(\beta) \leq \overleftarrow{\Phi}\left(\nu\llcorner B)=\pi_{1}\left(\mathrm{~d}^{\nu}\llcorner B \gamma)=\pi_{1}(\gamma\llcorner(X \times B))\right.\right.
$$

As a consequence, by definition

$$
\begin{aligned}
\alpha \wedge \overleftarrow{\Phi}(\beta)(X) & \leq\left(\mu \left\llcornerA \wedge \pi_{1}(\gamma\llcorner(X \times B)))(X) \leq \mu\left\llcorner A(X \backslash A)+\pi_{1}(\gamma\llcorner(X \times B))(A)\right.\right.\right. \\
& =\gamma\llcorner(X \times B)(A \times Y)=\gamma((X \times B) \cap(A \times Y))=\gamma(A \times B)=0,
\end{aligned}
$$

by assumption. Hence, the left implication in (2.2) is proved.
On the other hand, assume that $\gamma(A \times B)>0$. Thus, there exists $\varepsilon>0$ such that $\gamma\left(A_{\varepsilon} \times\right.$ $\left.B_{\varepsilon}\right)>0$, being $A_{\varepsilon}=\left\{x: \mathrm{d}_{\alpha}(x)>\varepsilon\right\}, B_{\varepsilon}=\left\{y: \mathrm{d}^{\beta}(y)>\varepsilon\right\}$. Therefore, we get

$$
\alpha \geq \varepsilon \mu\left\llcorner A_{\varepsilon}, \quad \beta \geq \varepsilon \nu\left\llcorner B_{\varepsilon}\right.\right.
$$

The first inequality yields the bound

$$
\alpha \geq \varepsilon \pi_{1}\left(\gamma\left\llcorner\left(A_{\varepsilon} \times Y\right)\right) \geq \varepsilon \pi_{1}\left(\gamma\left\llcorner\left(A_{\varepsilon} \times B_{\varepsilon}\right)\right)\right.\right.
$$

while the second inequality implies $\mathrm{d}^{\beta} \gamma \geq \varepsilon \gamma\left\llcorner\left(X \times B_{\varepsilon}\right) \geq \varepsilon \gamma\left\llcorner\left(A_{\varepsilon} \times B_{\varepsilon}\right)\right.\right.$, so that

$$
\overleftarrow{\Phi}(\beta) \geq \varepsilon \pi_{1}\left(\gamma\left\llcorner\left(A_{\varepsilon} \times B_{\varepsilon}\right)\right)\right.
$$

The last two estimates, together, ensure that $\alpha \wedge \overleftarrow{\Phi}(\beta)>0$, so also the right implication in (2.2) is proved and the proof is concluded.
2.2. The Finite Chain Lemma. This section is devoted to the proof of our main technical tool, which is a general abstract result on measures. Let us first explain what is its meaning.

Let $\gamma$ and $\gamma^{\prime}$ be two transport plans between $\mu$ and $\nu$, and assume for simplicity that all the measures are purely atomic. Assume that we can find $N+1$ pairs $\left(x_{i}, y_{i}\right) \in X \times Y$, for $0 \leq i \leq N$, with the property that each ( $x_{i}, y_{i}$ ) belongs to the support of $\gamma$, and each ( $x_{i}, y_{i+1}$ ) belongs to the support of $\gamma^{\prime}$, with the usual convention that $y_{N+1} \equiv y_{0}$. We can then modify the plan $\gamma^{\prime}$ by making use of $\gamma$, deciding that a small quantity $\varepsilon$ of the mass which was originally moved from every $x_{i}$ to $y_{i+1}$ should be instead moved from $x_{i}$ to $y_{i}$. This modified plan is clearly better than the original one if and only if

$$
\sum_{i=0}^{N} c\left(x_{i}, y_{i}\right) \leq \sum_{i=0}^{N} c\left(x_{i}, y_{i+1}\right)
$$

and this is a crucial property in mass transportation, called cyclical monotonicity, strongly related with the optimality of plans (see for instance [19, 21, 30]). This property is actually also fundamental in the construction of the Kantorovich potential, and we will then use it in Section 5, see Definition 5.1.

Notice that the strategy discussed above makes sense only if we can find a "chain" of pairs as the $\left(x_{i}, y_{i}\right)$ defined before, because this is the only way to guarantee that the modified measure is indeed a transport plan. And actually, the only problem is in "closing" the chain: indeed, we can arbitrarily choose some $\left(x_{0}, y_{0}\right)$ in the support of $\gamma$; since this means that $x_{0}$ is in the support of $\mu$, we find some $y_{1} \in Y$ such that $\left(x_{0}, y_{1}\right)$ is in the support of $\gamma^{\prime}$ (to do this formally, we need to make use of the map $\vec{\Phi}$ of Definition 2.3). In turn, this ensures that $y_{1}$ is in the support of $\nu$, hence (with the map $\overleftarrow{\Phi}$ ) we find some $x_{1} \in X$ such that $\left(x_{1}, y_{1}\right)$ is in the support of $\gamma$, and so on. This procedure will be concluded as soon as some $x_{N}$ has the property that $\left(x_{N}, y_{0}\right)$ is in the support of $\gamma^{\prime}$, and then everything works if we can prove that this will happen sooner or later: this is precisely the goal of our "chain Lemma" below. Actually, we do not make the assumption that the measures are purely atomic, so our procedure is technically a bit more complicate than what we just described, but the underlying idea is the same.

Lemma 2.10 (finite chain Lemma). Let $\gamma, \gamma^{\prime} \in \Pi(\mu, \nu)$, let $\gamma_{0} \leq \gamma$ be a non zero measure, and define $\mu_{0}=\pi_{1} \gamma_{0} \leq \mu$ and $\nu_{0}=\pi_{2} \gamma_{0} \leq \nu$ the marginals of $\gamma_{0}$. Define recursively the measures

$$
\nu_{i+1}=\vec{\Phi}^{\prime}\left(\mu_{i}\right), \quad \mu_{i+1}=\overleftarrow{\Phi}\left(\nu_{i+1}\right)
$$

where $\overleftarrow{\Phi}$ and $\vec{\Phi}^{\prime}$ are the functions of Definition 2.3 relative to $\gamma$ and to $\gamma^{\prime}$ respectively. Then, there exists $j>0$ such that $\nu_{j} \wedge \nu_{0}>0$.

Proof. Let us start by introducing the two sets (notice carefully the slight differences!)

$$
\begin{aligned}
\mathcal{B} & :=\left\{\tilde{\nu} \leq \nu: \text { if } \exists i>0, \nu_{i} \wedge \tilde{\nu}>0, \text { then } \exists j \geq i: \nu_{j} \wedge \nu_{0}>0\right\}, \\
\mathcal{A} & :=\left\{\tilde{\mu} \leq \mu: \text { if } \exists i \geq 0, \mu_{i} \wedge \tilde{\mu}>0, \text { then } \exists j>i: \nu_{j} \wedge \nu_{0}>0\right\} .
\end{aligned}
$$

Notice that both the sets are non-empty since $\mathcal{B}$ contains $\tilde{\nu}=\nu_{0}$, while $\mathcal{A}$ contains $\tilde{\mu}=0$. Set now

$$
\beta:=\sup \{\tilde{\nu}: \tilde{\nu} \in \mathcal{B}\}, \quad \alpha:=\sup \{\tilde{\mu}: \tilde{\mu} \in \mathcal{A}\} .
$$

We can immediately notice that both the suprema are actually maxima, that is, that $\beta \in \mathcal{B}$ and $\alpha \in \mathcal{A}$. To show the first property, assume that for some $i>0$ one has $\nu_{i} \wedge \beta>0$ : by definition of sup of measures, this means that there exists some $\tilde{\nu} \in \mathcal{B}$ such that $\nu_{i} \wedge \tilde{\nu}>0$; hence, by definition of $\mathcal{B}$ there is some $j \geq i$ such that $\nu_{j} \wedge \nu_{0}>0$, and finally this means exactly $\beta \in \mathcal{B}$. The proof that $\alpha \in \mathcal{A}$ is identical.

Assume that, for some $i \geq 0, \mu_{i} \wedge \overleftarrow{\Phi}^{\prime}(\beta)>0$; hence, Lemma 2.9 implies that $\beta \wedge \vec{\Phi}^{\prime}\left(\mu_{i}\right)>0$, that is, $\beta \wedge \nu_{i+1}>0$. Since $\beta \in \mathcal{B}$, this implies the existence of some $j \geq i+1$ such that $\nu_{j} \wedge \nu_{0}>0$, and in turn this establishes that $\overleftarrow{\Phi}^{\prime}(\beta) \in \mathcal{A}$, so $\overleftarrow{\Phi}^{\prime}(\beta) \leq \alpha$. Since the very same argument gives that $\vec{\Phi}(\alpha) \leq \beta$, we obtain $\vec{\Phi}\left(\overleftarrow{\Phi}^{\prime}(\beta)\right) \leq \beta$. But the maps $\vec{\Phi}$ and $\overleftarrow{\Phi}^{\prime}$ do not change the total mass, so the last inequality must be an equality, and this implies

$$
\alpha=\overleftarrow{\Phi}^{\prime}(\beta), \quad \beta=\vec{\Phi}(\alpha),
$$

hence in particular

$$
\begin{equation*}
\|\alpha\|=\|\beta\| . \tag{2.3}
\end{equation*}
$$

Let us now recall that $\alpha \leq \mu$ and $\beta \leq \nu$, so we have $\alpha=\mathrm{d}_{\alpha} \mu$ and $\beta=\mathrm{d}^{\beta} \nu$; if we now define

$$
A:=\left\{x \in X: \mathrm{d}_{\alpha}(x)>0\right\}, \quad B:=\left\{y \in Y: \mathrm{d}^{\beta}(y)>0\right\}, \quad \bar{\mu}:=\mu\llcorner A, \quad \bar{\nu}:=\nu\llcorner B,
$$

we have that for any $i \geq 0$

$$
\mu_{i} \wedge \bar{\mu}>0 \Longleftrightarrow \mu_{i} \wedge \alpha>0
$$

and since $\alpha \in \mathcal{A}$ we deduce that also $\bar{\mu} \in \mathcal{A}$, hence $\bar{\mu} \leq \alpha$. On the other hand, $\alpha \leq \bar{\mu}$ by definition, hence $\alpha=\bar{\mu}$ and then $\mathrm{d}_{\alpha} \equiv 1$ on $A$; similarly, $\beta=\bar{\nu}$ and $\mathrm{d}^{\beta} \equiv 1$ on $B$. Observe that

$$
\nu\left\llcorner B=\beta=\vec{\Phi}(\alpha)=\pi_{2}\left(\mathrm{~d}_{\alpha} \gamma\right)=\pi_{2}(\gamma\llcorner(A \times Y)),\right.
$$

which implies

$$
\gamma(A \times(Y \backslash B))=\pi_{2}(\gamma\llcorner(A \times Y))(Y \backslash B)=\nu\llcorner B(Y \backslash B)=0 .
$$

Then, from (2.3) we readily deduce

$$
\nu(B)=\|\beta\|=\|\alpha\|=\mu(A)=\gamma(A \times Y)=\gamma(A \times B) \leq \gamma(X \times B)=\nu(B),
$$

which also implies that

$$
\gamma((X \backslash A) \times B)=0
$$

We can then deduce that

$$
\overleftarrow{\Phi}(\beta)=\pi_{1}\left(\mathrm{~d}^{\beta} \gamma\right)=\pi_{1}\left(\gamma\llcorner(X \times B))=\pi_{1}\left(\gamma\llcorner(A \times B)) \leq \pi_{1}(\gamma\llcorner(A \times Y))=\mu\llcorner A=\alpha\right.\right.
$$

which again since $\|\alpha\|=\|\beta\|=\|\overleftarrow{\Phi}(\beta)\|$ allows us to infer

$$
\begin{equation*}
\alpha=\overleftarrow{\Phi}(\beta) \tag{2.4}
\end{equation*}
$$

Notice that previously we had found that $\alpha=\overleftarrow{\Phi}^{\prime}(\beta)$. Now, since of course $\nu_{0} \in \mathcal{B}$, then $\nu_{0} \leq \beta$ and so by $(2.4) \overleftarrow{\Phi}\left(\nu_{0}\right) \leq \overleftarrow{\Phi}(\beta)=\alpha$. Moreover, $\gamma_{0} \ll \gamma$ and then Lemma 2.6 ensures that $\gamma_{0} \ll \mathrm{~d}_{\pi_{2} \gamma_{0}} \gamma$, so we get

$$
\mu_{0}=\pi_{1} \gamma_{0} \ll \pi_{1}\left(\mathrm{~d}^{\pi_{2} \gamma_{0}} \gamma\right)=\pi_{1}\left(\mathrm{~d}^{\nu_{0}} \gamma\right)=\overleftarrow{\Phi}\left(\nu_{0}\right) \leq \alpha
$$

which implies $\mu_{0} \wedge \alpha>0$. Finally, since $\alpha \in \mathcal{A}$, the last estimate implies the existence of some $j>0$ for which $\nu_{j} \wedge \nu_{0}>0$, which is the thesis.

We can use the finite chain Lemma 2.10 to make the following decomposition, which is basically the formal way to define the modification that we described at the beginning of the section.

Proposition 2.11. Let $\gamma, \gamma^{\prime} \in \Pi(\mu, \nu), 0 \neq \gamma_{0} \leq \gamma$, and set $\mu_{0}=\pi_{1} \gamma_{0}, \nu_{0}=\pi_{2} \gamma_{0}$. Then, there exist $N \in \mathbb{N}$ and $\bar{\varepsilon}>0$ such that, for every $\varepsilon \leq \bar{\varepsilon}$, there are plans $\tilde{\gamma} \leq \gamma$ and $\tilde{\gamma}^{\prime} \leq \gamma^{\prime}$ satisfying

$$
\begin{array}{lll}
\pi_{1} \tilde{\gamma}=\pi_{1} \tilde{\gamma}^{\prime}=\tilde{\mu}+\mu_{A}, & \pi_{2} \tilde{\gamma}=\tilde{\nu}+\nu_{A}, & \pi_{2} \tilde{\gamma}^{\prime}=\tilde{\nu}+\nu_{B} \\
\mu_{A} \leq \mu_{0}, \quad \tilde{\mu} \leq \mu-\mu_{0}, & \nu_{A}, \nu_{B} \leq \nu_{0}, & \tilde{\nu} \wedge \nu_{0}=0 \\
\|\tilde{\gamma}\|=\left\|\tilde{\gamma}^{\prime}\right\|=(N+1) \varepsilon, & \|\tilde{\mu}\|=\|\tilde{\nu}\|=N \varepsilon, & \left\|\mu_{A}\right\|=\left\|\nu_{A}\right\|=\left\|\nu_{B}\right\|=\varepsilon
\end{array}
$$

In particular, $\tilde{\gamma}$ can be decomposed as $\tilde{\gamma}=\tilde{\gamma}_{0}+\tilde{\gamma}_{\infty}$, where

$$
\tilde{\gamma}_{0} \leq \gamma_{0}, \quad \tilde{\gamma}_{\infty} \leq \gamma-\gamma_{0}, \quad \pi_{1} \tilde{\gamma}_{0}=\mu_{A}, \quad \pi_{2} \tilde{\gamma}_{0}=\nu_{A}, \quad \pi_{1} \tilde{\gamma}_{\infty}=\tilde{\mu}, \quad \pi_{2} \tilde{\gamma}_{\infty}=\tilde{\nu}
$$

Proof. Let us use the same notations as in the proof of Lemma 2.10, and let us define $N$ as the smallest integer such that $\bar{\varepsilon}:=\left\|\nu_{N+1} \wedge \nu_{0}\right\|>0$. For any $\varepsilon \leq \bar{\varepsilon}$, we fix arbitrarily a measure $\tilde{\nu}_{N+1} \leq \nu_{N+1} \wedge \nu_{0}$ such that $\left\|\tilde{\nu}_{N+1}\right\|=\varepsilon$. Now, for any $0<i \leq N$, let us call $\gamma_{i}=\mathrm{d}^{\nu_{i}} \gamma$, so that

$$
\begin{equation*}
\pi_{1} \gamma_{i}=\pi_{1} \mathrm{~d}^{\nu_{i}} \gamma=\overleftarrow{\Phi}\left(\nu_{i}\right)=\mu_{i}, \quad \pi_{2} \gamma_{i}=\nu_{i} \tag{2.5}
\end{equation*}
$$

Analogously, for any $0 \leq i \leq N$ we let $\gamma_{i}^{\prime}=\mathrm{d}_{\mu_{i}} \gamma^{\prime}$, and as before

$$
\pi_{1} \gamma_{i}^{\prime}=\mu_{i}, \quad \pi_{2} \gamma_{i}^{\prime}=\pi_{2} \mathrm{~d}_{\mu_{i}} \gamma^{\prime}=\vec{\Phi}^{\prime}\left(\mu_{i}\right)=\nu_{i+1}
$$

Observe that we have not defined $\gamma_{0}$ because it was already defined; however, by definition, the result of (2.5) holds also for $i=0$. We claim now that

$$
\begin{equation*}
\gamma_{0}+\gamma_{1}+\cdots+\gamma_{N} \leq \gamma, \quad \quad \gamma_{0}^{\prime}+\gamma_{1}^{\prime}+\cdots+\gamma_{N}^{\prime} \leq \gamma^{\prime} \tag{2.6}
\end{equation*}
$$

If $N=0$, there is nothing to prove. Otherwise, by definition of $N$ we know that $\nu_{0} \wedge \nu_{1}=0$, which implies a fortiori that $\gamma_{0} \wedge \gamma_{1}=0$, thus $\gamma_{0}+\gamma_{1} \leq \gamma$. As a consequence, $\mu \geq \mu_{0}+\mu_{1}$, and this gives by definition that $\gamma^{\prime} \geq \gamma_{0}^{\prime}+\gamma_{1}^{\prime}$, which by taking the second marginals ensures $\nu \geq \nu_{1}+\nu_{2}$. If $N=1$, we have then concluded (2.6). If $N>1$, we have also that $\nu_{0} \wedge \nu_{2}=0$, and exactly as before this implies first that $\gamma_{0} \wedge\left(\gamma_{1}+\gamma_{2}\right)=0$, then that $\gamma_{0}+\gamma_{1}+\gamma_{2} \leq \gamma$, then that $\mu \geq \mu_{0}+\mu_{1}+\mu_{2}$, and finally that $\gamma_{0}^{\prime}+\gamma_{1}^{\prime}+\gamma_{2}^{\prime} \leq \gamma^{\prime}$. This proves (2.6) if $N=2$, and with an obvious induction argument we actually derive the validity of (2.6) whatever $N$ is.

We can now easily conclude the proof. Since $\tilde{\nu}_{N+1} \leq \nu_{N+1}=\pi_{2} \gamma_{N}^{\prime}$, we define $\tilde{\gamma}_{N}^{\prime}=\mathrm{d}^{\tilde{\nu}_{N+1}} \gamma_{N}^{\prime}$, and by construction $\tilde{\gamma}_{N}^{\prime} \leq \gamma_{N}^{\prime}$ and $\pi_{2} \tilde{\gamma}_{N}^{\prime}=\tilde{\nu}_{N+1}$. We set then $\tilde{\mu}_{N}=\pi_{1} \tilde{\gamma}_{N}^{\prime}$, and notice that $\tilde{\mu}_{N} \leq \pi_{1} \gamma_{N}^{\prime}=\mu_{N}$. Recalling then that $\mu_{N}=\pi_{1} \gamma_{N}$, we define $\tilde{\gamma}_{N}=\mathrm{d}_{\tilde{\mu}_{N}} \gamma_{N} \leq \gamma_{N}$, and $\tilde{\nu}_{N}=\pi_{2} \tilde{\gamma}_{N} \leq \nu_{N}$. We can now define $\tilde{\gamma}_{N-1}^{\prime}=\mathrm{d}^{\tilde{\nu}_{N}} \gamma_{N-1}^{\prime}$ and continue this construction backward in the obvious way. In the end, for every $i \in\{0,1, \cdots, N\}$ we have built measures $\tilde{\gamma}_{i}, \tilde{\gamma}_{i}^{\prime}, \tilde{\mu}_{i}$ and $\tilde{\nu}_{i}$, with the properties that

$$
\tilde{\mu}_{i}=\pi_{1} \tilde{\gamma}_{i}^{\prime}=\pi_{1} \tilde{\gamma}_{i}, \quad \quad \tilde{\nu}_{i}=\pi_{2} \tilde{\gamma}_{i-1}^{\prime}=\pi_{2} \tilde{\gamma}_{i}
$$

We define now

$$
\tilde{\gamma}=\sum_{i=0}^{N} \tilde{\gamma}_{i}, \quad \tilde{\gamma}^{\prime}=\sum_{i=0}^{N} \tilde{\gamma}_{i}^{\prime}, \quad \tilde{\nu}=\sum_{i=1}^{N} \tilde{\nu}_{i}, \quad \tilde{\mu}=\sum_{i=1}^{N} \tilde{\mu}_{i}, \quad \mu_{A}=\tilde{\mu}_{0}, \quad \nu_{A}=\tilde{\nu}_{0}, \quad \nu_{B}=\tilde{\nu}_{N+1} .
$$

Observe that $\tilde{\gamma} \leq \gamma$ and $\tilde{\gamma}^{\prime} \leq \gamma^{\prime}$ thanks to (2.6): then, by setting $\tilde{\gamma}_{\infty}=\sum_{i=1}^{N} \tilde{\gamma}_{i}$, it is immediate to check all the properties and the proof is concluded.

## 3. Proof of Theorem C

This section is devoted to show our first main result, Theorem C, which says that for supercritical speeds only a negligible set of points moves at maximal speed; we will show this under the assumption that the cost is highly relativistic (see Definition 3.2) and that $\mu \ll \mathscr{L}^{n}$, and Example 4.2 in next Section will show that both these assumptions are sharp. After that, we will give a precise quantitative estimate of how many points are moved at almost maximal speed, in Proposition 3.3.

To start, we need to give the formal definition of the (highly) relativistic cost functions, together with some notation.

Definition 3.1 (Relativistic cost function). Let $\mathscr{C}$ be a strictly convex closed subset of $\mathbb{R}^{n}$, containing the origin in its interior, and let $h: \mathbb{R}^{n} \rightarrow[0,+\infty]$ be a strictly convex function with $h(0)=0$, continuous and bounded in $\mathscr{C}$, and identically $+\infty$ in $\mathbb{R}^{n} \backslash \mathscr{C}$. Then, for every $t>0$ we call relativistic cost function the function

$$
c_{t}(x, y)=h\left(\frac{y-x}{t}\right) .
$$

Definition 3.2 (Highly relativistic cost function). Let $c_{t}$ be a relativistic cost function. For every $\eta$, $\rho \ll 1$ we define

$$
\begin{equation*}
\kappa(\eta, \rho):=\inf \left\{\frac{h(v)-h(v-\lambda w)}{\lambda}: v \in \partial \mathscr{C},|w-v| \leq \rho, 0<\lambda<\eta\right\} \tag{3.1}
\end{equation*}
$$

We say that $c_{t}$ is a highly relativistic cost function if, for $\rho$ small enough, one has $\kappa(\eta, \rho) \rightarrow \infty$ for $\eta \searrow 0$. Notice that this only depends on $h$, and not on $t$.

Notice that the above definition has the following meaning: we are asking that the derivatives of $h$ explode at $\partial \mathscr{C}$, and we write this by introducing the notation $\kappa(\eta, r)$ only in order not to ask differentiability of the function $h$. For instance, the standard Brenier's relativistic heat cost (1.1) satisfies this assumption. We are now able to give the proof of Theorem C, but first we briefly
explain our strategy. The idea is to take an optimal plan $\gamma$ relative to the cost $c_{t}$ and assume that a positive quantity of mass moves exactly of maximal distance; for simplicity, let us think that there is a small cube $Q_{\bar{x}}$ of side $r$ and mass $\varepsilon$ which is translated by $\gamma$ by a vector $v \in t \partial C$ onto another cube $Q_{\bar{y}}$; let us also think that $\mu$ is a constant density on this cube. We can now take the right half of $Q_{\bar{x}}$ and decide to move it to the left half of $Q_{\bar{y}}$ : in this way, the points are not moving of a distance $|v|$, but only of a distance more or less equal to $|v|-r / 2$. Thanks to the fact that $c_{t}$ is a highly relativistic cost, the gain is much bigger than $\varepsilon r$ (and then in the end it appears evident why the Theorem fails for just relativistic costs). Now, we cannot send the left part of $Q_{\bar{x}}$ onto the right part of $Q_{\bar{y}}$, because the points would move of a distance $|v|+r / 2$, and this would have an infinite cost. Here the assumption $t>T$ comes into play: indeed, we can take a plan $\gamma^{\prime}$ relative to some $T<t^{\prime}<t$, and use it to modify $\gamma$ according to Proposition 2.11. Since $t^{\prime}<t$, the modification given by $\gamma^{\prime}$ has finite cost for $\mathcal{C}_{t}$ even after a change of order $r \ll t-t^{\prime}$, and then this gives an admissible plan which pays something of order $N \varepsilon$ for the modification. Since the loss $N \varepsilon$ is much smaller than the gain that we had before, our modified transport plan has a cost strictly smaller than $\gamma$, and the contradiction will conclude the proof. Of course, the proof will be quite more complicate without all the simplifications that we assumed in this discussion.

Proof (of Theorem C). To keep the proof simple to read, we split it in four steps.
Step I. Setting of the construction and definitions.
Let $t>T$, and assume that there exists an optimal plan $\gamma$ for which

$$
\gamma(\{(x, y): y-x \in t \mathscr{C} \mathscr{C}\})>0 .
$$

Let then $(\bar{x}, \bar{y})$ be a Lebesgue point (with respect to $\gamma$ ) of the set $\{y-x \in t \partial \mathscr{C}\}$, and let $r \ll 1$ to be specified during the proof. Without loss of generality, and up to a rescaling of the problem, we can assume that $\bar{y}-\bar{x}=\mathrm{e}_{1}$; let then $Q_{\bar{x}}$ and $Q_{\bar{y}}$ be two cubes, centered at $\bar{x}$ and $\bar{y}$ respectively, with sides parallel to the coordinate axes and of length $r$. Since $(\bar{x}, \bar{y})$ is a $\gamma$-Lebesgue point of the pairs $\{(x, y): y-x \in t \mathscr{C}\}$, and since $\mu \ll \mathscr{L}^{n}$, there is a big constant $K$ such that the measure

$$
\gamma_{0}:=\gamma\left\llcorner\left\{(x, y) \in Q_{\bar{x}} \times Q_{\bar{y}}: y-x \in t \partial \mathscr{C}, \rho(x) \leq K\right\}\right.
$$

is a non-zero measure, being $\rho$ the density of $\mu$ with respect to the Lebesgue measure. As usual, we call $\mu_{0} \leq \mu$ and $\nu_{0} \leq \nu$ the marginals of $\gamma_{0}$.

Let us denote the generic point of $\mathbb{R}^{n}$ as $x=(\sigma, \tau) \in \mathbb{R} \times \mathbb{R}^{n-1}$, and call $\pi_{\tau}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n-1}$ the projection on the $\tau$ variable; for simplicity of notation, let us assume that the $\sigma$-coordinates of the points of $Q_{\bar{x}}$ are between 0 and $r$. We disintegrate now $\mu_{0}=\mu^{\tau} \otimes \alpha$, being $\alpha=\pi_{\tau \#} \mu_{0}$; recall that each $\mu^{\tau}$ is a probability measure concentrated in the segment [0,r], and that by Fubini Theorem $\alpha$-a.e. measure $\mu^{\tau}$ is absolutely continuous with respect to $\mathscr{L}^{1}$. Let us call now $m=\left\|\mu_{0}\right\|$ and let $\delta=\delta(r) \ll 1$ be a small quantity, to be specified later; for $\alpha$-a.e. $\tau$, we can define the function

$$
f_{\tau}:\left\{\sigma \in[0, r]: \mu^{\tau}([0, \sigma]) \leq 1-\delta\right\} \longrightarrow\left\{\sigma \in[0, r]: \mu^{\tau}([0, \sigma]) \geq \delta\right\}
$$

as $f_{\tau}(\sigma)=\sigma^{\prime}$ where

$$
\mu^{\tau}\left(\left[0, \sigma^{\prime}\right]\right)=\mu^{\tau}([0, \sigma])+\delta .
$$

In words, this means that we are taking all the points of the segment $\{(\sigma, \tau): 0 \leq \sigma \leq r\}$ except the last portion $\delta$ on the right, and we are calling $f_{\tau}$ the map associating to each $\sigma$ a point $f_{\tau}(\sigma)$ "after" a portion $\delta$ of mass. In average, we should expect points to be translated by $f_{\tau}$ of a quantity $\delta r$ to the right: of course this would be true only if $\mu^{\tau}$ were constant, and actually all what we can say is that every point is moved to the right because $\mu^{\tau}\left(\left[0, \sigma^{\prime}\right]\right)>\mu^{\tau}([0, \sigma])$ implies $\sigma^{\prime}>\sigma$; nevertheless, it is useful to keep in mind this "average" idea. In fact, let us fix a big constant $M=M(r)$, to be specified later, and let us set

$$
Z:=\left\{\tau: \exists \sigma \in[0, r]: f_{\tau}(\sigma)<\sigma+\frac{\delta r}{M}\right\} .
$$

So, the elements $\tau$ of $Z$ are those for which at least some point $\sigma$ of the horizontal segment moves only very little to the right compared with the "average". We can expect $Z$ to contain only few points if $M$ is big enough, and this is precisely the content of next step.
Step II. One has $\alpha(Z) \leq m / 3$.
By the Measurable Selection Theorem (see for instance [17, 18]), there exists a measurable function $\tau \mapsto \sigma(\tau)$ on $Z$ such that, for any $\tau \in Z$,

$$
f_{\tau}(\sigma(\tau))-\sigma(\tau)<\frac{\delta r}{M}
$$

Define then the box

$$
\Gamma:=\left\{(\sigma, \tau) \in Q_{\bar{x}}: \tau \in Z, \sigma(\tau) \leq \sigma \leq \sigma(\tau)+\frac{\delta r}{M}\right\} .
$$

Then, on the one hand we have

$$
\mu_{0}(\Gamma) \leq K \mathscr{L}^{n}(\Gamma)=K \frac{\delta r}{M} \mathscr{L}^{n-1}(Z) \leq \frac{K \delta r^{n}}{M}
$$

and on the other hand

$$
\mu_{0}(\Gamma)=\int_{Z}\left(\int_{\sigma(\tau)}^{\sigma(\tau)+\frac{\delta r}{M}} 1 d \mu^{\tau}(\sigma)\right) d \alpha(\tau) \geq \int_{Z}\left(\int_{\sigma(\tau)}^{f_{\tau}(\sigma(\tau))} 1 d \mu^{\tau}(\sigma)\right) d \alpha(\tau)=\delta \alpha(Z) .
$$

Putting together the last two estimates, it is clear that the claim holds true if we choose, for instance,

$$
\begin{equation*}
M=\frac{3 K r^{n}}{m} . \tag{3.2}
\end{equation*}
$$

Notice carefully that such $M$ depends on $r$ but not on $\delta$.
Step III. The measures $\xi_{1}$ and $\xi_{2}$.
Let us start to build the competitor $\xi \in \Pi(\mu, \nu)$ for $\gamma$. First of all, we set $W=\{\tau \notin Z\}$ and

$$
L:=\left\{(\sigma, \tau): \tau \in W, \mu^{\tau}([0, \sigma]) \leq 1-\delta\right\}, \quad R:=\left\{(\sigma, \tau): \tau \in W, \mu^{\tau}([0, \sigma]) \geq \delta\right\} .
$$

We define now the functions $\tilde{g}: L \rightarrow Q_{\bar{x}}$ and $g: L \times \mathbb{R}^{n} \rightarrow \mathbb{R}^{n} \times \mathbb{R}^{n}$ as

$$
\tilde{g}(\sigma, \tau):=\left(f_{\tau}(\sigma), \tau\right), \quad g(x, y):=(\tilde{g}(x), y) .
$$

Notice that $\tilde{g}$ is a $\mu_{0}$-essentially bijective function from $L$ to $R$. We can now define the first two pieces of $\xi$, namely,

$$
\xi_{1}:=\gamma_{0}\left\llcorner\left\{(x, y) \in Q_{\bar{x}} \times Q_{\bar{y}}: x \in[0, r] \times Z\right\}, \quad \xi_{2}:=g_{\#}\left(\gamma_{0}\left\llcorner\left(L \times \mathbb{R}^{n}\right)\right)\right.\right.
$$

The meaning of the last two definitions is simple, keeping in mind the discussion before the beginning of this proof: for pairs $(x, y) \in Q_{\bar{x}} \times Q_{\bar{y}}$ for which $x=(\sigma, \tau)$ and $\tau \in Z$, we decide not to do any change, so $\xi_{1}$ equals $\gamma_{0}$ for those points. For the points for which $\tau \in W$, instead, we want to shorten the path: $\xi_{2}$ takes the "right points" which belong to $R$, and move them as the corresponding "left points" in $L$ were doing; it is easy to imagine that this will shorten the paths, and that in the end we will remain with a very left part of the cube $Q_{\bar{x}}$ which must go onto a very right part of the cube $Q_{\bar{y}}$ : we will take care of this in the next step with $\xi_{3}$ and $\xi_{4}$. Let us now calculate the projections of $\xi_{2}$ (since the projections of $\xi_{1}$ are obvious). For the first one, for any Borel set $A \subseteq \mathbb{R}^{n}$ we have

$$
\begin{aligned}
\pi_{1} \xi_{2}(A) & =\xi_{2}\left(A \times \mathbb{R}^{n}\right)=\gamma_{0}\left(\left(L \times \mathbb{R}^{n}\right) \cap g^{-1}\left(A \times \mathbb{R}^{n}\right)\right)=\gamma_{0}\left(g^{-1}\left(A \times \mathbb{R}^{n}\right)\right)=\gamma_{0}\left(\tilde{g}^{-1}(A) \times \mathbb{R}^{n}\right) \\
& =\mu_{0}\left(\tilde{g}^{-1}(A)\right)=\iint_{(\sigma, \tau) \in Q_{\bar{x}}} \chi_{\tilde{g}^{-1}(A)}(\sigma, \tau) d \mu_{0}(\tau, \sigma) \\
& =\int_{\tau \in W} \int_{\sigma \in(0, r)} \chi_{\tilde{g}^{-1}(A)}(\sigma, \tau) d \mu^{\tau}(\sigma) d \alpha(\tau)=\int_{\tau \in W} \mu^{\tau}\left(\left\{\sigma:(\sigma, \tau) \in \tilde{g}^{-1}(A)\right\}\right) d \alpha(\tau)
\end{aligned}
$$

Now we observe that, by definition of $f_{\tau}$ and of $g$, for any $\tau$ one clearly has

$$
\mu^{\tau}\left(\left\{\sigma:(\sigma, \tau) \in \tilde{g}^{-1}(A)\right\}\right)=\mu^{\tau}(\{\sigma:(\sigma, \tau) \in A \cap R\}) .
$$

Hence, from last estimate we finally get

$$
\pi_{1} \xi_{2}(A)=\int_{\tau \in W} \mu^{\tau}(\{\sigma:(\sigma, \tau) \in A \cap R\}) d \alpha(\tau)=\mu_{0}(A \cap R)
$$

In the very same way, using the fact the the second component of $g$ is the identity, one directly obtains that

$$
\pi_{2} \xi_{2}(B)=\gamma_{0}(L \times B)
$$

Since the projections of $\xi_{1}$ are obviously the same of the restriction of $\gamma_{0}$ to points $(\sigma, \tau)$ with $\tau \in Z$, we simply have

$$
\begin{equation*}
\pi_{1}\left(\xi_{1}+\xi_{2}\right)=\mu_{0}-\mu_{\mathrm{rem}}, \quad \pi_{2}\left(\xi_{1}+\xi_{2}\right)=\nu_{0}-\nu_{\mathrm{rem}} \tag{3.3}
\end{equation*}
$$

where the "remaining measures" are

$$
\begin{aligned}
& \mu_{\mathrm{rem}}=\mu_{0}\left\llcorner\left(\left(Q_{\bar{x}} \backslash R\right) \cap\{(\sigma, \tau): \tau \in W\}\right),\right. \\
& \nu_{\mathrm{rem}}=\pi_{2}\left(\gamma_{0}\left\llcorner\left[\left(\left(Q_{\bar{x}} \backslash L\right) \cap\{(\sigma, \tau): \tau \in W\}\right) \times \mathbb{R}^{n}\right]\right) .\right.
\end{aligned}
$$

Let us call $\varepsilon:=\left\|\mu_{\text {rem }}\right\|=\left\|\nu_{\text {rem }}\right\|$, and notice that by Step II, we have

$$
\begin{equation*}
\varepsilon=\delta \alpha(W) \in\left[\frac{2}{3} m \delta, m \delta\right] \tag{3.4}
\end{equation*}
$$

To conclude this step, we need to evaluate also the costs of $\xi_{1}$ and $\xi_{2}$. To do so, take any $x=(\sigma, \tau) \in L$, and let $y \in Q_{\bar{y}}$ be a point such that $(x, y)$ is in the support of $\gamma_{0}$. Call now $v=(y-x) / t \in \partial \mathscr{C}$ and $w=\mathrm{e}_{1} / t$, and notice that by construction $\tilde{g}(x)-x=\lambda w t$ for some

$$
\frac{\delta r}{M} \leq \lambda \leq r
$$

where the second inequality comes from geometric arguments, while the first one is true because $x \in L$ implies $\tau \in W$; observe also that $|w-v| \leq \sqrt{n} r / t$ (keep in mind that $|\bar{y}-\bar{x}|=1$ ). Notice also that, since $h$ is strictly convex, we readily obtain that the function $\lambda \mapsto h(v-\lambda w)$ is decreasing for $\lambda \in(\delta r / M, r)$ as soon as $r$ is small enough. We can then evaluate

$$
\begin{align*}
c_{t}(g(x, y))-c_{t}(x, y) & =c_{t}(\tilde{g}(x), y)-c_{t}(x, y)=h\left(\frac{y-\tilde{g}(x)}{t}\right)-h\left(\frac{y-x}{t}\right) \\
& =h(v-\lambda w)-h(v) \leq h\left(v-\frac{\delta r}{M} w\right)-h(v) \leq-\frac{\delta r}{M} \kappa\left(\frac{\delta r}{M}, \frac{\sqrt{n} r}{t}\right) \tag{3.5}
\end{align*}
$$

As a consequence, the cost of $\xi_{1}+\xi_{2}$ is

$$
\begin{align*}
\mathcal{C}_{t}\left(\xi_{1}+\xi_{2}\right) & =\iint_{\mathbb{R}^{n} \times \mathbb{R}^{n}} c_{t}(x, y) d \xi_{1}(x, y)+\iint_{\mathbb{R}^{n} \times \mathbb{R}^{n}} c_{t}(x, y) d \xi_{2}(x, y) \\
& =\int_{[0, r] \times Z \times Q_{\bar{y}}} c_{t}(x, y) d \gamma_{0}(x, y)+\iint_{L \times \mathbb{R}^{n}} c_{t}(g(x, y)) d \gamma_{0}(x, y) \\
& \leq \int_{[0, r] \times Z \times Q_{\bar{y}}} c_{t}(x, y) d \gamma_{0}(x, y)+\iint_{L \times \mathbb{R}^{n}} c_{t}(x, y)-\frac{\delta r}{M} \kappa\left(\frac{\delta r}{M}, \frac{\sqrt{n} r}{t}\right) d \gamma_{0}(x, y)  \tag{3.6}\\
& \leq \mathcal{C}_{t}\left(\gamma_{0}\right)-\frac{\delta r}{M} \kappa\left(\frac{\delta r}{M}, \frac{\sqrt{n} r}{t}\right) \gamma_{0}\left(L \times \mathbb{R}^{n}\right)=\mathcal{C}_{t}\left(\gamma_{0}\right)-\frac{\delta r}{M} \kappa\left(\frac{\delta r}{M}, \frac{\sqrt{n} r}{t}\right) \mu_{0}(L) \\
& =\mathcal{C}_{t}\left(\gamma_{0}\right)-\frac{\delta r}{M} \kappa\left(\frac{\delta r}{M}, \frac{\sqrt{n} r}{t}\right)(1-\delta) \alpha(W)=\mathcal{C}_{t}\left(\gamma_{0}\right)-\frac{\varepsilon r}{2 M} \kappa\left(\frac{\delta r}{M}, \frac{\sqrt{n} r}{t}\right)
\end{align*}
$$

where we have used also (3.4) and the fact that $\delta \leq 1 / 2$.
Step IV. The measures $\xi_{3}$ and $\xi_{4}$.
We now present that last two pieces of the competitor $\xi \in \Pi(\mu, \nu)$. Let us start by recalling that by (3.3) and (3.4) the marginals of the measure $\xi_{1}+\xi_{2}$ cover almost all $\mu_{0}$ and $\nu_{0}$, and then it would be easy to conclude the competitor simply by adding the whole $\gamma-\gamma_{0}$ plus a plan transporting $\mu_{\text {rem }}$ onto $\nu_{\text {rem }}$. Unfortunately, this would have an infinite cost, because points of $\mu_{\text {rem }}$ and point of $\nu_{\text {rem }}$ have distance more or less $\mathrm{e}_{1}+r$, which is outside $t \mathscr{C}$. To overcome this problem, we use the assumption that $t>T$, then we select some $T<t^{\prime}<t$ and a transport plan $\gamma^{\prime} \in \Pi(\mu, \nu)$ which has finite $\mathcal{C}_{t^{\prime}}$ cost (we could take, for instance, some optimal plan for $\mathcal{C}_{t^{\prime}}$, but the finiteness of the cost is enough). Then, we apply Proposition 2.11 and we get some $N \in \mathbb{N}$ and $\bar{\varepsilon}>0$, depending on $\gamma, \gamma^{\prime}$ and $\gamma_{0}$, also on $r$, but not on $\delta$; if $\varepsilon \leq \bar{\varepsilon}$, then we get also plans $\tilde{\gamma}$ and $\tilde{\gamma}^{\prime}$ satisfying the claim of the proposition. But actually, by (3.4) we can assume $\varepsilon \leq \bar{\varepsilon}$ as soon as we choose $\delta$ small enough; thus, Proposition 2.11 provides us with two plans $\tilde{\gamma}=\tilde{\gamma}_{0}+\tilde{\gamma}_{\infty}$ and $\tilde{\gamma}^{\prime}$, being $\tilde{\gamma}_{0} \leq \gamma_{0}$ and $\tilde{\gamma}_{\infty} \leq \gamma-\gamma_{0}$. We can then define $\xi_{3}=\gamma-\gamma_{0}-\tilde{\gamma}_{\infty}$, which is a positive measure whose marginals are

$$
\pi_{1} \xi_{3}=\mu-\mu_{0}-\tilde{\mu}, \quad \pi_{2} \xi_{3}=\nu-\nu_{0}-\tilde{\nu}
$$

so that by (3.3) we have now

$$
\begin{equation*}
\pi_{1}\left(\xi_{1}+\xi_{2}+\xi_{3}\right)=\mu-\mu_{\mathrm{rem}}-\tilde{\mu}, \quad \pi_{2}\left(\xi_{1}+\xi_{2}+\xi_{3}\right)=\nu-\nu_{\mathrm{rem}}-\tilde{\nu} \tag{3.7}
\end{equation*}
$$

Since $\xi=\xi_{1}+\xi_{2}+\xi_{3}+\xi_{4}$ must be in $\Pi(\mu, \nu)$, we need then the marginals of $\xi_{4}$ to be $\tilde{\mu}+\mu_{\text {rem }}$ and $\tilde{\nu}+\nu_{\text {rem }}$. Since the marginals of $\tilde{\gamma}^{\prime}$ are $\tilde{\mu}+\mu_{A}$ and $\tilde{\nu}+\nu_{B}$, the plan $\tilde{\gamma}^{\prime}$ does the job almost perfectly, the only mistake being the presence of $\mu_{A}$ and $\nu_{B}$ instead of $\mu_{\mathrm{rem}}$ and $\nu_{\mathrm{rem}}$. However, notice that both $\mu_{A}$ and $\mu_{\mathrm{rem}}$ are measures smaller than $\mu_{0}$ and of mass $\varepsilon$, and similarly both $\nu_{B}$ and $\nu_{\mathrm{rem}}$ are measures smaller than $\nu_{0}$ and of mass $\varepsilon$. Let us then introduce two auxiliary transport plans, namely

$$
\alpha=(\mathrm{Id}, \mathrm{Id})_{\#} \tilde{\mu}+\mu_{\mathrm{rem}} \otimes \mu_{A}, \quad \beta=(\mathrm{Id}, \mathrm{Id})_{\#} \tilde{\nu}+\nu_{B} \otimes \nu_{\mathrm{rem}}
$$

whose marginals are precisely $\tilde{\mu}+\mu_{\text {rem }}$ and $\tilde{\mu}+\mu_{A}$, and $\tilde{\nu}+\nu_{B}$ and $\tilde{\nu}+\nu_{\text {rem }}$ respectively. Let us finally define $\xi_{4}=\beta \circ \tilde{\gamma}^{\prime} \circ \alpha$, so that

$$
\pi_{1} \xi_{4}=\tilde{\mu}+\mu_{\mathrm{rem}}, \quad \pi_{2} \xi_{4}=\tilde{\nu}+\nu_{\mathrm{rem}}
$$

so that by (3.7) we finally get that $\xi \in \Pi(\mu, \nu)$, so $\xi$ is an admissible competitor. To conclude, we need to estimate the costs of $\xi_{3}$ and $\xi_{4}$. While for $\xi_{3}$ we simply have

$$
\begin{equation*}
\mathcal{C}_{t}\left(\xi_{3}\right)=\mathcal{C}_{t}(\gamma)-\mathcal{C}_{t}\left(\gamma_{0}\right)-\mathcal{C}_{t}\left(\tilde{\gamma}_{\infty}\right) \leq \mathcal{C}_{t}(\gamma)-\mathcal{C}_{t}\left(\gamma_{0}\right) \tag{3.8}
\end{equation*}
$$

concerning $\xi_{4}$ we must observe what follows: since $\tilde{\gamma}^{\prime} \leq \gamma^{\prime}$, and $\gamma^{\prime}$ has a finite $\mathcal{C}_{t^{\prime}}$ cost, we know that $y-x \in t^{\prime} \mathscr{C}$ for $\gamma^{\prime}$-a.e. $(x, y)$. Hence, since $\mu_{A}$ and $\mu_{\text {rem }}$ are both supported in $Q_{\bar{x}}$ and $\nu_{B}$ and $\nu_{\text {rem }}$ are both supported in $Q_{\bar{y}}$, we derive that $\alpha$ and $\beta$ move points at most of a distance $\sqrt{n} r$, and then for $\xi_{4}$-a.e. $(x, y)$ one has $y-x \in t^{\prime} \mathscr{C}+2 \sqrt{n} r$. Since we are free to chose $r$ as small as we wish, and $t^{\prime}<t$, it is admissible to assume that $t^{\prime} \mathscr{C}+2 \sqrt{n} r \subseteq t \mathscr{C}$, thus $\xi_{4}$ has a finite $\mathcal{C}_{t}$ cost (and maybe an infinite $\mathcal{C}_{t^{\prime}}$ cost, but this is not a problem); notice that formally we are first choosing $t^{\prime}$, then $r$, then $M$, and only at the end $\delta$. Recalling that $h$ is bounded where it is finite, and calling $C$ the maximum of $h$ in $\mathscr{C}$, we get then

$$
\begin{equation*}
\mathcal{C}_{t}\left(\xi_{4}\right) \leq C\left\|\xi_{4}\right\|=C(N+1) \varepsilon . \tag{3.9}
\end{equation*}
$$

Putting together (3.6), (3.8) and (3.9), we finally get

$$
\begin{equation*}
\mathcal{C}_{t}(\xi) \leq \mathcal{C}_{t}(\gamma)+\varepsilon\left(C(N+1)-\frac{r}{2 M} \kappa\left(\frac{\delta r}{M}, \frac{\sqrt{n} r}{t}\right)\right) \tag{3.10}
\end{equation*}
$$

Let us finally use the assumption that $c_{t}$ is a highly relativistic cost function: when we choose $r$ only $t^{\prime}$ has been fixed, hence we are allowed to chose it so small that $\kappa(\eta, \sqrt{n} r / t) \rightarrow \infty$ for $\eta \searrow 0$. Having chosen $r$, we get $M$ by Step II, and $N$ by Proposition 2.11. Thus, we are free to chose $\delta$ depending on $N$ and $r$; since $c_{t}$ is highly relativistic, we can do this in such a way that the parentesis in (3.10) is strictly negative. We have then found a competitor with cost strictly smaller than the optimal plan $\gamma$, and the contradiction gives the thesis.

We can now easily give a "quantitative" version of Theorem C, where we state precisely how many pairs $(x, y)$ can be moved from an optimal plan $\gamma$ of a vector which is inside $t \mathscr{C}$ but very close to $t \partial \mathscr{C}$; it is important to notice that the next result holds for any relativistic cost
function, regardless whether or not it is also highly relativistic. To state our result, we have to generalize the definition (3.1) of $\kappa(\eta, \rho)$ as follows: for every $s \ll 1$, we set

$$
\begin{equation*}
\kappa_{s}(\rho):=\lim _{\eta \searrow 0} \kappa_{s}(\eta, \rho)=\lim _{\eta \searrow 0} \inf \left\{\frac{h(v)-h(v-\lambda w)}{\lambda}: v \in \overline{\mathscr{C}} \backslash(1-s) \mathscr{C},|w-v| \leq \rho, 0<\lambda<\eta\right\} . \tag{3.11}
\end{equation*}
$$

Proposition 3.3. Let $c_{t}$ be a relativistic cost function, $t>T$, and assume that $\mu \in L^{\infty}$ and the supports of $\mu$ and $\nu$ are bounded. Then, for any $t>T$ and any $\rho$ small enough, there exists a constant $\bar{C}=\bar{C}(\mu, \nu, t, \rho)$ such that, for s small enough, any optimal plan $\gamma$ for $\mathcal{C}_{t}$ satisfies

$$
\begin{equation*}
\gamma\left(\left\{(x, y) \in \mathbb{R}^{n} \times \mathbb{R}^{n}: \frac{y-x}{t} \in \overline{\mathscr{C}} \backslash(1-s) \mathscr{C}\right\}\right) \leq \frac{\bar{C}}{\kappa_{s}(\rho)} . \tag{3.12}
\end{equation*}
$$

Proof. We will make use of the same main idea of the proof of Theorem C, with some important modifications. Let us start from the beginning: first of all, we fix once for all $T<t^{\prime}<t$, a plan $\gamma^{\prime}$ with finite $\mathcal{C}_{t^{\prime}}$ cost, and a constant $r$ such that the inclusion $t^{\prime} \mathscr{C}+2 \sqrt{n} r \subseteq t \mathscr{C}$ holds.

We can find a finite number points $x_{i}$ and $y_{j}$ such that the cubes centered at $x_{i}$ (resp., $y_{i}$ ) with side-length $r$ cover the support of $\mu$ (resp., of $\nu$ ): here we use the fact that both the supports are bounded. The support of $\gamma$ is then covered by a finite number $C_{1}$ of products of cubes $Q_{x_{i}} \times Q_{y_{j}}$. Let us fix such a product, say $Q_{x_{i}} \times Q_{y_{j}}$ and consider the measure $\gamma_{1}=\gamma_{1}(i, j)$ given by

$$
\gamma_{1}:=\gamma\left\llcorner\left(Q_{x_{i}} \times Q_{y_{j}}\right) .\right.
$$

If this measure is non-zero, we apply Proposition 2.11 finding constants $N(i, j)$ and $\bar{\varepsilon}(i, j)$. We call simply $N$ and $\bar{\varepsilon}$ the biggest and smallest of the constants $N(i, j)$ and $\bar{\varepsilon}(i, j)$. For any $s<1$, now, we define

$$
\gamma_{s}:=\gamma\left\llcorner\left\{(x, y) \in Q_{x_{i}} \times Q_{y_{j}}, \frac{y-x}{t} \in \overline{\mathscr{C}} \backslash(1-s) \mathscr{C}\right\} .\right.
$$

Notice that $\gamma_{s} \leq \gamma_{1}$. We will now repeat the very same construction as in the proof of Theorem C , using $\gamma_{s}$ in place of $\gamma_{0}$ and $\left(x_{i}, y_{j}\right)$ in place of $(\bar{x}, \bar{y})$. In particular, we give the same definition of $f_{\tau}$ and $Z$ as in Step I there, with $\gamma_{0}$ replaced by $\gamma_{s}$; then, we repeat exactly the Step II, so we find that $\alpha(Z) \leq m_{i j} / 3$ as soon as we have chosen

$$
\begin{equation*}
M=\frac{3\|\mu\|_{L^{\infty}} r^{n}}{m_{i j}}, \tag{3.13}
\end{equation*}
$$

where $m_{i j}=\left\|\gamma_{s}\right\|$-compare with the estimate (3.2). Observe carefully that the choice of $M$ depends on $m_{i j}$, which in turn depends also on $s$. Notice also that the assumption that $\mu \in L^{\infty}$ is used in (3.13) only.

Now, we repeat the Step III, defining the measures $\xi_{1}$ and $\xi_{2}$; keep in mind that, in the whole proof of Theorem C, we used only once the fact that the measure $\gamma_{0}$ in that proof was concentrated on pairs $(x, y)$ with $y-x \in t \partial \mathscr{C}$, that is, to obtain the estimate (3.5) on the gain in the cost of a single pair, and in turn (3.5) was used only immediately after, to obtain (3.6). Since now the measure $\gamma_{s}$ is concentrated in pairs $(x, y)$ such that $(y-x) / t$ does not necessarily
belong to $\partial \mathscr{C}$, but to $\overline{\mathscr{C}} \backslash(1-s) \mathscr{C}$, we just have to substitute $\kappa$ with $\kappa_{s}$, so this time instead of (3.5) we get

$$
c_{t}(g(x, y))-c_{t}(x, y) \leq-\frac{\delta r}{M} \kappa_{s}(\delta r / M, \rho)
$$

where we write for brevity $\rho=r \sqrt{n} / t$, and then instead of (3.6) we get

$$
\begin{equation*}
\mathcal{C}_{t}\left(\xi_{1}+\xi_{2}\right) \leq \mathcal{C}_{t}\left(\gamma_{s}\right)-\frac{\varepsilon r}{2 M} \kappa_{s}(\delta r / M, \rho) . \tag{3.14}
\end{equation*}
$$

We have now to repeat Step IV: first recall that, in the construction of $\xi_{1}$ and $\xi_{2}$, we have used the measure $\gamma_{s}$ in place of $\gamma_{0}$; nevertheless, we have applied Proposition 2.11 to $\gamma_{0}$ : this is of primary importance, since in this way $N$ and $\bar{\varepsilon}$ do not depend on $s$. Then, this time we have that the two projections of $\xi_{1}+\xi_{2}$ cover almost completely the two projections of $\gamma_{s}$; hence similarly as in Step IV of last proof, we set now $\xi_{3}=\gamma-\gamma_{s}-\tilde{\gamma}_{\infty}$, which is positive since $\tilde{\gamma}_{\infty} \leq \gamma-\gamma_{0} \leq \gamma-\gamma_{s}$, and finally $\xi_{4}$ is again exactly as before a plan having the correct marginals so that $\xi \in \Pi(\mu, \nu)$. Exactly as before, for almost every pair $(x, y)$ in the support of $\xi_{4}$ we have that $y-x \in t^{\prime} \mathscr{C}+2 \sqrt{n} r$, which by assumption is in $t \mathscr{C}$ (and again, the choice of $s$ has no influence on this). Therefore, (3.8) and (3.9) perfectly generalize to

$$
\mathcal{C}_{t}\left(\xi_{3}\right)=\mathcal{C}_{t}(\gamma)-\mathcal{C}_{t}\left(\gamma_{s}\right), \quad \mathcal{C}_{t}\left(\xi_{4}\right) \leq C_{2}(N+1) \varepsilon
$$

which together with (3.14) tells us that (3.10) becomes now

$$
\mathcal{C}_{t}(\xi) \leq \mathcal{C}_{t}(\gamma)+\varepsilon\left(C_{2}(N+1)-\frac{r}{2 M} \kappa_{s}(\delta r / M, \rho)\right)
$$

Since $\gamma$ is an optimal measure, we get that the term in parenthesis must be positive, which by letting $\delta \searrow 0$ means

$$
M \geq \frac{r}{2 C_{2}(N+1)} \kappa_{s}(\rho) .
$$

Recalling now the choice of $M$ in (3.13), we deduce that

$$
\left\|\gamma_{s}\right\|=m_{i j} \leq \frac{6 C_{2}(N+1)\|\mu\|_{L^{\infty}} r^{n-1}}{\kappa_{s}(\rho)}=\frac{C_{3}}{\kappa_{s}(\rho)} .
$$

Basically, we have found that the pairs in $C_{x_{i}} \times C_{y_{j}}$ which are moved of ( $t$ times) a vector in $\overline{\mathscr{C}} \backslash(1-s) \mathscr{C}$ are only few. Putting now together all the pairs of cubes covering the support of $\gamma$, we precisely find (3.12), with $\bar{C}=C_{1} C_{3}$, and the proof is then concluded.

Remark 3.4. Observe that the above result is valid for any relativistic cost, not necessarily highly relativistic: in fact, in the proof of Theorem $C$ we used the highly relativistic assumption only at the very end, to use that $\kappa(\eta, \rho) \rightarrow \infty$ for $\eta \searrow 0$; instead, this time the constant $\kappa_{s}(\rho)$ is surely finite for every $s>0$. Nevertheless, if the cost is highly relativistic, then $\kappa_{s}(\rho) \rightarrow+\infty$ for $s \searrow 0$ if $\rho$ is small enough; hence, in the particular case of a highly relativistic cost the result of Proposition 3.3 is stronger than Theorem C. Notice also that, in the particular case of the relativistic heat cost (1.1), formula (3.12) takes the nice form

$$
\gamma\left(\left\{(x, y) \in \mathbb{R}^{n} \times \mathbb{R}^{n}: \frac{y-x}{t} \in \overline{\mathscr{C}} \backslash(1-s) \mathscr{C}\right\}\right) \leq \bar{C} \sqrt{s}
$$

## 4. Continuity of the total cost function

In this section, we prove our result on the continuity of the total cost function $\mathcal{C}(t)$, that is, Theorem A. First of all, we need the following axiliary definition.

Definition 4.1. Let $c_{t}$ be a relativistic cost function. Then, for every $\delta>0$ we call

$$
\operatorname{exc}(\delta):=\sup \left\{\frac{h(v)}{h((1-\delta) v)}: v \in \mathscr{C}\right\}-1 .
$$

Notice that $\operatorname{exc}(\delta) \searrow 0$ when $\delta \searrow 0$.
Proof (of Theorem A). The proof of this Theorem is quite similar to that of Theorem C and Proposition 3.3, and in particular it makes use of Proposition 2.11.

Notice first that the function $c_{t}$ decreases with $t$, so the inequality $\mathcal{C}(t) \leq \mathcal{C}(t-\delta)$ is obvious for every $\delta$ and we only have to show the second inequality in (1.3). Let us then fix, as usual, some $t>T$, some $T<t^{\prime}<t$, call $\gamma$ and $\gamma^{\prime}$ two optimal plans corresponding to $\mathcal{C}_{t}$ and $\mathcal{C}_{t^{\prime}}$, and let $r$ be a small side-length with the property that the inclusion $t^{\prime} \mathscr{C}+2 \sqrt{n} r \subset \subset t \mathscr{C}$ holds. Then, let us fix some $\bar{\delta}>0$ with the property that

$$
\begin{equation*}
t^{\prime} \mathscr{C}+2 \sqrt{n} r \subseteq(t-\bar{\delta}) \mathscr{C} . \tag{4.1}
\end{equation*}
$$

For every small $\delta>0$, we define

$$
S_{\delta}:=\left\{(x, y) \in \mathbb{R}^{2 n}: \frac{y-x}{t} \in(1-\delta) \mathscr{C}\right\} .
$$

We can now cover the support of $\gamma\left\llcorner\left\{(x, y) \notin S_{\bar{\delta}}\right\}\right.$ with $C_{1}$ cubes $Q_{i}$ in $\mathbb{R}^{2 n}$ with side-length $r$; for every $1 \leq i \leq C_{1}$ and for every $\delta<\bar{\delta}$, we set then

$$
\gamma_{\bar{\delta}}^{i}:=\gamma\left\llcorner Q_{i} \backslash S_{\bar{\delta}}, \quad \gamma_{\delta}^{i}:=\gamma\left\llcorner Q_{i} \backslash S_{\delta} \leq \gamma_{\bar{\delta}}^{i},\right.\right.
$$

and let us call $\mu_{\bar{\delta}}^{i}, \nu_{\bar{\delta}}^{i}, \mu_{\delta}^{i}$ and $\nu_{\delta}^{i}$ their projections. Up to discard useless cubes, we can assume that each $\gamma_{\frac{i}{\delta}}$ is a non-zero measure. As usual, we apply Proposition 2.11 to each of the measures $\gamma \frac{i}{\delta}$, finding constants $N(i)$ and $\bar{\varepsilon}(i)$, and again we simply call $N=\max \{N(i)\}$ and $\bar{\varepsilon}=\min \{\bar{\varepsilon}(i)\}$. Moreover, up to decrease the value of $\bar{\varepsilon}$, we have

$$
\begin{equation*}
\min \left\{\left\|\gamma_{\bar{\delta}}^{i}\right\|, 1 \leq i \leq C_{1}\right\}>2 \bar{\varepsilon} \tag{4.2}
\end{equation*}
$$

Since $c_{t}$ is highly relativistic, we know that $\kappa_{s}(\rho) \rightarrow \infty$ when $s \rightarrow 0$, as soon as $\rho=r \sqrt{n} / t$ has been chosen small enough. Then Proposition 3.3 ensures us that

$$
\begin{equation*}
\sum_{i=1}^{C_{1}} \varepsilon_{i}:=\sum_{i=1}^{C_{1}}\left\|\gamma_{\delta}^{i}\right\| \leq \frac{\bar{\varepsilon}}{N} \tag{4.3}
\end{equation*}
$$

as soon as $\delta \ll 1$, depending on $\bar{\varepsilon}$ and on $\bar{\delta}$.
For each $1 \leq i \leq C_{1}$, we apply now Proposition 2.11, relative to the measure $\gamma \frac{i}{\delta}$, to the constant $C_{1} \varepsilon_{i}$, and we find as usual measures $\tilde{\gamma}^{i} \leq \gamma$ and $\tilde{\gamma}^{\prime i} \leq \gamma^{\prime}$ with total mass $(N(i)+1) C_{1} \varepsilon_{i}$ satisfying all the properties listed in the proposition. By definition, each measure $\tilde{\gamma}_{\infty}^{i}$ is smaller than $\gamma-\gamma_{\delta}^{i}$, hence it is orthogonal to the measure $\gamma_{\delta}^{i}$; nevertheless, it could have some common
part with $\gamma_{\delta}^{j}$ for some $j \neq i$. As a consequence, we need now to slightly modify each $\gamma_{\delta}^{i}$ : more precisely, we call $\gamma_{\delta, A D}^{i}$ the "already done" part, that is,

$$
\gamma_{\delta, A D}^{i}:=\left(\sum_{j=1}^{C_{1}} \frac{\tilde{\gamma}_{\infty}^{j}}{C_{1}}\right)\left\llcorner Q_{i} \backslash S_{\delta} \leq \gamma_{\delta}^{i},\right.
$$

we define arbitrarily a "replacement part" $\gamma_{\delta, R}^{i}$ satisfying

$$
\begin{equation*}
\left\|\gamma_{\delta, R}^{i}\right\|=\left\|\gamma_{\delta, A D}^{i}\right\|, \quad \quad \gamma_{\delta, R}^{i} \leq\left(\gamma-\sum_{j=1}^{C_{1}} \frac{\tilde{\gamma}_{\infty}^{j}}{C_{1}}\right)\left\llcorner\left(\left(Q_{i} \backslash S_{\bar{\delta}}\right) \cap S_{\delta}\right),\right. \tag{4.4}
\end{equation*}
$$

so that $\gamma_{\delta, R}^{i} \leq \gamma_{\delta}^{i}$ but $\gamma_{\delta, R}^{i} \wedge \gamma_{\delta}^{i}=0$, and we finally set the "modified version" of $\gamma_{\delta}^{i}$ as

$$
\widehat{\gamma}_{\delta}^{i}:=\gamma_{\delta}^{i}-\gamma_{\delta, A D}^{i}+\gamma_{\delta, R}^{i},
$$

calling $\widehat{\mu}_{\delta}^{i}$ and $\widehat{\nu}_{\delta}^{i}$ its projections. Notice that the existence of some replacement part as in (4.4) is ensured by (4.2) and (4.3), and by construction we have that $\left\|\widehat{\gamma}_{\delta}^{i}\right\|=\left\|\gamma_{\delta}^{i}\right\|=\varepsilon_{i}$. We can now set

$$
\xi_{1}:=\gamma-\sum_{i=1}^{C_{1}} \widehat{\gamma}_{\delta}^{i}-\sum_{i=1}^{C_{1}} \frac{\tilde{\gamma}_{\infty}^{i}}{C_{1}}
$$

which is a positive measure by construction (this could have been false if we had used $\gamma_{\delta}^{i}$ instead of $\widehat{\gamma}_{\delta}^{i}$ ). Notice carefully that the measure $\xi_{1}$ is concentrated on $S_{\delta}$. The projections of $\xi_{1}$ are then

$$
\pi_{1} \xi_{1}=\mu-\sum_{i=1}^{C_{1}} \widehat{\mu}_{\delta}^{i}-\sum_{i=1}^{C_{1}} \frac{\tilde{\mu}^{i}}{C_{1}}, \quad \quad \pi_{2} \xi_{1}=\nu-\sum_{i=1}^{C_{1}} \widehat{\nu}_{\delta}^{i}-\sum_{i=1}^{C_{1}} \frac{\tilde{\nu}^{i}}{C_{1}}
$$

and we look now for a measure $\xi_{2}^{i}$ satisfying

$$
\begin{equation*}
\pi_{1} \xi_{2}^{i}=\widehat{\mu}_{\delta}^{i}+\frac{\tilde{\mu}^{i}}{C_{1}}, \quad \quad \pi_{2} \xi_{2}^{i}=\widehat{\nu}_{\delta}^{i}+\frac{\tilde{\nu}^{i}}{C_{1}} \tag{4.5}
\end{equation*}
$$

Notice that the measure $\tilde{\gamma}^{i} / C_{1}$ has projections $\tilde{\mu}^{i} / C_{1}+\mu_{A} / C_{1}$ and $\tilde{\nu}^{i} / C_{1}+\nu_{B} / C_{1}$, and the measures $\mu_{A} / C_{1}$ and $\nu_{B} / C_{1}$ have mass $\varepsilon_{i}$, exactly as $\widehat{\mu}_{\delta}^{i}$ and $\widehat{\nu}_{\delta}^{i}$. Arguing exactly as in the proof of Theorem C, then, we can define $\xi_{2}^{i}=\beta \circ \tilde{\gamma}^{\prime i} / C_{1} \circ \alpha$ so that (4.5) is satisfied. Notice that, exactly as in Theorem C and keeping in mind (4.1), the measure $\xi_{2}^{i}$ is concentrated on $S_{\delta}$ : since the same was already observed for $\xi_{1}$, we derive that the plan $\xi=\xi_{1}+\sum_{i} \xi_{2}^{i}$ belongs to $\Pi(\mu, \nu)$, and it has a finite $\mathcal{C}_{t-\delta t}$ cost. More precisely, calling again $C=\max \{h(v): v \in \mathscr{C}\}$, we have immediately that

$$
\sum_{i=1}^{C_{1}} \mathcal{C}_{t-\delta t}\left(\xi_{2}^{i}\right) \leq C \sum_{i=1}^{C_{1}}\left\|\xi_{2}^{i}\right\|=C \sum_{i=1}^{C_{1}} N(i) \varepsilon_{i} \leq C N \sum_{i=1}^{C_{1}} \varepsilon_{i}
$$

Concerning $\xi_{1}$, we have

$$
\begin{aligned}
\mathscr{C}_{t-\delta t}\left(\xi_{1}\right) & \leq \mathscr{C}_{t-\delta t}\left(\gamma-\sum_{i} \gamma_{\delta}^{i}\right)=\iint_{S_{\delta}} c_{t-\delta t}(x, y) d \gamma(x, y)=\iint_{S_{\delta}} h\left(\frac{y-x}{t(1-\delta)}\right) d \gamma(x, y) \\
& \leq \iint_{S_{\delta}}(1+\operatorname{exc}(\delta)) h\left(\frac{y-x}{t}\right) d \gamma(x, y) \leq(1+\operatorname{exc}(\delta)) \mathcal{C}_{t}(\gamma)=(1+\operatorname{exc}(\delta)) \mathcal{C}(t)
\end{aligned}
$$

Putting together the last two estimates and applying Proposition 3.3 we then find

$$
\begin{aligned}
\mathcal{C}(t-\delta t) & \leq \mathcal{C}_{t-\delta t}(\xi) \leq(1+\operatorname{exc}(\delta)) \mathcal{C}(t)+C N \sum_{i} \varepsilon_{i} \\
& =(1+\operatorname{exc}(\delta)) \mathcal{C}(t)+C N \gamma\left(\mathbb{R}^{2 n} \backslash S_{\delta}\right) \leq(1+\operatorname{exc}(\delta)) \mathcal{C}(t)+\frac{C N \bar{C}}{\kappa_{\delta}(\rho)} .
\end{aligned}
$$

This gives (1.3) which in turn, since $\kappa_{\delta}(\rho) \rightarrow \infty$ for $\delta \searrow 0$, also implies the continuity of the map $t \mapsto \mathcal{C}(t)$.

We can conclude this section by showing a simple situation, which gives counterexamples to all our previos results without their assumptions, and then shows that basically all our assumptions were sharp.


Figure 1. Situation in Example 4.2

Example 4.2. Let us consider the situation drawn in Figure 1, where $\mu$ (resp., $\nu$ ) is the sum of three Dirac masses of weight $1 / 3$ at the points $x_{1}, x_{2}$ and $x_{3}$ (resp., $y_{1}, y_{2}$ and $y_{3}$ ). Assume that, as in the figure, the distance between $x_{1}$ and $y_{1}$ is 1 , the distances between $x_{1}$ and $y_{2}, x_{2}$ and $y_{3}$, and $x_{3}$ and $y_{1}$ are all 0.9 , and the distances between $x_{2}$ and $y_{2}$ and between $x_{3}$ and $y_{3}$ are some small $\varepsilon \ll 1$. Let $h$ be a relativistic cost function such that $h(v)=f(|v|)$, with $f(0)=0, f(\varepsilon) \approx \varepsilon^{2}, f(1)=1$ and $f(r)=+\infty$ for $r>1$. We can assume that

$$
\begin{equation*}
f(1)+2 f(\varepsilon)<3 f(0.9) . \tag{4.6}
\end{equation*}
$$

It is clear that there are basically two possible ways of transporting $\mu$ onto $\nu$ : either sending each $x_{i}$ onto the corresponding $y_{i}$, or sending each $x_{i}$ onto $y_{i+1}$, where we denote for simplicity $y_{4} \equiv y_{1}$. These two possibilities correspond to paying once the price of moving of a distance 1 and twice of a distance $\varepsilon$, or to paying thrice the price of moving of a distance 0.9. It is obvious that $T=0.9$, and that for $T \leq t<1$ the second transport plan is the unique having a finite cost; instead, for $t \geq 1$ both the transports (and all their convex combinations) have finite cost; by keeping in mind (4.6), this gives that the total cost is given by

$$
\mathcal{C}(t)= \begin{cases}f(0.9 / t) & \text { if } 0.9 \leq t<1 \\ \frac{1}{3} f(1 / t)+\frac{2}{3} f(\varepsilon / t) & \text { if } t \geq 1\end{cases}
$$

Hence, the total cost function $t \mapsto \mathcal{C}(t)$ can be discontinuous if $\mu$ is a singular measure, and this gives a counterexample to Theorem A without the assumption that $\mu \ll \mathscr{L}^{n}$ (while the additional fact that $\mu$ is in $L^{\infty}$ is basically only needed to get the nice estimate (1.3)). The same example gives also a counterexample to Theorem $C$ without the assumption $\mu \ll \mathscr{L}^{n}$ : indeed, if we take $t=1$, then $t>T=0.9$, and nevertheless there is a strictly positive mass moved exactly with directions in t $\partial \mathscr{C}$ by the optimal plan relative to $\mathcal{C}_{t}$.

Finally, it is easy to show that a simple modification of the same example gives also a counterexample to the validity of Theorem C for costs which are relativistic but not highly relativistic. Indeed, substitute the Dirac masses in the example above with uniform measures concentrated in very small balls, the centers still being the points $x_{i}$ and $y_{i}$. It is then still true that $T=0.9$, and we want to show that for $t=1$ (and with a good choice of f) the tranport plan $\gamma_{1}$ translating each ball centered at $x_{i}$ onto the ball centered at $y_{i}$ is still optimal. If we do so, then again we have a counterexample to Theorem $C$ for $\mu$ absolutely continuous but with relativistic and not highly relativistic cost functions. We just underline how can one prove this: arguing as in the proof of Theorem C, if another transport plan $\tilde{\gamma}$ moves a portion $1-\delta$ of the points of the ball at $x_{1}$ onto the ball at $y_{1}$ by making a distance $1-\delta$ instead of 1 , then the remaining portion $\delta$ of the ball at $x_{1}$ must move of a distance 0.9 onto the ball at $y_{2}$, then a portion $\delta$ of the ball at $x_{2}$ must go onto the ball at $y_{3}$, and finally a portion $\delta$ of the ball at $x_{3}$ must fill the remaining portion $\delta$ of the ball at $y_{1}$. The costs of the plans $\gamma_{1}$ and $\tilde{\gamma}$ are then

$$
3 \mathcal{C}_{1}\left(\gamma_{1}\right)=f(1)+2 f(\varepsilon), \quad 3 \mathcal{C}_{1}(\tilde{\gamma}) \approx(1-\delta)(f(1-\delta)+2 f(\varepsilon))+3 \delta f(0.9)
$$

A simple calculation, then, shows that basically $\tilde{\gamma}$ is better than $\gamma_{1}$ if

$$
f^{\prime}(1)>3 f(0.9)-f(1)
$$

(recall that the right term is positive by (4.6)). Now, if the cost is highly relativistic, then this means $f^{\prime}(1)=+\infty$, and thus the above inequality is surely true and then $\gamma_{1}$ is not optimal, according with Theorem C. Instead, for a cost which is relativistic but not highly relativistic, it is possible that the above inequality fails, which implies the optimality of $\gamma_{1}$, and thus the claim of Theorem $C$ is seen to be false without the highly relativistic assumption.

Remark 4.3. Roughly speaking, both Theorems $C$ and $A$ have two main assumptions, namely, the fact that $c_{t}$ is highly relativistic, and the fact that $\mu \ll \mathscr{L}^{n}$. The above Example 4.2 has shown that the absolute continuity assumption on $\mu$ is essential for both the results, and that the highly relativistic assumption on $c_{t}$ is essential for Theorem C. Instead, we do not have an example showing that the plain relativistic assumption is not enough to get the continuity of the total cost function.

## 5. Existence of Kantorovich potentials

The goal of this section is to prove the existence of a Kantorovich potential, namely, Theorem B. To do so, we need some classical notation, which we briefly present here. The first definition is the $c$-cyclical monotonicity, which was introduced by Knott and Smith in [26, 27],
generalizing the classical cyclical monotonicity defined by Rockafellar in [31], and later used by several authors.

Definition 5.1 (c-cyclically monotone set). $A$ set $S \subseteq \mathbb{R}^{n} \times \mathbb{R}^{n}$ is said c-cyclically monotone if for any $m \in \mathbb{N}$ and any choice of $\left(x_{i}, y_{i}\right) \in S, 0 \leq i \leq m$, one has

$$
\sum_{i=0}^{m} c\left(x_{i}, y_{i}\right) \leq \sum_{i=0}^{m} c\left(x_{i}, y_{i+1}\right)
$$

as usual denoting $y_{m+1}=y_{0}$.
It is extremely simple to show that, whenever a mass transport problem with a l.s.c. cost $c$ is considered, if the total cost is finite then any optimal transport plan is concentrated on a $c$-cyclical monotone set. The second notation that we will use is that of $c$-transform, $c$-concavity and $c$-subdifferential.

Definition 5.2 ( $c$-transform, $c$-concave function and $c$-subdifferential). Let $\varphi: \operatorname{supp} \mu \rightarrow \mathbb{R} \cup$ $\{-\infty\}$ (resp., $\psi: \operatorname{supp} \nu \rightarrow \mathbb{R} \cup\{-\infty\}$ ) be a Borel function; its $c$-transform $\varphi^{c}: \operatorname{supp} \nu \rightarrow$ $\mathbb{R} \cup\{ \pm \infty\}$ (resp., $\psi^{c}: \operatorname{supp} \mu \rightarrow \mathbb{R} \cup\{ \pm \infty\}$ ) is defined as

$$
\varphi^{c}(y):=\inf _{x \in \operatorname{supp} \mu} c(x, y)-\varphi(x), \quad \quad \psi^{c}(x):=\inf _{y \in \operatorname{supp} \nu} c(x, y)-\psi(y)
$$

The function $\varphi$ is said $c$-concave if $\varphi^{c}(y)<+\infty$ for every $y \in \operatorname{supp} \nu$ and $\varphi^{c c}=\varphi$. Finally, for every $c$-concave function $\varphi$, we define the $c$-subdifferential as the set

$$
\partial_{c} \varphi=\left\{(x, y) \in \operatorname{supp} \mu \times \operatorname{supp} \nu: \varphi(x)+\varphi^{c}(y)=c(x, y)\right\} .
$$

It is well-known that the $c$-subdifferential of any $c$-concave function is a $c$-monotone set, see for instance [34]. We can now recall the definition of Kantorovich potential, and that of the approximate gradient.

Definition 5.3 (Kantorovich potential). A Kantorovich potential is a c-concave map $\varphi$ : $\operatorname{supp} \mu \rightarrow \mathbb{R} \cup\{-\infty\}$ with the property that any optimal transport plan $\gamma \in \Pi(\mu, \nu)$ is concentrated on $\partial_{c} \varphi$.

Definition 5.4 (Approximate gradient). Given a function $f: X \subseteq \mathbb{R}^{n} \rightarrow \mathbb{R} \cup\{-\infty\}$ and a point $x$ of density 1 in $X$, we say that $v \in \mathbb{R}^{n}$ is the approximate gradient of $f$ at $x$ if there exists a map $g: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ differentiable at $x$, with $\nabla g(x)=v$ and such that $x$ has density 1 in the set $\{f=g\}$. If there exists such a vector, then it is necessarily unique, and we denote it by $\widetilde{\nabla} f(x)$; moreover, the function $f$ is said to be approximately differentiable at $x$.

The existence of a Kantorovich potential is a fundamental issue in the mass transportation problem, also because it is an important aid in showing the existence of an optimal transport map. It is classically known that a Kantorovich potential exists in the most common cases of transport problems, though not always.

To show Theorem B, we start with another "chain" result, which is a sort of discrete version of the Chain Lemma 2.10.

Lemma 5.5. Under the same assumptions of Theorem B, there exists $M \in \mathbb{N}$ such that the following holds. Let $\gamma_{t}$ be an optimal transport plan for the cost $c_{t}$, let $\Gamma \subseteq \operatorname{supp} \gamma_{t}$ be a set on which $\gamma_{t}$ is concentrated, and fix $\left(x_{0}, y_{0}\right) \in \operatorname{supp} \gamma_{t}$. Then, for every $(x, y) \in \operatorname{supp} \gamma_{t}$, there is some $0 \leq k \leq M$ and points $\left(x_{i}, y_{i}\right) \in \Gamma$ for $0<i \leq k$ such that, calling $\left(x_{k+1}, y_{k+1}\right)=(x, y)$, for every $0 \leq i \leq k$ one has $y_{i}-x_{i+1} \in t \mathscr{C}$.

Proof. The proof is somehow similar to that of the chain Lemma 2.10, but in a discretized setting. For simplicity, we divide it in some steps.
Step I. The discretized setting and the cycles.
We start by selecting some $T<t^{\prime}<t^{\prime \prime}<t$ and a constant $\varepsilon>0$, much smaller than $t-t^{\prime \prime}, t^{\prime \prime}-t^{\prime}$, and $t^{\prime}-T$, and we cover $\operatorname{supp} \mu($ resp., $\operatorname{supp} \nu)$ with finitely many essentially disjoint cubes $Q_{j}^{1}$ (resp., $Q_{k}^{2}$ ) of side length $\varepsilon>0$. Let us call $N+1$ the number of the pairs $(j, k)$ such that $\gamma_{t}\left(Q_{j}^{1} \times Q_{k}^{2}\right)>0$, and let us enumerate these cubes $Q_{j}^{1} \times Q_{k}^{2}$ as $Q_{0}, Q_{1}, Q_{2}, \ldots, Q_{N} \subseteq \mathbb{R}^{n} \times \mathbb{R}^{n}$. Notice that the number of cubes can be bounded only in terms of the supports of $\mu$ and $\nu$, and of the value of $\varepsilon$, thus in turn only in terms of $\mu, \nu$ and $t$; hence, there is some $M \in \mathbb{N}$, depending only on $\mu, \nu$ and $t$, such that $N \leq M$. For every $0 \leq i \leq N$, writing $Q_{i}=Q_{j}^{1} \times Q_{k}^{2}$ we call $z_{i}$ and $w_{i}$ the centers of $Q_{j}^{1}$ and $Q_{k}^{2}$. Notice that the pairs $\left(z_{i}, w_{i}\right)$ are all different, but of course $z_{i}=z_{m}$ (resp., $w_{i}=w_{m}$ ) whenever the first (resp., the second) projection of the $2 n$-dimensional cubes $Q_{i}$ and $Q_{m}$ is the same $n$-dimensional cube.

Let us now set the "discretized problem" as follows: we define

$$
\gamma_{d}=\sum_{i=0}^{N} \delta_{\left(z_{i}, w_{i}\right)} \gamma_{t}\left(Q_{i}\right), \quad \quad \mu_{d}=\pi_{1} \gamma_{d}, \quad \nu_{d}=\pi_{2} \gamma_{d}
$$

and we observe that the critical speed $T_{d}$ of the discretized problem corresponding to $\mu_{d}$ and $\nu_{d}$ is at most $t^{\prime}$. Indeed, for every $j, k$ let us call

$$
\alpha_{j}=\mu_{d}\left\llcornerQ _ { j } ^ { 1 } \otimes \mu \left\llcorner Q_{j}^{1}, \quad \quad \beta_{k}=\nu\left\llcornerQ _ { k } ^ { 2 } \otimes \nu _ { d } \left\llcorner Q_{k}^{2}\right.\right.\right.\right.
$$

so that $\alpha=\sum \alpha_{j}$ (resp., $\beta=\sum \beta_{k}$ ) is a transport plan between $\mu_{d}$ and $\mu$ (resp., $\nu$ and $\left.\nu_{d}\right)$; by construction, for every $\left(x^{\prime}, x^{\prime \prime}\right) \in \operatorname{supp} \alpha$ (resp., for every $\left(y^{\prime}, y^{\prime \prime}\right) \in \operatorname{supp} \beta$ ) one has $\left|x^{\prime \prime}-x^{\prime}\right| \leq \varepsilon \sqrt{n}$ (resp., $\left|y^{\prime}-y^{\prime \prime}\right| \leq \varepsilon \sqrt{n}$ ). As a consequence, if we call $\gamma_{T}$ an optimal transport plan corresponding to $c_{T}$, we have that $\gamma^{\prime}:=\beta \circ \gamma_{T} \circ \alpha$ is a transport plan between $\mu_{d}$ and $\nu_{d}$, and for every $(x, y) \in \operatorname{supp} \gamma^{\prime}$ one has $y-x \in T \mathscr{C}+\mathcal{B}(2 \varepsilon \sqrt{n}) \subseteq t^{\prime} \mathscr{C}$, where $\mathcal{B}(r)$ denotes the ball of radius $r$ centered at the origin, and the last inclusion holds as soon as $\varepsilon$ is small enough.

Let us now take any two indices $0 \leq i, i^{\prime} \leq N$ : we say that $i^{\prime}$ follows $i$ if $\left(z_{i^{\prime}}, w_{i}\right) \in \operatorname{supp} \gamma^{\prime}$; notice that for every index $i$ there is at least one $i^{\prime}$ which follows $i$, possibly $i$ itself. Then, we call cycle any finite sequence of not necessarily distinct indices $I=\left\{i_{1}, i_{2}, \ldots, i_{p}\right\}$ such that every $i_{m+1}$ follows $i_{m}$, and $i_{1}$ follows $i_{p}$; in particular, if $i$ follows itself, then the singleton $I=\{i\}$ is a cycle.
Step II. Every index $0 \leq i \leq N$ belongs to at least a cycle.
For every $0 \leq i \leq N$, we can start a sequence of indices from $i$, so that each one follows the preceding one, until a cycle is found; notice that this cycle does not necessarily contain $i$ (otherwise, the claim would be trivial); notice also that a cycle might contain twice the same
index, but any cycle contains a subcycle with all distinct indices. Take then an arbitrary cycle $I=\left\{i_{1}, i_{2}, \ldots, i_{p}\right\}$, containing all distinct indices, and set

$$
\eta=\min _{\ell=1, \ldots, p} \gamma_{d}\left(\left\{\left(z_{i_{\ell}}, w_{i_{\ell}}\right)\right\}\right) \wedge \min _{\ell=1, \ldots, p} \gamma^{\prime}\left(\left\{\left(z_{i_{\ell+1}}, w_{i_{\ell}}\right)\right\}\right),
$$

considering as usual $i_{p+1}=i_{1}$; notice that $\eta$ is strictly positive by construction. Now, define the "reduced transport plans"

$$
\gamma_{\mathrm{red}}=\gamma_{d}-\eta \sum_{\ell=1}^{p} \delta_{\left(z_{i_{\ell}}, w_{i_{\ell}}\right)}, \quad \quad \gamma_{\mathrm{red}}^{\prime}=\gamma^{\prime}-\eta \sum_{\ell=1}^{p} \delta_{\left(z_{i_{\ell+1}}, w_{i_{\ell}}\right)} .
$$

Observe that these are two positive measures with the same marginals; observe also that the number of the pairs in the support of $\gamma_{\mathrm{red}}$ and $\gamma_{\text {red }}^{\prime}$ is at most the corresponding number for $\gamma_{d}$ and $\gamma^{\prime}$, and at least one of them has strictly decreased. We can repeat the same construction with $\gamma_{\text {red }}$ and $\gamma_{\text {red }}^{\prime}$ in place of $\gamma_{d}$ and $\gamma^{\prime}$, so to find again a cycle: notice that, since $\gamma_{\text {red }} \leq \gamma_{d}$ and $\gamma_{\text {red }}^{\prime} \leq \gamma^{\prime}$, then any cycle with respect to $\gamma_{\text {red }}$ and $\gamma_{\text {red }}^{\prime}$ is a fortiori a cycle with respect to $\gamma_{d}$ and $\gamma^{\prime}$. Going on in the same way, after at most $2 M+2$ steps the reduced transport plans become null: this means that every index belongs to some of the cycles which have been found, so this step is concluded.
Step III. The "regions" and the "networks".
For any index $0 \leq i \leq N$, we call region the union $\mathcal{R}$ of all the indices which belong to some cycle containing $i$. It is clear that for any two elements $m, p \in \mathcal{R}$ there is a sequence of indices, each following the preceding one, which starts with $m$ and ends in $p$; it is also clear that, if we had started with some $m \in \mathcal{R}$, instead than with $i$, we would have found the same region. In other words, the set of all the indices is subdivided in finitely many disjoint regions. Take now two indices $i$ and $m$ such that $z_{i}=z_{m}$; then, $m$ and $i$ belong to the same region: indeed, for any index $p$ one has that $i$ follows $p$ if and only if $m$ follows $p$, and then we can trivially use two cycles, one containing $i$ and the other containing $m$, to build another cycle containing both $i$ and $m$. As a consequences, the first projections of the regions form also a disjoint cover of $\operatorname{supp} \mu_{d}$.

Let us now introduce another piece of notation: given two indices $m, p$ we say that $m$ can be hooked to $p$ if $w_{p}-z_{m} \in t^{\prime \prime} \mathscr{C}$. By definition, if $m$ follows $p$ then it is also true that $m$ can be hooked to $p$, but the converse is not necessarily true. Moreover, we say that the region $\mathcal{R}^{\prime}$ can be hooked to the region $\mathcal{R}$ if there exist two indices $m \in \mathcal{R}^{\prime}, p \in \mathcal{R}$ such that $m$ can be hooked to $p$. Finally, for any region $\mathcal{R}$ we call its network the union of all the regions $\mathcal{R}^{\prime}$ such that there is a finite sequence of regions, each one hooked to the preceding one, starting with $\mathcal{R}$ and ending with $\mathcal{R}^{\prime}$. Notice that, in principle, the fact that $\mathcal{R}^{\prime}$ can be hooked to $\mathcal{R}$ does not imply that $\mathcal{R}$ can be hooked to $\mathcal{R}^{\prime}$; more in general, if the network of $\mathcal{R}$ contains $\mathcal{R}^{\prime}$, in principle it is possible that the network of $\mathcal{R}^{\prime}$ does not contain $\mathcal{R}$ : in other words, the networks are not automatically a partition of the indices. Nevertheless, in the next step we see that something much stronger is actually true.
Step IV. Every network contains all the regions.
Let us fix a region $\mathcal{R}$, and call $\mathcal{N}$ the network of $\mathcal{R}$ : we aim to prove that $\mathcal{N}$ contains all the
regions. Since we already observed, in Step III, that whenever $z_{i}=z_{m}$ the two indices $i$ and $m$ are in the same region, it is enough to show that the union $\mathcal{N}_{1}$ of all the cubes centered at some $z_{i}$ with $i \in \mathcal{N}$ contains the whole support of $\mu_{d}$.

Suppose that it is not so: this means that the measure $\mu$ is not concentrated on $\mathcal{N}_{1}$. Since the support of $\mu$ is connected, there must be some cube $Q_{j}^{1}$ which does not belong to $\mathcal{N}_{1}$ but is adjacent to one of the cubes of $\mathcal{N}_{1}$, say $Q_{l}^{1}$, such that $\mu\left(Q_{j}^{1}\right)>0$. Assuming for simplicity of notations that $Q_{j}^{1}$ and $Q_{l}^{1}$ are contained in the $2 n$-dimensional cubes $Q_{j}$ and $Q_{l}$, we have then that $\left|z_{j}-z_{l}\right| \leq \varepsilon \sqrt{n}$; recall that by construction we have $l \in \mathcal{N}$ and $j \notin \mathcal{N}$. Let then $p$ be an index such that $l$ follows $p$ : this implies that $\left(z_{l}, w_{p}\right) \in \operatorname{supp} \gamma^{\prime}$, hence $w_{p}-z_{l} \in t^{\prime} \mathscr{C}$, so in turn $w_{p}-z_{j} \in t^{\prime} \mathscr{C}+\mathcal{B}(\varepsilon \sqrt{n}) \subseteq t^{\prime \prime} \mathscr{C}$, where again the last inclusion holds if $\varepsilon$ has been chosen small enough. Since this means that $j$ can be hooked to $p$, and since the fact that $l$ follows $p$ implies that $l$ and $p$ are in the same region, thus also $p$ belongs to $\mathcal{N}$, we derive that the index $j$ is an element of $\mathcal{N}$, which gives the desired contradiction. Hence, the network of each region is made of all the indices, and this conlcudes the step.
Step V. Conclusion.
We are finally in position to conclude the proof of this lemma. Fix a pair $\left(x_{0}, y_{0}\right)$ in $\operatorname{supp} \gamma_{t}$, and take any other pair $(x, y) \in \operatorname{supp} \gamma_{t}$. Thanks to Step IV, we can find $k \geq 0$ and indices $j(0), j(1), \ldots, j(k+1)$ such that $\left(x_{0}, y_{0}\right) \in Q_{j_{0}},(x, y) \in Q_{j_{k+1}}$, and for every $0 \leq i \leq k$ the index $j_{i+1}$ can be hooked to $j_{i}$; of course, we can do this in such a way that all the indices $j_{i}$ are distinct, except for $j_{0}$ and $j_{k+1}$ which could coincide. As a consequence, since there exist $N+1$ indices, for sure we have $k \leq N$, hence $k \leq M$ because we know that $N \leq M$. Now, set $x_{k+1}=x$ and $y_{k+1}=y$, and for every $0<i \leq k$ choose arbitrarily some pair $\left(x_{i}, y_{i}\right) \in \Gamma \cap Q_{i}$ : this is possible because, by construction, $\gamma_{t}\left(Q_{i}\right)>0$, and $\gamma_{t}$ is concentrated on $\Gamma$. We claim that the sequence $\left(x_{i}, y_{i}\right)$ for $0<i \leq k$ fulfills the requirements of the Lemma.

Indeed, the fact that every pair $\left(x_{i}, y_{i}\right)$ with $0<i \leq k$ belongs to $\Gamma$ is true by construction. Moreover, for every $i$ we have that $x_{i}$ (resp., $y_{i}$ ) is in a cube of side $\varepsilon$ centered at $z_{i}$ (resp., $w_{i}$ ). As a consequence, for every $0 \leq i \leq k$ we have

$$
y_{i}-x_{i+1} \in w_{i}-z_{i+1}+\mathcal{B}(2 \varepsilon \sqrt{n}) \subseteq t^{\prime \prime} \mathscr{C}+\mathcal{B}(2 \varepsilon \sqrt{n}) \subseteq t \mathscr{C}
$$

where as usual the last inclusion is true if $\varepsilon$ was chosen small enough.
The next step to prove Theorem B is to show the existence of a Kantorovich potential for $c_{t}$ whenever $t>T$. This is the content of the next result. We underline that the same result was proven in [12] only for almost every $t>T$, and under additional assumptions on the measure $\mu$.

Proposition 5.6. Under the same assumptions of Theorem B, for any supercritical speed $t>T$ there exists a Kantorovich potential $\varphi_{t}$ for $c_{t}$, which is approximately differentiable $\mu$-almost everywhere.

Proof. Let us fix a supercritical speed $t>T$, and let us assume without loss of generality -and just for the simplicity of notations- that $t=1$. Let us call $\gamma$ an optimal transport plan relative to the cost $c=c_{1}$, let $\Gamma \subseteq \operatorname{supp} \gamma$ be a $c$-cyclical monotone set on which $\gamma$ is concentrated, and
let us fix arbitrarily a point $\left(x_{0}, y_{0}\right) \in \Gamma$ : since the time $t=1$ is supercritical and $\gamma$ is an optimal transport plan, we know that $y_{0}-x_{0} \in \mathscr{C}$. Let us then define the map $\varphi: \operatorname{supp} \mu \rightarrow \overline{\mathbb{R}}$ as

$$
\begin{equation*}
\varphi(x)=\inf _{k \in \mathbb{N}} \inf \left\{\sum_{i=1}^{k} c\left(x_{i}, y_{i-1}\right)+c\left(x, y_{k}\right)-\sum_{i=0}^{k} c\left(x_{i}, y_{i}\right):\left(x_{i}, y_{i}\right)_{i=1}^{k} \in \Gamma^{k}\right\} . \tag{5.1}
\end{equation*}
$$

The proof will by concluded once we show that this map is a $\mu$-almost everywhere approximately differentiable Kantorovich potential.
Step I. The map $\varphi$ is well-defined and bounded from above and from below.
First of all notice that, since the set $\left\{(x, y) \in \mathbb{R}^{n} \times \mathbb{R}^{n}: c(x, y)=+\infty\right\}$ is open, then $c(x, y)<$ $+\infty$ for every $(x, y) \in \operatorname{supp} \gamma$, hence in particular for every $(x, y) \in \Gamma$. As a consequence, for every $k \in \mathbb{N}$ and every choice of the sequence $\left(x_{i}, y_{i}\right)_{i=1}^{k} \in \Gamma^{k}$, the second sum in the braces above is real valued, while the first sum is possibly $+\infty$. Hence, the term in braces is well defined for every sequence and then the function $\varphi$ is well-defined (and easily Borel).

We aim now to show that $\varphi$ is bounded from above and from below. First of all, take any $x \in \operatorname{supp} \mu$, and let $y$ be a point such that $(x, y) \in \operatorname{supp} \gamma$. By Lemma 5.5, we can find a finite chain $\left(x_{i}, y_{i}\right)_{i=1}^{k} \in \Gamma^{k}$ such that, writing $x_{k+1}=x$, for every $0 \leq i \leq k$ one has $y_{i}-x_{i+1} \in \mathscr{C}$, hence $c\left(x_{i+1}, y_{i}\right)<\|h\|_{L^{\infty}}$. This implies that $\varphi(x)<(k+1)\|h\|_{L^{\infty}}$, and since $x$ was arbitrary and Lemma 5.5 gives $k \leq M$ for some $M$ depending only on $\mu, \nu$ and $t$, we obtain that $\varphi \leq(M+1)\|h\|_{L^{\infty}}$, so the upper bound is found.

We now deal with the lower bound. Take any $x \in \operatorname{supp} \mu$, any $k \in \mathbb{N}$, and any choice of pairs $\left(x_{i}, y_{i}\right)_{i=1}^{k} \in \Gamma^{k}$. Let us apply Lemma 5.5 with starting point $\left(x_{k}, y_{k}\right)$ and final point $\left(x_{0}, y_{0}\right)$. We get some $0 \leq \ell \leq M$ and points $\left(x_{i}, y_{i}\right) \in \Gamma$ for $k<i \leq k+\ell$ such that $t \mathscr{C}$ contains $y_{i-1}-x_{i}$ for every $k<i \leq k+\ell$, as well as $y_{k+\ell}-x_{0}$. As a consequence $c\left(x_{i}, y_{i-1}\right)$ is bounded by $\|h\|_{L^{\infty}}$ for every $k<i \leq k+\ell$, as well as $c\left(x_{0}, y_{k+\ell}\right)$. Therefore, we can estimate

$$
\begin{aligned}
& \sum_{i=1}^{k} c\left(x_{i}, y_{i-1}\right)+c\left(x, y_{k}\right)-\sum_{i=0}^{k} c\left(x_{i}, y_{i}\right) \\
& \quad \geq \sum_{i=1}^{k+\ell} c\left(x_{i}, y_{i-1}\right)-\sum_{i=0}^{k+\ell} c\left(x_{i}, y_{i}\right)-\sum_{i=k+1}^{k+\ell} c\left(x_{i}, y_{i-1}\right)+c\left(x_{0}, y_{k+\ell}\right)-c\left(x_{0}, y_{k+\ell}\right) \\
& \quad \geq \sum_{i=1}^{k+\ell} c\left(x_{i}, y_{i-1}\right)+c\left(x_{0}, y_{k+\ell}\right)-\sum_{i=0}^{k+\ell} c\left(x_{i}, y_{i}\right)-(M+1)\|h\|_{L^{\infty}} \geq-(M+1)\|h\|_{L^{\infty}}
\end{aligned}
$$

where the last inequality comes because all the points $\left(x_{i}, y_{i}\right)$ are in the $c$-cyclically monotone set $\Gamma$. Keeping in mind the definition (5.1) of $\varphi$ and the arbitrariness of $x \in \operatorname{supp} \mu$ and of the sequence $\left(x_{i}, y_{i}\right)_{i=1}^{k}$, we obtain that $\varphi \geq-(M+1)\|h\|_{L^{\infty}}$, so also the lower bound is found. Step II. $\varphi$ is c-concave.
In this step we want to show that $\varphi$ is $c$-concave, that is, $\varphi^{c}(y)<+\infty$ for every $y \in \operatorname{supp} \nu$, and $\varphi^{c c}=\varphi$. First of all, take any $y \in \operatorname{supp} \nu$ and let $x \in \operatorname{supp} \mu$ be such that $(x, y) \in \operatorname{supp} \gamma$, hence $c(x, y) \leq\|h\|_{L^{\infty}}$. Thus, also by the estimates of Step I,

$$
\varphi^{c}(y) \leq c(x, y)-\varphi(x) \leq(M+2)\|h\|_{L^{\infty}} .
$$

We have then found that $\varphi^{c}$ is bounded from above, so the function $\varphi^{c c}$ is well defined, and we have to check that $\varphi^{c c}=\varphi$. Actually, since the inequality $\varphi^{c c} \geq \varphi$ is always trivially true, we just have to take care of the opposite inequality. Let us now define the function

$$
\begin{equation*}
\psi(y)=\sup _{k \in \mathbb{N}} \sup \left\{\sum_{i=0}^{k} c\left(x_{i}, y_{i}\right)-\sum_{i=1}^{k} c\left(x_{i}, y_{i-1}\right):\left(x_{i}, y_{i}\right)_{i=1}^{k} \in \Gamma^{k}, y_{k}=y\right\} \tag{5.2}
\end{equation*}
$$

which is well-defined and Borel by the very same argument as for $\varphi$ in Step I. We claim that $\psi(y)<+\infty$ for every $y \in \operatorname{supp} \nu$. If this is the case, then the $c$-transform $\psi^{c}$ is well-defined, and comparing (5.2) with (5.1) it is immediate to realize that $\psi^{c}=\varphi$. Since we know that $\varphi=\psi^{c}$ is bounded, we obtain then that $\psi \leq \psi^{c c}=\varphi^{c}$, and this implies that $\psi^{c} \geq \varphi^{c c}$, that is, the inequality $\varphi \geq \varphi^{c c}$ follows. Summarizing, to prove that $\varphi$ is $c$-concave we just have to check that $\psi(y)<+\infty$ for every $y \in \operatorname{supp} \nu$.

We can immediately notice that $\psi\left(y_{0}\right)=c\left(x_{0}, y_{0}\right)$ : indeed, the inequality $\psi\left(y_{0}\right) \geq c\left(x_{0}, y_{0}\right)$ comes simply by choosing $k=1$ and $\left(x_{1}, y_{1}\right)=\left(x_{0}, y_{0}\right)$ in (5.2); on the other hand, for every sequence $\left(x_{i}, y_{i}\right)_{i=1}^{k}$ in $\Gamma^{k}$ with $y_{k}=y_{0}$, one has

$$
\sum_{i=0}^{k} c\left(x_{i}, y_{i}\right)-\sum_{i=1}^{k} c\left(x_{i}, y_{i-1}\right)=\sum_{i=0}^{k-1} c\left(x_{i}, y_{i}\right)+c\left(x_{k}, y_{0}\right)-\sum_{i=1}^{k} c\left(x_{i}, y_{i-1}\right) \leq c\left(x_{0}, y_{0}\right)
$$

by the $c$-cyclical monotonicity of $\Gamma$, and then we get $\psi\left(y_{0}\right) \leq c\left(x_{0}, y_{0}\right)$ directly by the definition.
Let us now take any $y \in \operatorname{supp} \nu$ : we must show that $\psi(y)<+\infty$. If $y$ does not belong to the second projection $\pi_{2}(\Gamma)$ of $\Gamma$, then (5.2) readily gives $\psi(y)=-\infty$, so there is nothing to prove. Otherwise, let $x \in \operatorname{supp} \mu$ be such that $(x, y) \in \Gamma$. First of all, consider any other $(\tilde{x}, \tilde{y}) \in \Gamma$ : for each sequence $\left(x_{i}, y_{i}\right)_{i=1}^{k} \in \Gamma^{k}$ with $y_{k}=y$, we can then add $\left(x_{k+1}, y_{k+1}\right)=(\tilde{x}, \tilde{y})$ to get a sequence of length $k+1$ which is admissible in the definition of $\psi(\tilde{y})$, and we then $\psi(\tilde{y}) \geq \psi(y)+c(\tilde{x}, \tilde{y})-c(\tilde{x}, y)$, which of course implies, for any choice of $(\tilde{x}, \tilde{y}) \in \Gamma$, the inequality

$$
\begin{equation*}
\psi(y) \leq \psi(\tilde{y})+c(\tilde{x}, y)-c(\tilde{x}, \tilde{y}) \leq \psi(\tilde{y})+c(\tilde{x}, y) \tag{5.3}
\end{equation*}
$$

Let us now apply once again Lemma 5.5 , starting from $(x, y)$ and reaching ( $x_{0}, y_{0}$ ). We get some $0 \leq k \leq M$ and point $\left(x_{i}, y_{i}\right)$ in $\Gamma$ for $1 \leq i \leq k$ in such a way that $\mathscr{C}$ contains all the vectors

$$
y-x_{1}, y_{1}-x_{2}, \ldots, y_{k-1}-x_{k}, y_{k}-x_{0} .
$$

Therefore, we can apply several times (5.3) to get

$$
\begin{aligned}
\psi(y) & \leq \psi\left(y_{1}\right)+c\left(x_{1}, y\right) \leq \psi\left(y_{2}\right)+c\left(x_{2}, y_{1}\right)+c\left(x_{1}, y\right) \leq \cdots \\
& \leq \psi\left(y_{k}\right)+c\left(x_{k}, y_{k-1}\right)+\cdots+c\left(x_{1}, y\right) \leq \psi\left(y_{0}\right)+c\left(x_{0}, y_{k}\right)+c\left(x_{k}, y_{k-1}\right)+\cdots+c\left(x_{1}, y\right) \\
& \leq c\left(x_{0}, y_{0}\right)+(M+1)\|h\|_{L^{\infty}} \leq(M+2)\|h\|_{L^{\infty}} .
\end{aligned}
$$

As a consequence, we have proved that $\psi$ is bounded from above, so this step is concluded.
Step III. $\gamma$ is concentrated on $\partial_{c} \varphi$.
To conclude the proof that $\varphi$ is a Kantorovich potential, we have to check that every optimal transport plan is concentrated on $\partial_{c} \varphi$; since Theorem 1.3 gives the uniqueness of an optimal transport plan, we just have to check that the plan $\gamma$ is concentrated on $\partial_{c} \varphi$ (actually, we do not
need to rely on Theorem 1.3, see Remark 5.7 at the end). Since we know that $\gamma$ is concentrated on $\Gamma$, we will just check that $\Gamma \subseteq \partial_{c} \varphi$.

To do so, let us fix $(\bar{x}, \bar{y}) \in \Gamma$ : by (5.3), for any other $y \in \pi_{2}(\Gamma)$ we have the inequality

$$
\begin{equation*}
\psi(\bar{y})-c(\bar{x}, \bar{y}) \geq \psi(y)-c(\bar{x}, y)=-(c(\bar{x}, y)-\psi(y)) . \tag{5.4}
\end{equation*}
$$

We notice that the same inequality holds also for every $y \in \operatorname{supp} \nu$, because if $y \in \operatorname{supp} \nu \backslash \pi_{2}(\Gamma)$ then we have already noticed that $\psi(y)=-\infty$, so (5.4) is emptily true. Passing then to the supremum in $y \in \operatorname{supp} \nu$, we get $\psi(\bar{y})-c(\bar{x}, \bar{y}) \geq-\psi^{c}(\bar{x})$. We now observe that, since $\psi^{c}=\varphi$, then $\varphi^{c}=\psi^{c c} \geq \psi$, so the last estimate gives

$$
\varphi(\bar{x})+\varphi^{c}(\bar{y}) \geq \varphi(\bar{x})+\psi(\bar{y}) \geq c(\bar{x}, \bar{y})
$$

and since the inequality $\varphi(x)+\psi(y) \leq c(x, y)$ is always true for any $x \in \operatorname{supp} \mu$ and $y \in \operatorname{supp} \nu$, we have proved that $(\bar{x}, \bar{y}) \in \partial_{c} \varphi$, and this step is concluded.
Step IV. $\varphi$ is approximatively differentiable $\mu$-almost everywhere.
To conclude the proof, we have now only to show that $\varphi$ is approximatively differentiable $\mu$ almost everywhere. To do so, let us define the sets

$$
\Theta_{n}:=\left\{x \in \operatorname{supp} \mu: \exists y \in \operatorname{supp} \nu,(x, y) \in \Gamma, y-x \in\left(1-n^{-1}\right) \mathscr{C}\right\}
$$

and let us observe that the union of the sets $\Theta_{n}$ has full measure in supp $\mu$ since $\gamma$ is concentrated in $\Gamma$ and by Theorem C. Let now $x, z$ be two points in $\Theta_{n}$, let $y$ be such that $(x, y) \in \operatorname{supp} \gamma$ with $y-x \in\left(1-n^{-1}\right) \mathscr{C}$, and observe that, since

$$
\varphi(x)+\psi(y)=c(x, y), \quad \varphi(z)+\psi(y) \leq c(z, y),
$$

then we have

$$
\begin{equation*}
\varphi(z)-\varphi(x) \leq c(z, y)-c(x, y) \tag{5.5}
\end{equation*}
$$

Let now $\varepsilon_{n}>0$ be a geometrical constant such that $B\left(0,2 \varepsilon_{n}\right) \subseteq(1 / n) \mathscr{C}$ : then, since $c$ is a relativistic cost and by the properties of convex functions, by (5.5) we immediately get the existence of a constant $K_{n}$ such that, as soon as $|z-x| \leq \varepsilon_{n}$,

$$
|\varphi(z)-\varphi(x)| \leq K(n)|z-x| .
$$

In other words, $\varphi$ is locally Lipschitz in $\Theta_{n}$ and, since it is also bounded by Step I, we get that it is actually Lipschitz there. By Kirszbraun's Theorem, there is a Lipschitz extension $\varphi_{n}$ of $\varphi_{\mid \Theta_{n}}$. So, by Definition 5.4, we obtain that $\varphi$ is approximately differentiable at $x$ for every point $x$ of density 1 in $\Theta_{n}$ where $\varphi_{n}$ is differentiable, hence for $\mu$-almost every $x \in \Theta_{n}$ (keep in mind that $\mu$ is absolutely continuous with respect to the Lebsegue measure). Since this holds for every $n \in \mathbb{N}$, we have obtained that $\varphi$ is approximatively differentiable $\mu$-almost everywhere.

Having the existence of the Kantorovich potential at hand, the existence of an optimal transport map is easy and classical. We give them for the sake of completeness.

Proof (of Theorem B). Let $\varphi_{t}$ be the Kantorovich potential given by Proposition 5.6; let us also call $\Theta_{n}$ the sets defined in Step IV of the proof of that proposition. Also recalling Theorem C we know that, for $\mu$-a.e. $x$, there exists some $y$ such that $(x, y) \in \operatorname{supp} \gamma_{t}, y-x \in t \mathscr{C}$, and $\varphi_{t}$ is approximatively differentiable at $x$. For every $z \in \cup_{n} \Theta_{n}$ close enough to $x$ we know the validity of (5.5), which can be rewritten as

$$
h\left(\frac{y-z}{t}\right)-h\left(\frac{y-x}{t}\right) \geq \varphi_{t}(z)-\varphi_{t}(x)=\widetilde{\nabla} \varphi_{t}(x) \cdot(z-x)+o(|z-x|) .
$$

As a consequence, since $x$ has density 1 in $\cup_{n} \Theta_{n}$, we get

$$
-t \widetilde{\nabla} \varphi_{t}(x) \in \partial h\left(\frac{y-x}{t}\right),
$$

hence

$$
y=x+t \nabla h^{*}\left(-t \widetilde{\nabla} \varphi_{t}(x)\right),
$$

being $h^{*}$ the Legendre transform of the strictly convex function $h$. This shows that the optimal transport plan $\gamma_{t}$ is actually an optimal transport map, in particular $\gamma=\left(I d, F_{t}\right)_{\#} \mu$ for the optimal transport map $F_{t}(x)=x+t \nabla h^{*}\left(-t \widetilde{\nabla} \varphi_{t}(x)\right)$. The proof is then concluded.

Remark 5.7. As underlined in the Introduction, our construction does not rely on the result of Theorem 1.3; in particular, we do not need to know a priori that there is a unique optimal transport plan and that it is given by a map, since we are able to show it in Theorem B. Nevertheless, we have seemingly used this existence during the proof of Proposition 5.6, at the beginning of Step III. Therefore, we have now to explain how the proof really works if one does not want to use the uniqueness given by Theorem 1.3. In Step III of the proof of Proposition 5.6, we are only able to check that the optimal plan $\gamma$ used to build $\varphi$ is concentrated on $\partial_{c} \varphi$; hence, we are still not sure that $\varphi$ is a Kantorovich potential, because we would not know how to deal with other optimal plans, if there were. Thus, the proof of Theorem B still ensures that the plan $\gamma$ is actually a transport map, but it gives no information on possible other plans. However, we can now conclude in a classical way: indeed, by the arbitrariness of $\gamma$, we have proved that every optimal transport plan is a transport map. If now $\gamma_{1}$ and $\gamma_{2}$ are two optimal transport plans, then so is also $\gamma=\left(\gamma_{1}+\gamma_{2}\right) / 2$, thus it must be a map. But it is immediate to realize that $\gamma$ can be a map only if $\gamma_{1}=\gamma_{2}$ : this shows the uniqueness of the optimal transport plan, and in turn only now we can conclude that the map $\varphi$ built before is actually a Kantorovich potential.

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